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INTRODUCTION TO MODERN PHYSICS

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INTRODUCTION TO MODERN PHYSICS

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CHAPTER I
HISTORICAL SKETCH
FIRST PERIOD: EARLIEST TIMES TO 1560 A.D.

1. The Greeks.—Relatively speaking, the contributions made by the Greeks to the natural sciences were far less than their contributions to mathematics, literature, art, and metaphysics. Nevertheless, in spite of their vague and misty philosophizing concerning natural phenomena and in spite of their general failure to test theory by experiment, the Greeks gave to the world much of the physics that was known up to 1400 A.D. In their writings, one finds, here and there, the germ of such fundamental modern principles as the conservation of matter, inertia, atomic theory, the finite velocity of light, and the like.

2. Thales of Miletus (624-547 B.C.), according to Aristotle, was acquainted with the attractive power of magnets and of rubbed amber. He is often said to have discovered the inclination of the ecliptic and the spherical shape of the earth,¹ but Aristotle credited him with the doctrine that the earth was cylindrical in shape and rested on water.

3. Pythagoras (580-500 B.C.) was one of the greatest of the early Greek philosophers and the founder of the Pythagorean school. He held that the earth is spherical, although the basis of this belief is not known. According to Heath,² his argument was probably a "mathematico-esthetic" one based on the idea that "the sphere is the most perfect of all figures." Pythagoras himself, and probably his immediate successors among the Pythagoreans, believed that the entire universe was spherical in shape with the earth at its center and that the sun, stars, and planets moved in independent circles around the earth as a center.

4. Anaxagoras (500-428 B.C.) and Empedocles (484-424 B.C.). According to Plato, Anaxagoras neglected his possessions in order to devote himself to science. He is credited with the view that the moon does not shine by its own light but that "the sun places the brightness in the moon" and "the moon is eclipsed through the interposition of the earth." Also, "The sun is eclipsed at new moon through the

¹ ROSENBERGER, F., "Geschichte der Physik," vol. I, p. 6.

* HEATH, T. L., "Aristarchus of Samos."

interposition of the moon."¹ Apparently, however, human nature, as well as human curiosity, has not changed for 2,500 years, for Anaxagoras was accused of impiety because he taught that the sun was a red-hot stone and that the moon was simply earth, and for holding this doctrine he was banished from Athens.

To Anaxagoras is due the germ of the idea of the atomic hypothesis of Democritus, who lived in the next generation. Anaxagoras denied the contention of the earlier Greeks regarding the creation or destruction of matter. He taught that changes in matter are due to combinations or separations of small, invisible particles (spermata). These particles themselves were conceived to be unchangeable and imperishable but different from each other in form, color, and taste. This doctrine foreshadowed the law of the conservation of matter.

Empedocles, on the other hand, reduced the elements to four—earth, water, air, and fire—through combinations and separations of which the All exists. He also held definite views concerning the phenomena of light. According to him, light is due to the emission by the luminous or visible body of small particles that enter the eye and are then returned from the eye to the body, the two "streams" giving rise to the sense of form, color, etc.

According to Aristotle, Empedocles believed that light "takes time to travel from one point to another." This idea was rejected by Aristotle, who stated that "though a movement of light might elude our observation within a short distance, that it should do so all the way from east to west is too much to assume."²

5. **Democritus** (460-370 B.C.) gave more definite form to the atomic hypothesis of Anaxagoras by postulating that *the universe consists of empty space and an (almost) infinite number of indivisible and invisible particles* which differ from each other in *form, position, and arrangement*. In support of this hypothesis, Democritus argues that the creation of matter is impossible, since *nothing* can come from *nothing* and, further, nothing which is can cease to exist. Aristotle³ puts the argument in this form: "If, then, some one of the things which are is constantly disappearing, why has not the whole of what is been used up long ago and vanished away?" But Aristotle rejects the atomic hypothesis which, indeed, on the basis of speculative reasoning alone, could not evolve beyond the point where Democritus left it.

6. **Aristotle** (384-332 B.C.), a pupil of the philosopher Plato, contributed so much to all branches of knowledge—logic, rhetoric, ethics,

¹ *Ibid.*, p. 78. Quotations cited from later Greek writers.

² Quoted by Heath, *op. cit.*, p. 93.

³ "De Generatione et Corruptione," Book 1, Cap. III, translated by H. H. Ieachim, the Clarendon Press, Oxford, 1922.

metaphysics, psychology, natural science—that it is difficult to sift out that which is germane to a brief history of physics. Perhaps the most important single fact is the tremendous influence which, as a result of his intellectual brilliance and achievements in *many* branches of learning, he exerted for many succeeding centuries in *all* branches, physics included. Viewed from our twentieth-century vantage point, however, not a little of his reasoning concerning the physical universe sounds like piffle. For example, in "De Generatione et Corruptions," he discusses the "coming-to-be" and the "passing-away" of things, and argues for the indestructibility of matter by saying that (Book I I, Cap. X) "it is far more reasonable (to assume) that *what is* should cause the coming-to-be of *what is not* than that *what is not* should cause the being of *what is*," which is understandable. But then follows the curious argument: "Now that which is being moved *is*, but that which is coming to be *is not*; hence, also, motion is prior to coming-to-be . . . and we assert that motion causes *coming-to-be*." But *coming-to-be* and *passing-away* are two processes contrary to one another. Therefore, says Aristotle, we must look for *two motions*, likewise contrary, as the cause of both *coming-to-be* and *passing-away*. Since these processes go on continuously, we must look for continuous motion. Only motion in a circle is continuous, and motion in an inclined circle has the necessary duality of opposing movements. Such a motion is that of the sun along the ecliptic, which, as it approaches (spring), causes coming-to-be and, as it retreats (autumn), causes decay.

And yet Aristotle frequently calls in observed facts to substantiate his speculations. For example, in "De Caelo" (Book I I, Cap. XIV), after proving, by a more or less abstract argument, that the earth is spherical, he says:

The evidence of all the senses further corroborates this. How else would eclipses of the moon show segments as we see them? . . . since it is the interposition of the earth that makes the eclipse, the form of this line (*i.e.*, the earth's shadow on the moon) will be caused by the form of the earth's surface, which is therefore spherical.

He also points to the apparent change in altitude of the stars as one travels north or south and concludes that "not only is the earth circular, but it is a circle of no great size."

Indeed, in theory if not in his own practice, Aristotle places great emphasis on the importance of facts in connection with scientific development. In a paragraph in "De Generatione et Corruptione" (Book I, Cap. II), he says:

Lack of experience diminishes our power of taking a comprehensive view of the admitted facts. Hence, those who dwell in intimate association with

nature and its phenomena grow more and more able to formulate, as the foundation of their theories, principles such as to admit of a wide and coherent development; while those whom devotion to abstract discussions has rendered unobservant of the facts are too ready to dogmatize on the basis of a few observations.

This is surely good doctrine even for twentieth-century scientists!

An attempt to summarize Aristotle's views on physics is beyond the scope of this book, but reference may be made to two of his doctrines because of their bearing upon subsequent history.

The first is his supposed views on falling bodies. The statement is commonly made that Aristotle held that a heavy body would fall from a given height with greater velocity than a light body. It is difficult, however, to be sure from Aristotle's extant writings just what he actually held in regard to this point. The passages that seem to refer to it occur in the course of his arguments against the possibility of the existence of a void. For example, he states¹:

. . . We see the same weight or body moving faster than another for two reasons, either because there is a difference in what it moves through, as between water, air, and earth, or because, other things being equal, the moving body differs from the other owing to excess of weight or lightness. . . . And always, by so much as the medium is more incorporeal and less resistant and more easily divided, the faster will be the movement.

Here, as elsewhere, Aristotle speaks always of movement of a body through a medium. It has been suggested by Greenhill² that he may have meant *terminal* velocity, such as the constant velocity of rain drops as they approach the ground. The *heavier* drops *do* fall faster. It seems perhaps more probable that Aristotle, believing that a medium of some sort must always be present, was unaware of such distinctions as that between terminal and nonterminal velocity and actually did believe that in all stages of its motion the heavier body falls faster.³

The second doctrine referred to is that of the motion of the earth, sun, and planets. In his "De Caelo" (Book II, Cap. XIV), after a series of abstract arguments, in the course of which he states that "heavy bodies forcibly thrown quite straight upward return to the point from which they started even if they be thrown to an infinite (!) distance," Aristotle concludes "that the earth does not move and does not lie elsewhere than at the center." He supposed that the sun, planets, and stars are carried by a series of concentric spheres which

¹ This and the following quotation from Aristotle are taken from "The Works of Aristotle translated into English," vol. II, the Clarendon Press, Oxford, 1930.

² GREENHILL, *Nature*, vol. 92, p. 584.

³ Cf: COOPER, L., "Aristotle, Galileo, and the Tower of Pisa," 1935.

revolve around the earth as a center. The authority of Aristotle was so great as to render sterile the brilliant work of Aristarchus in the next century.

7. Aristarchus (about 310-230 B.C.) enunciated a cosmogony identical with that proposed by Copernicus nearly 2,000 years later. No mention of this hypothesis is made in his only extant work, "On the Sizes and Distances of the Sun and Moon," but Archimedes tells us, in a book called "The Sand-reckoner," that "Aristarchus of Samos brought out a book" containing the hypothesis "that the fixed stars and the sun remained unmoved; that the earth revolves around the sun in the circumference of a circle, the sun lying in the middle of the orbit"; and that "the sphere of the fixed stars" is very great compared with the circle in which the earth revolves. The prestige of Aristotle was too great, however, and the geocentric hypothesis that he supported was so completely satisfactory to the ancient mind that Aristarchus' theory was practically lost for nearly 2,000 years.¹

8. Archimedes (287-212 B.C.), whose name is known to every student of elementary physics because of the famous principle of hydrostatics that bears his name, was one of the most noted physicists of antiquity. He was a man of great ability in what would now be called "theoretical (or mathematical) physics" as well as a practical engineer—a sort of ancient Lord Kelvin. In one of his books already mentioned, "The Sand-reckoner," he computes that 10^{63} grains of sand would fill the sphere of the universe as fixed by Aristarchus. In another, "On Floating Bodies," he lays the foundations of hydrostatics. His Proposition 7, in this book, enunciates the famous principle: "A solid heavier than a fluid will, if placed in it, descend to the bottom of the fluid, and the solid will, when weighed in the fluid, be lighter than its true weight by the weight of the fluid displaced."

9. From the Greeks to Copernicus.—To give but a passing comment to the 17 centuries between Archimedes and Copernicus would seem to give the reader the impression that no developments of moment occurred during that long period. This impression is almost correct. During ancient times Ptolemy of Alexandria (70-147 A.D.) collected the optical knowledge of his time in a book in which he discusses, among other things, reflection from mirrors—plane, convex, concave—and, particularly, refraction, which Ptolemy evidently studied experimentally. He gives, in degrees, relative values of angles of incidence

¹ Heath's "Aristarchus of Samoa" is an exceedingly interesting and valuable book. It contains a review of early Greek astronomy up to the time of Aristarchus, a discussion of his work, and a translation of his only extant book, "On the Sizes and Distances of the Sun and Moon."

and of refraction for air-water, air-glass, and water-glass surfaces and describes an apparatus by which he determined these quantities; he states that for a given interface these two angles are proportional. He also mentions atmospheric refraction as affecting the apparent position of stars. He invented a complicated theory of the motions of the planets in their orbits about the earth in order to explain their apparent motions among the stars.

From Ptolemy to the Arabian Alhazen, is a span of nine centuries—twice the total lapse of time from the discovery of America to the present—during which there was stagnation in almost all lines of intellectual pursuits. But about the eighth century A.D., as an indirect result of religious activity, the Arabs began to cultivate chemistry, mathematics, and astronomy, in large part by translating into Arabic the works of the Greeks but also, in a few instances, by making original contributions. About the year 1000, Alhazen produced a work on optics in seven books. This treatise sets forth a clear description of the optical system of the eye, discusses the laws of concave and convex mirrors, both spherical and cylindrical, and carries the subject of refraction a little further by recognizing that the proportionality between the angles of incidence and refraction holds only for small angles.

During the next 500 years, a very few advances in physics were made. Roger Bacon (1214-1294), British philosopher and scientist and a monk of the Franciscan Order, taught that in order to learn the secrets of nature *we must first observe*. He believed in mathematics and in deductive science, but he clearly realized that only as scientific conclusions are based on observed phenomena and tested by experiment can useful knowledge result.

About the same time Petrus Peregrinus recognized that magnetic poles are of two kinds, like poles repelling and unlike attracting each other.

Then there was Leonardo da Vinci (1452-1519), Italian painter, architect, sculptor, engineer, and philosopher, whose greatness as a scientist has come to be appreciated only in recent years, for his works were left in manuscript form and were probably not widely known among his contemporaries—for which reason his influence on early science is comparatively insignificant. His belief in the value of experiment is worthy of the twentieth century: "Before making this case a general rule, test it by experiment two or three times and see if the experiment produces the same effect." Although expressed in the vague language of his time, some of his ideas concerning what we now refer to as "force," "inertia," "acceleration,"

the "laws of motion," etc., were qualitatively correct. Concerning perpetual motion, he wrote: "Oh, speculators on perpetual motion, how many vain projects of the like character you have created! Go and be the companions of the searchers after gold." Rejecting the Ptolemaic theory, he held that "the sun does not move." That he was not persecuted or even burned at the stake, as was Bruno a century later, for holding such revolutionary and, therefore (!), Heretical views is probably due to the fact that his doctrines were given so little publicity; for, holding no academic position, he did not teach, and he published nothing.

Finally, in the sixteenth century, the full force of that period of intense intellectual activity known as the Renaissance began to be felt in the field of physics. Then were produced such men as Copernicus, Tycho, Kepler, Galileo, Newton, who with their contemporaries and colleagues, in a space of hardly more than a century, completely broke the "spell" of Aristotle and made possible the beginnings of modern experimental science. In so far as the heliocentric theory completely revolutionized man's conception of the universe and his place in it, it is quite correct to regard the work of Copernicus as beginning a new era in scientific thought. But had it not been for other discoveries coming immediately after Copernicus, such as the telescope, Kepler's laws, Galileo's famous experiments on falling bodies, and many others, it is quite possible that the theory of Copernicus would have had the same fate as that of Aristarchus centuries earlier. It is, therefore, fitting to regard the birth of the Copernican theory as *closing* the first period in the history of physics.

10. The Copernican System.—Copernicus (1473-1543), a younger contemporary of Columbus, spent most of his life as one of the leading canons in the monastery at Frauenburg, near the mouth of the Vistula. As a student, however, he had studied, among other subjects, mathematics and astronomy; at the University of Bologna, he was a pupil of a famous astronomer, Novara, who supported the Pythagorean system of the universe as against the Ptolemaic. Copernicus's new theory of the universe is set forth in his "De Revolutionibus Orbium Coelestium," which he allowed to be printed near the close of his life. The new theory appears to have been the result of long reflection upon the difficulties of the current astronomical theory of his day, **and** upon the various speculations of ancient philosophers.

Copernicus perceived that, by assuming that the earth is a planet like the others and that **all** the planets move **in** circles around the sun, **a great simplification, both philosophical and mathematical,**

could be made with regard to the world system. He could thus easily account for the seasons and for the apparent retrograde motion, at times, of the planets. The rotation of the earth on its axis causes the *apparent* daily motion of the sun, moon, and stars; and he pointed out that, probably, the stars are too far away for any motion of the earth to affect their apparent positions. He gave the correct order of the planets from the sun outward.

Whatever the system as proposed by Copernicus lacked quantitatively, it was correct, in its main outline, qualitatively. Its reasonableness set a few men thinking and did much to usher in a new era in science, an era that could come only when truth could have the opportunity of standing alone, *unaided* or *unhindered* by the "authority" of 2,000 years.

SECOND PERIOD (1650-1800 A.D.): RISE OF THE EXPERIMENTAL METHOD

11. Galileo Galilei (1564-1642).—Galileo is widely, and quite correctly, regarded as the father of modern physics. To be sure, modern physics has grandfathers and still more remote ancestors, but none of them gave to experimental physics so much as Galileo. Physicists should be acquainted with the main details not only of his scientific work but also of his life. The student is urged, therefore, to peruse one of the several biographies and to read some of Galileo's writings, at least in translation. Even a short time spent in following his deductions and in reading *firsthand* something he wrote will prove both interesting and instructive.

Galileo was descended from a noble family, and it is quite probable that he inherited from his father the spirit of free inquiry which characterized his life. For, in the writings of the elder Galileo, who was well educated and was an accomplished musician, is the statement: "It appears to me that they who in proof of any assertion rely simply on the weight of authority, without adducing any argument in support of it, act very absurdly."

As a student in the monastery of Vallombrosa, near Florence, the young Galileo excelled in the classics and literature and was something of a poet, musician, and art critic. He also showed an aptitude for science and exhibited considerable mechanical inventiveness. At the age of seventeen he was sent to the University of Pisa to study medicine. It was here that he made his first discovery and invention. One day, in 1581, he noticed the regular oscillations of the great hanging lamp in the cathedral at Pisa. Although the amplitude of these oscillations became less and less, they were all

reached Galileo in June, 1609. Immediately grasping the principle involved, he made a telescope and exhibited it in Venice "for more than a month, to the astonishment of the chiefs of the republic." By January, 1610, Galileo had made a telescope with a power of 30 diameters,¹ and with this instrument he made a number of fundamental discoveries. He saw that the number of fixed stars was vastly greater than could be seen by the unaided eye, and thus he was able to explain the agelong puzzle of the Milky Way. He saw that the planets appeared in his telescope as luminous disks while the stars still remained points of light, and he discovered the satellites of Jupiter. These discoveries naturally made Galileo famous, and he soon accepted an invitation to return to Pisa as "First Mathematician and Philosopher," at a very substantial increase in salary, but at a sacrifice, unfortunately, of his "academic freedom" in Padua. Continuing his astronomical investigations, he discovered the crescent phases of Venus, sunspots and the rotation of the sun, the faculæ of the solar atmosphere, and the libration of the moon. In 1612, he published his "Discourse on Floating Bodies."

At first, it seemed as if his fame had silenced all opposition from the church. But the support that his discoveries gave to the hated Copernican theory and his vigorous attacks on Aristotelian philosophy roused his enemies to fury, with the result that in 1615 he was hauled before the Pope and, under threat of imprisonment and torture, was "enjoined . . . to relinquish altogether the said opinion that the sun is the center of the world and immovable . . . nor henceforth to hold, teach, or defend it in any way" Simultaneously, it was decreed that the works of Copernicus "be suspended until they be corrected." Galileo acquiesced in these decrees and was allowed to return to Pisa, where he continued his researches along such lines as would not give offense.

In 1623, one of Galileo's friends, Barberini, became Pope Urban VIII, from whom Galileo received assurances of "pontifical good will." Thereupon, thinking that the decree of 1615 would no longer be enforced, he began the writing of his great book, "Dialogues on the Ptolemaic and Copernican Systems." This was published in 1632, under formal permission from the censor. The form of these dialogues is ingeniously contrived to abide by the *letter* of the decree of 1615. Three "characters" carry on the discussion: Salviati, a Copernican; Simplicio, an Aristotelian; and Sagredo, a witty, impartial, good-

¹ Galileo's telescopes were similar to the modern opera glass—a double-convex (or plano-convex) object glass and a double-concave eyepiece. Thus, they had an erect image.

natured chairman. The dialogues cover four "Days," during which the arguments for and against each system are set forth with apparent impartiality and without reaching any *stated* conclusion. Nevertheless, the general effect of the book was "a powerful plea for Copernicanism."¹

In spite of its enthusiastic reception by the public, the form of the book did not deceive his enemies, who were now determined that he must be silenced. This was effectually accomplished by representing to the Pope that the Simplicio of the dialogues, whose ignorance was very apparent, was simply a caricature of the Pope himself. In spite of the absurdity of the argument—for Galileo would hardly have risked offending Urban VIII, his one friend in the church—the Pope, who seems to have been both ambitious and vain, became convinced that Galileo "had made game of him." Whereupon he was ready to join Galileo's enemies in persecuting that great scientist, ostensibly for "the safety of the church and the vindication of its decrees."

This tragic incident is illustrative of the fact that, in the seventeenth just as in the twentieth century, much of the "warfare between science and theology"—as Andrew D. White calls it²—has been based upon personal motives rather than upon a sincere wish to uphold theological doctrines. Intolerance, even when it is sincere, is to be condemned; but *insincere* intolerance is to be despised. Yet, one cannot fully comprehend the forward march of science unless one recognizes the seriousness of the obstacles of this kind that have had to be overcome. Great as is the fame of Galileo, how much more might he have accomplished if the energy that he was forced to spend in overcoming opposition could have been directed toward his researches. Without doubt, the giant strides being taken by science in the twentieth century are possible only because of the freedom that the scientist now enjoys.

Galileo was presently called before the Inquisition. He was sixty-seven years old, impaired in health and in spirit. Bowing to the inevitable because of the magnitude of the forces arrayed against him, he followed the advice of his friends and indicated his "free and unbiased" willingness to recant, to "abjure, curse, and detest the said heresies and errors and every other error and sect contrary to the Holy Church," and he further agreed "never more in future to say or assert anything, verbally or in writing, which may give rise to a similar suspicion." Thereafter Galileo was kept a prisoner under suspended

¹ FAHIE, *-loc. cit.*

² WHITE, ANDKEW D., "The History of the Warfare between Science and Theology."

sentence, first at Rome, then at his home in Arcetri. Here, during the last years of his life, he prepared and, in 1636, published his "Dialogues on Two New Sciences"¹ (i.e. Cohesion and Motion).

The dialogues on "Motion" sum up Galileo's earlier experiments and his more mature deliberations. He states that "if the resistance of the media be taken away, all matter would descend with equal velocity." He deduces the formulas of uniformly accelerated motion. He shows that the path of a projectile is parabolic under suitable limiting conditions and states that, if all resistance were removed, a body projected along a horizontal plane would continue to move forever. His work on mechanics paved the way for the enunciation by Newton of the famous three laws of motion, which form the foundation of mechanics.

12. Tycho Brahe (1546-1601) and Kepler (1571-1630).—The work of Tycho and Kepler is particularly interesting not only because of its direct bearing on the development of physics but more particularly because of the mutual dependence of the work of each one upon that of the other, a relation very common in present-day science. Tycho was the experimentalist, the observer, who supplied the accurate data upon which Kepler, the theorist, built a new theory of planetary motion. Without a Kepler to build a theory from them, Tycho's observations would have attracted hardly more than passing notice. Kepler, in turn, might have theorized to his heart's content, but, without the accurate data of a Tycho, those theories would ultimately have shared the fate of Aristotle's. Sometimes theory precedes, sometimes experiment. But neither can get far without the other.

Tycho Brahe, born of a noble family in Sweden, was educated for a career as a statesman, but he developed a consuming interest in astronomy. By means of observations of his own, he soon found that the current astronomical tables were incorrect. In 1575, he was put in charge of the observatory of Uraniborg by King Frederick II of Denmark, one of his duties being to make *astrological* calculations for the royal family. Here he spent 20 years making systematic observations of the planets, constructing a star catalogue, and accumulating

¹ See the excellent translation by Crew and de Salvio.

Some writers have severely censured Galileo for yielding to the Inquisition. They say that "had Galileo added the courage of a martyr to the wisdom of a sage . . . science would have achieved a memorable triumph" (see Brewster, "Life of Newton"). Whatever opinions on this question one may hold, one fact stands out indisputable: Had Galileo *not* yielded he would surely have been cast into the dungeon and would probably have been burned at the stake. We should not then have had handed down to us these dialogues on Motion, so fundamental to our modern physics.

other astronomical data, always with the highest possible order of accuracy that could be attained without a telescope. In 1599, Tycho undertook to establish a new observatory at Prague for the German emperor, Rudolph the Second, but in the midst of this work he suddenly died (1601).

Now among Tycho's assistants at Prague had been, during the last few months, a brilliant young mathematician, Johann Kepler. He succeeded Tycho as principal mathematician to the emperor and undertook to carry to completion the new set of astronomical tables based on elaborate observations, which Tycho had begun. Kepler remained at Prague until 1612; from then until his death in 1630, he held a professorship at Linz.

Kepler believed thoroughly in the Copernican system, which Tycho had rejected for a geocentric system of his own. It is one of the dramatic situations in science that Tycho's data on planetary motions, taken in support of his own theory, became, in the hands of Kepler, the clinching argument for the Copernican system. Using Tycho's observations, Kepler made a special study of the motion of Mars. He tried to reconcile the various recorded positions of the planet by assuming circular orbits for Mars and for the earth, trying various positions of these orbits relative to the sun. None worked. By resorting to the Ptolemaic notion of epicycles and deferents, some improvement resulted, but still the observed positions differed from the computed, in some cases by as much as 8 minutes of arc. Kepler knew that Tycho's observations could not be in error by that amount. Some new concept regarding planetary motion was necessary.

Then Kepler gave up *uniform* circular motion and assumed that the speed varied inversely as the planet's distance from the sun. This assumption is his famous "second law," that *the radius vector from the sun to the planet describes equal areas in equal times*. It worked approximately, but still there were systematic errors which exceeded the possible errors of observation. Finally, he cast aside the last traditions of the Ptolemaic system and tried orbits of other forms, first, an oval path, and, then, an ellipse, with the sun at one focus. At last, his years of computation bore fruit. The path was an ellipse. Theory and observation agreed! And one of the most important and far-reaching laws in all science had been discovered, all because of a discrepancy of 8 minutes of arc between observation and theory! In fact, one of the striking things in the growth of science, particularly physical science, is the fact that many fundamental advances have come about because of just such discrepancies, frequently very small ones, between observation and theory. Studying Tycho's observa-

tions further, Kepler finally hit upon the true relation between the periods of the planets and the radii of their orbits, a relation now known as Kepler's "third law," that *the squares of the times of revolution around the sun are as the cubes of the mean orbital radii*.

Thus were completed the three laws of planetary motion that Kepler handed down to posterity and which, sweeping away all remnants of the Ptolemaic system, paved the way for modern astronomy:

1. The planets move around the sun in orbits which are ellipses, with the sun at one focus.

2. The radius vector (from sun to planet) sweeps over equal areas in equal times.

3. The squares of the periods of revolution of the planets around the sun are proportional to the cubes of the (mean) radii of their respective orbits.

But what makes the planets move? *Why* do the outer ones go more slowly? Is there "one moving intelligence in the sun, the common center, forcing them all around, but those most violently which are nearest?" Kepler speculated long on this question and actually arrived at the idea of an attraction acting between any two material bodies. This *qualitative* idea of Kepler's was later developed by Newton into his *quantitative* theory of universal gravitation. Kepler himself, however, seems to have had no idea that it is this very attractive force which keeps the planets themselves in their orbits.

In passing, it may be mentioned that Kepler also made substantial contributions in the field of optics. He understood clearly the principle of total reflection and how to determine what we now call the "critical angle." He studied atmospheric refraction as affecting the apparent position of the heavenly bodies and worked out an approximate formula to allow for this error for all positions, from zenith to horizon. He was the first to propose the meniscus type of lens. And he proposed the Keplerian or astronomical type of telescope, in which a *real* image is formed, thus making possible accurate measurements by means of cross hairs in the focal plane of the objective.

13. The Experimental Method Spreads.—The impetus given to science by that great trio, Galileo, Tycho, and Kepler, resulted in an ever increasing number of investigators in the generations that followed. We can mention only a few of them. Of great significance, too, is the fact that at about this time there were formed in several European centers various learned societies which brought together, for argument and discussion, men of kindred interests. The Lincean Society was founded in Italy, in 1603; the Royal Academy of Sciences,

in France, in 1666; and the Royal Society for the Advancement of Learning, in England, in 1662. The continued improvement of the art of printing enormously facilitated the diffusion of scientific knowledge.

In 1600, Gilbert, an English physician, published his famous work, "De Magnete," based largely upon his own experiments, in which he showed the fallacy of such popular fancies as the belief that lodestones lost their magnetic power when rubbed with garlic and regained it again when smeared with goat's blood. He was the first to recognize that the earth is a great spherical magnet, and he actually magnetized a small sphere of iron and showed that it produced a magnetic field similar to that of the earth.

Among other workers in magnetism may be mentioned Kircher (1601-1680), who, by measuring the force required to pull a piece of iron from either pole of a magnet, demonstrated the equality of the two poles; Cabeo (1585-1650), who showed that an *unmagnetized* iron needle floated freely on water would place itself along the earth's magnetic meridian; and Gellibrand (1597-1637), who discovered the secular variation of the magnetic declination.

In the field of optics, there was Schemer (1575-1650), who studied the optics of the eye; Snell (1591-1626), who discovered the true law of refraction; and Cavalieri (1598-1647), who gave the correct formula for the focal length of a thin glass lens in terms of the radii of curvature of the two sides. Studies in acoustics, also, were not wanting. For example, Mersenne (1588-1648), after having investigated the laws of vibrating strings, determined, in absolute measure, the frequency of a tone. He also measured the velocity of sound by observing the time interval between the flash of a gun and the arrival of the report.

In the field of fluid mechanics, there was Torricelli (1608-1647), who studied the flow of liquids from orifices, discovered the principle of the barometer, and observed variation in barometric height with altitude. Working independently, Guericke (1602-1686) invented the air pump. Pascal (1623-1662) measured the difference in barometric height between the base and the top of a mountain, correctly explaining the reason for the difference, and, later, announced the famous principle of hydrostatics that bears his name.

Not only was physics, as a subject, beginning to assume definite form, but even the usual subdivisions such as mechanics, light, sound, etc., were beginning to crystallize out.

Then came a man

. . . towering head and shoulders above all his contemporaries, a veritable giant among the giants, a man whose intellect and whose contributions to

knowledge are incomparably greater than those of any other scientist of the past, that prince of philosophers, Sir Isaac Newton.¹

The other "giants" referred to, contemporaries of Newton, are such men as Boyle, Huygens, and Hooke.

14. Sir Isaac Newton (1642-1727).—Newton was born in the little hamlet of Woolsthorpe, England, in 1642, less than a year after the death of Galileo. In the public school at Grantham, he showed at first no exceptional aptitude for his studies, but ultimately he rose to the highest place in the class. Then, at the age of fifteen, he was removed from school to assist his widowed mother in running the family estate at Woolsthorpe. But he had little taste for farming. Rather, he was interested in studying and in devising various mechanisms. He made a water clock, water wheels, sundials, and a working model of a windmill. And one morning his uncle found him under a hedge studying mathematics when he should have been farming. Thereupon Newton's mother wisely decided that an educational career was more suitable for her son, and he was sent back to school and, ultimately, to Cambridge, which he entered in 1661. Here his fondness for mathematics continued, and soon his creative genius began to appear. While still a student, he discovered the binomial theorem, developed the methods of infinite series, and discovered "fluxions," or the differential calculus.

Soon thereafter, an outbreak of the plague closed the University for some months, during which time Newton, at the family estate at Woolsthorpe, began his speculations on the subject of gravity, which later led to his enunciation of the inverse-square law. It was here that the much-quoted "falling apple" episode is said to have occurred, which is supposed to have given Newton the basic idea of *universal* gravitation. But Newton himself makes no mention of the incident, and it seems far more probable that at Cambridge he had read Kepler's qualitative proposal of a general principle of gravitation. Certainly, Newton was familiar with the three laws of planetary motion that Kepler had discovered.

In 1667, Newton returned to Cambridge as Fellow of Trinity. Two years later, at the age of twenty-six, he was appointed Lucasian Professor of Mathematics, a chair which he held for nearly 30 years. In 1703, he resigned his professorship to devote himself to his duties as Master of the Mint, to the scientific work of his contemporaries, and to defending his own work against the attacks of jealous rivals. In this same year, he was elected President of the Royal Society, an

¹ HART, "Makers of Science."

office which he held until his death. In 1705, he was knighted by Queen Anne.

Most of Newton's important scientific work was done before he vacated the professorship, although he remained thereafter "a power of the first magnitude in the world of science." In his later years, he devoted much time to theological studies. He died Mar. 20, 1727, at the ripe old age of eighty-five. Throughout his life, he shunned publicity and retained a modesty and simplicity which are indicated by a sentiment uttered shortly before his death:

I do not know what I may appear to the world, but to myself I seem to have been only like a boy playing on the seashore, and diverting myself in now and then finding a smother pebble or a prettier shell than ordinary, whilst the great ocean of truth lay all undiscovered before me.

Any brief account of Newton's work must inevitably give a very inadequate impression of his contributions to science. The student is urged to read some of his writings firsthand or, at least, some extensive biographical discussion of his life and work. We can here refer only to a very few of his researches on optics and on mechanics.

Newton's work on *optics* arose out of an attempt to improve lenses. The inability of a lens with spherical surfaces to bring parallel rays to a point focus was early recognized. In 1629, Descartes had shown that lenses with hyperbolic or, under certain conditions, parabolic surfaces should be free from the defect which we now call "spherical aberration." Newton found, however, that such lenses produced only a very slight improvement in the image, and he conjectured that, perhaps, the trouble lay not in the lens but in the light itself.

Accordingly, he procured a prism of glass and, placing it over a hole $\frac{1}{4}$ inch in diameter through which sunlight was shining into a darkened room, he observed the "very vivid and intense colors" produced on a wall some 20 feet distant. Newton was surprised to find that this "spectrum," as we now call it, was so much longer than it was wide $5(13\frac{1}{4})$ by 2 inches). The *width* subtended at the hole an angle corresponding *exactly* to the sun's angular diameter, but the length could not be so explained. He made various surmises as to the origin of the colors, such as the varying thickness of the prism from apex to base, the unevenness of the glass, a curvilinear motion of the light after leaving the prism, etc. One by one, experiment proved these hypotheses wrong. Finally, he isolated one ray, or "color," by suitable screens and caused it to pass through a second prism. In this way, he could measure the refrangibility of a single ray. And he found that the refrangibility increased from red to violet, that,



PLATE 2.—Newton,

therefore, the first prism simply "sorted out" the colors, which, in combination, made "white" light. In other words, he discovered that so-called "white light" is made up of the spectral colors—a very elementary concept to us of the twentieth century but very new and of far-reaching importance in 1666.

Newton at once saw that this dispersion of light was the cause of his failure to effect any substantial improvement in telescopes by use of paraboloidal lenses. Furthermore, he concluded, on the basis of a hurried experiment, that, in different mediums, dispersion was always proportional to refracting power. If this were so, it would follow that by no combination of lenses of different materials could chromatic aberration be eliminated. This singular error of Newton's retarded the development of refracting telescopes for many years. In 1730, Hall made several achromatic combinations of crown and flint glasses, but he published no account of his work, so that, when Dolland, about 1757, rediscovered the method of making such combinations, he was able to secure a patent on it—an invention that had been within the grasp of Newton three-quarters of a century before.

Newton's theories concerning the nature of light are of great historical interest. And much has been written concerning the extent to which he is supposed by some to have retarded the development of optics by espousing the corpuscular theory as against the wave theory of his contemporaries, Huygens (1629-1695) and Hooke (1635-1703). Accordingly, it may be of interest to point out, by a quotation or two, that Newton was by no means dogmatic in his support of the corpuscular theory and that later writers may have taken him, in this regard, more seriously than he intended.

In a communication to the Royal Society, in 1675, concerning "An Hypothesis Explaining the Properties of Light," Newton states:

... I have here thought fit to send you a description ... of this hypothesis . . . though I shall not assume either this or any other hypothesis, not thinking it necessary to concern myself whether the properties of light discovered by men be explained by this or any other hypothesis capable of explaining them; yet while I am describing this, I shall sometimes, to avoid circumlocution . . . speak of it as if I assumed it.

He then proceeds to describe "an setherial medium, much of the same constitution with air, but far rarer, subtiler and more strongly elastic" and supposes that

... light is neither aether, nor its vibrating motion, but something of a different kind propagated from lucid bodies. They that will may suppose it an aggregate of various peripatetic qualities. Others may suppose it multi-

tudes of unimaginable small and swift corpuscles of various sizes springing from shining bodies . . . and continually urged forward by a principle of motion which in the beginning accelerates them, till the resistance of the aetherial medium equals the force of that principle much after the manner that bodies let fall in water are accelerated till the resistance of the water equals the force of gravity.

Nearly 33 years later (1704), Newton published his optical researches in book form in his well-known "Opticks," bringing out a third edition in 1721. His first sentence reads: "My Design in this Book is not to explain the Properties of Light by Hypotheses, but to propose and prove them by Reason and Experiment." He then, in some 300 pages, gives his researches on refraction, reflection, colors of thin plates, etc., and he concludes the book by "proposing only some queries in order to further search to be made by others." One of these queries expresses his objections to the wave theory of light:

28. Are not all hypotheses erroneous in which light is supposed to consist in pression or motion propagated through a fluid medium? If light consists only in pression propagated without actual motion, it would not be able to agitate and heat the bodies which refract and reflect it, and . . . it would bend into the shadow. For pression or motion cannot be propagated in a fluid in right lines beyond an obstacle . . . but will bend and spread every way into the quiescent medium which lies beyond the obstacle.

In order to account for the colors of thin films, which are now regarded as strong evidence for the wave nature of light, he supposes that "every ray of light in its passage through any refracting surface is put into a certain transient constitution or state, which in the progress of the ray returns at equal intervals and disposes the ray at every return to be easily refracted through the next refracting surface and between the returns to be easily reflected by it." He even suggests that the effect might be due to vibrations excited by the "rays" in the material medium which vibrations "move faster than the rays so as to overtake them; and that when any ray is in that part of the vibration which conspires with its motion, it easily breaks through a refracting surface, but when it is in the contrary part of the vibration which impedes its motion, it is easily reflected . . . But whether this hypothesis be true or false, I do not here consider."¹

It is clear that Newton regarded his corpuscular theory as a tentative one, subject to confirmation or rejection on the basis of further experiments. If his theory really did retard progress in optics, the fault lay rather with those who attached too great weight to his opin-

¹ Quoted in Preston's "Light," p. 22.

j ions, perhaps because they retained something of that medieval respect for authority which had kept Aristotle enthroned so long.

Newton's speculations may also serve as an example to illustrate the rule that even the greatest intellect works on the basis of the knowledge and viewpoints of its age. Had Newton lived a century later, he would probably have been one of the first believers in the wave theory. The fact that great scientists share the limitations of their age is an added reason for treating their speculative opinions chiefly as sources of inspiration for further experiment.

Newton's researches on optics alone would have given him a high rank among the scientists of his time. But of still greater value was his work in *mechanics*. [in announcing that "every particle of matter in the universe attracts every other particle with a force inversely proportional to the square of the distance between the two particles," in showing that this one universal and comparatively simple law governs not only the motions of the planets around the sun and of the satellites round their planets but, probably, also the relative motions of all the heavenly bodies, he gave to the world a truth which exercised an enormous influence upon thought] This achievement of Newton's played a large part in bringing about the general conviction that the physical universe in its entirety is governed by law, not by caprice.

Newton himself has told us how he came to discover the law of gravitation. First he attacked the problem of finding a law of attraction between two bodies, such as the sun and a planet, which would result in Kepler's third law, viz., that squares of the periods of rotation of the planets around the sun are proportional to the cubes of their mean distances from the sun. He found that a gravitational attraction varying as the inverse square of the distance gives this law of planetary motion.

Then he saw that a test of this inverse-square law could easily be made by comparing the acceleration of the moon toward the earth with the acceleration of falling bodies at the surface of the earth. It was known that the distance between the moon and the earth's center is about sixty times the earth's radius. By the inverse-square law, therefore, the moon should "drop" toward the earth, in 1 second, $1/60^2$ as far as a body at the surface of the earth drops in 1 second. ^{The} latter distance being, from observations on falling bodies, 16 feet, the former should be $16/60^2$ feet, or 16 feet in 1 minute. But the ^{acceleration} of the moon could be determined directly by applying to the moon the expression he had used for the motion of the planets in their orbits, viz.,

$$a = \frac{v^2}{r} = 4\pi^2 \frac{r}{T^2},$$

where v is the velocity of the moon in its orbit or T is the period of its motion around the earth and r is the radius of the orbit. Now r is equal to sixty times the earth's radius, which was taken as 3,436 miles, on the then common assumption that a degree of latitude is 60 miles. On this basis, the moon is found to "drop" 13.9 feet toward the earth in a minute, instead of 16 feet, as should be the case if the inverse-square law were obeyed. Thus the two results did not agree.

Newton was twenty-three years old at the time, and he laid this calculation aside, not mentioning it to anyone. Some years later, however, he learned of a new and more accurate determination of the length of a degree which had been made by Picard, who found not 60 but more nearly 70 miles. In the meantime Newton had also succeeded in proving that a homogeneous sphere attracts an external body as if all its mass were concentrated at its center, thereby removing one small uncertainty in the calculation. So, on the basis of Picard's new value for the length of a degree, Newton revised his computations on the moon's acceleration and, to his great joy, found that it falls toward the earth 16 feet in a minute, just as predicted by the inverse-square law. At last, he had discovered the true law of gravitation. On the basis of this law, he could now *derive* all three of Kepler's laws. Theory and observation checked perfectly.

These results, together with some propositions on the motion of the planets, were communicated in 1683 to the Royal Society, which requested permission to publish Newton's complete researches on the subject of motion and gravitation. In 1687, there appeared the first edition of his "Principia" or, in full, "Philosophiae Naturalis Principia Mathematica" (Mathematical Principles of Natural Philosophy), "without exception the most important work in natural philosophy extant."¹ The original is in Latin, but English translations are available. The treatise is divided into three books, the subject matter of each being presented by propositions, theorems, and corollaries. The first two books deal with general theorems concerning the motions of bodies, whereas the third contains applications to the solar system. The entire treatise is characterized by the exposition of the principle of universal gravitation and its ramifications but, as the author carefully points out, without attempting any hypothesis as to the *cause* of gravitation.

In this treatise the famous three laws of motion are assumed as axioms. Their greatest merit lies in the fact that they contain just enough to constitute a complete basis for the science of mechanics and no more. The laws can be expressed with greater clarity in modern terminology, and we now realize that in payt they express

¹ HART; *loc. cit.*

definitions rather than experimental facts, but it is commonly felt that no more *convenient* basis for mechanics has been proposed.

In addition to the formulation of the laws, Newton's treatise makes several other important contributions to mechanics. He was the first to give precision to the concept of mass; and he describes a series of experiments showing that the period of a pendulum is independent of the material of which it is made, from which he concluded that, for different bodies, mass and weight are proportional to each other. He also gave precision to the idea of force, and formulated in general terms the principle of the parallelogram of forces.

Of Newton's invention of the method of fluxions (i.e., the calculus), of his very interesting miscellaneous writings, particularly those on theology, of the many controversies with his contemporaries into which he was unwillingly drawn in defense of his scientific work, we cannot take space to write. We can only urge our readers to make further study of the life and works of this renowned physicist.

15. Newton's Contemporaries.—The true productive period of Newton's life ended about 1700. His biography is so full of interest and inspiration that it is a temptation to discuss, similarly, the work of his contemporaries, themselves eminent scientists—the Honorable Robert Boyle (1627-1691), the discoverer of "Boyle's law"; Huygens (1629-1695), whose wave theory of light was to triumph a century and a half later; Robert Hooke (1635-1703), proponent of the undulatory theory and originator of "Hooke's law" of elasticity; Leibnitz (1646-1716), whose calculus ultimately replaced Newton's fluxions. But, remembering that the main business of this book is "modern" physics, we must pass on to a rapid review of the developments of physics during the *eighteenth* century.

16. Mechanics during the Eighteenth Century.¹—The subsequent history of mechanics has consisted of the derivation, from Newton's three laws, of various secondary principles which are convenient for special purposes, and of the solution of all sorts of problems. Among the prominent workers during the eighteenth century, we find such names as Daniel Bernoulli (1700-1782), who worked on hydrodynamics, the kinetic theory of gases, and the transverse vibrations of rods; Euler (1707-1783), who shares with Bernoulli the honor of discovering the general law of the conservation of angular momentum (1746); and Lagrange (1736-1813), who gave, in the equations that bear his name, a general method of attacking any problem, using any sort of coordinates that may be convenient.

¹ The reader will find in Mach's "The Science of Mechanics" a very valuable discussion of the history and the fundamentals of the subject of mechanics.

17. Heat during the Eighteenth Century.—Galileo had invented an air thermometer in 1597, but the first mercury thermometer was used by Kircher in 1643. About 1724, Fahrenheit proposed the temperature scale now known by his name; this was followed by the Reaumur scale and, in 1742, by the Celsius scale. James Black (1728-1799), a professor of chemistry at Glasgow and Edinburgh, made measurements of the heat of fusion and of vaporization of water, which led to modern calorimetry, and gave definite form to the previously hazy distinctions between temperature and heat.

With regard to theories of heat, however, there was retrogression during the eighteenth century. From his writings, it is clear that Newton regarded heat as intimately connected with the motion of the small particles of which bodies are composed, and this view seems to have been shared by his contemporaries. But, early in the eighteenth century, there was a return to the "caloric" theory, which held heat to be a fluid that could be extracted from or added to a body, thereby causing its changes of temperature. This heat fluid was indestructible, its particles were self-repellent but were attracted by ordinary matter, and it was all-pervading. The expansion of bodies when heated was the natural result of "swelling" due to forcing caloric into matter. The production of heat by percussion was due to the releasing or "pounding loose" of some of the caloric naturally condensed in or absorbed by the body, thereby increasing the amount of free caloric. Black explained latent heats and specific heats on the basis of this theory. Indeed, by the end of the eighteenth century, the caloric theory of heat was generally accepted.

18. Light during the Eighteenth Century.—An event of special importance in the history of science was the discovery of the aberration of light by Bradley in 1728. The absence of any measurable stellar parallax had been one of the stumbling blocks in the way of the Copernican system and was, therefore, one of the outstanding problems of astronomy. Tycho had recognized that, viewed from opposite sides of the earth's orbit, the stars should show a perspective displacement, but his careful observations convinced him that no such displacement so great as 1 minute of arc existed. Later observers, likewise, sought such an effect in vain.

In hopes of being able to measure stellar distances, Bradley began, in 1725, systematic observations on the position of a zenith star, 7 Draconis. If stellar parallax existed, this star should be farthest south in December and should then move north, reaching its maximum northerly position 6 months later. The position of the star was found to change, but not in the manner expected. It reached farthest south

in *March* and farthest north in *September*, the angular distance between the two positions being about 40 seconds of arc. Bradley continued his observations on other stars, and, in 1728, he came to the conclusion that the observed displacement was not due to parallax at all but to an apparent shift in the star's position due to a combination of the velocity of light with that of the earth in its orbit.¹ He was thus enabled to deduce a value for the velocity of light. The value so found was in substantial agreement with that determined by Romer a half-century earlier, from a study of the motion of Jupiter's moons, which constituted the first determination of the velocity of light. This discovery of Bradley's was the first in the series that formed the basis for the modern theory of relativity.

Theories as to the nature of light made no material progress during the eighteenth century. Some writers are inclined to ascribe this to the prestige given to the corpuscular theory by the fact that it was supported by Newton, whose preeminence "seemed to act like a spell," as had Aristotle's centuries before. But if so, then, likewise, the kinetic theory of heat held by Newton should have been uppermost during the eighteenth century, whereas, as has already been pointed out, the reigning theory of heat during this period was the caloric theory. Lack of progress in the theory of light was more likely due to the lack of any crucial experiment, just as was the case with theories of heat. Science has never progressed on the basis of speculation only.

19. Electricity during the Eighteenth Century.—Electricity received a great deal of attention during the eighteenth century, but research was concerned principally with electrostatics. Stephen Gray (1670-1736) distinguished clearly between conductors and nonconductors of electricity and showed that even conducting bodies may be electrified provided they are insulated. Du Fay (1698-1739) showed that flames exercise a discharging power and that there are two kinds of electricity, which he called "vitreous" and "resinous." He was thus led to propose the two-fluid theory of electricity. During the first half of the eighteenth century, the electroscope was invented (by Hawksbee, in 1705), frictional electric machines were developed, the Leyden jar was discovered (1745), and there was considerable popular interest in electrical phenomena. During the latter half of the century, three names stand out preeminent: Benjamin Franklin (1706-1790), Henry Cavendish (1731-1810), and Charles A. Coulomb (1736-1806).

Franklin's experiments began about 1745. One of his first observations was the effect of points "in drawing off and throwing off the electrical fire." He proposed the one-fluid theory of electricity, some-

¹ See any text on astronomy for further explanation of aberration.

what similar to the caloric theory of heat. This theory supposed that all bodies contained naturally a certain amount of the electric fluid. When a body had an excess of the fluid, it exhibited one of the two kinds of electrification, which Franklin chose to call for this reason *positive*; when it had a deficit, it exhibited the other kind, which he called *negative*. Certain features in the appearance of electric sparks led him to identify his positive electrification with that which had been called vitreous. (In the light of the modern electron theory, we should prefer that he had applied the signs in the opposite order!)

About 1750, Franklin began to speculate on the identity of electricity and lightning, pointing out many similarities and proposing, by means of a pointed iron rod, to "draw off the fire from a cloud." Franklin's writings were published in Europe, and in 1752 Dalibard tried the experiment in Paris, confirming Franklin's prediction. A short time later, Franklin performed the famous kite experiment, so well known to every schoolboy; this led to his study of atmospheric electricity and to his invention of the lightning rod. Franklin's researches occupied but a small portion of his long and busy life, but they were sufficient to give him a high standing among the scientists of the world.

Quantitative researches in electricity began with Cavendish and Coulomb. Cavendish is known not only for his work in electrostatics but also for his researches in chemistry and for the well-known "Cavendish experiment," in 1798, in which he determined the constant of gravitation. His electrical researches were very extensive, but most of his work remained unknown, for he published only one paper of importance. He left behind a large amount of material, however, in the form of manuscript notes, and these were edited and published in 1879 by Maxwell. In these experiments, Cavendish proved the inverse-square law of electrostatic force; measured capacity, recognized the principle of the condenser, and measured the specific inductive capacity of several substances; arrived at a reasonably clear idea of the quantity which we now call "potential"; and anticipated Ohm's law by 50 years. Had these important measurements been communicated to his scientific contemporaries, the history of electricity might have been substantially modified. As it is, the credit of *discovery* seems fairly to belong to others, for a discovery is of no importance to any one else if it is kept secret.

Coulomb's work in electricity grew out of his development of the torsion balance, originally used for studying the torsional elasticity of wires. In the period 1785-1789, he published seven papers on electricity and magnetism in the *Memoires de l'Academie Royale des*

Sciences. In these papers, he showed by means of the torsion balance that electrostatic forces obey the inverse-square law; that, on conducting bodies, the charge exists only on the surface; and that the capacity of a body is independent of the nature of the material of which it is composed. He advocated the two-fluid theory of electricity.

20. Close of the Second Period.—The end of the eighteenth century found rival theories contending in each of three of the subdivisions of physics: the caloric vs. the kinetic theory in heat; the corpuscular vs. the undulatory theory in light; and the one-fluid vs. the two-fluid theory in electricity. The very fact that these issues were raised in rather clean-cut fashion is an indication of the tremendous strides that had been taken since Galileo. But most important of all, men had learned the value of experiment and observation and the fallacy of blindly following "authority."

During the next century the caloric theory was definitely ruled out, apparently never to return. The corpuscular theory of light seemed likewise to be effectively disproved by many new experiments. But further progress in regard to the nature of electrification waited, for the most part, until the twentieth century.

THIRD PERIOD (1800-1890 A.D.): THE RISE OF CLASSICAL PHYSICS

21. The Nineteenth Century in Physics.—So much was added to physical knowledge during the nineteenth century that an adequate history of this period would almost constitute a textbook of physics. We can only comment briefly upon the principal lines of advance and a few of the most important discoveries, selecting especially those that form the background for the characteristic advances of the present century.

In mechanics, there was Hamilton, who discovered in the "Hamiltonian equations" a new form of the equations of motion which is especially convenient in attacking theoretical questions. The theory of the motion of rigid bodies, including the gyroscope, was worked out as well as the mathematical theory of elasticity. The subject of hydrodynamics, dealing with the motion of fluid of all sorts, was developed. In dealing with the flow of viscous fluids, however, only simple problems could be solved; extensive study of such fluids, by half-empirical methods, was not made until during the present century, after the invention of the airplane.

The work in other fields was more striking. The most significant discoveries and advances were: the establishment of the kinetic theory of heat, and the development of the kinetic theory of gases; the

victory (?) of the wave theory of light over the corpuscular theory; the formulation of the general law of the conservation of energy; the discovery of the second law of thermodynamics; and, above all, the discovery, by Faraday and others, of the whole range of electromagnetic phenomena, culminating in Maxwell's theory of the electromagnetic field and of light.

Of these lines of advance we shall select three for detailed presentation, choosing those which bear more or less directly upon modern developments in physics; and as typical scientists of the period, we shall discuss especially Faraday and Maxwell.

22. Heat and Energy.—The law of the conservation of energy is one of the most fundamental and far-reaching of all physical laws, and yet, curiously enough, it is of comparatively recent origin, for it was not announced until the middle of the nineteenth century. As exemplified in mechanics, it had been recognized during the eighteenth century, in the theory of the "vis viva"; but its announcement as a law of universal application awaited experimental work demonstrating the definite equivalence of heat and mechanical work.

The first *qualitative* experiment bearing on the nature of heat was performed in 1798 by Count Rumford, an American who had fled to England in 1775 and eventually became a sort of military engineer to the Bavarian government. He became impressed by the large amount of heat that was developed in boring cannon, and he performed experiments indicating that this heat was too great to be accounted for plausibly by the caloric theory. He could find no loss of weight, when the chips made by boring were included, and showed that the specific heat of the chips was the same as that of the block from which they had come. He concluded that heat "cannot possibly be a material substance" such as caloric but must be a form of "motion."

A still more difficult experiment for the caloric theory to explain was one performed by Sir Humphry Davy, Director of the Royal Institution, which had been founded by Count Rumford. Davy rubbed together two pieces of ice in a vacuum surrounded by walls kept below the freezing point and found that the ice was melted. Here the mechanical work of rubbing accomplished exactly the same effects that could have been produced by the addition of a certain quantity of heat from the outside, yet there was no way in which caloric could have entered the ice.

The majority of the supporters of the caloric theory were, however, unconvinced. Even Carnot (1796-1832), the founder of the modern science of thermodynamics, when he proposed the now famous Carnot's cycle in 1824, based his reasoning on the caloric theory. A given

quantity of caloric "falling" from a higher to a lower temperature was analogous to a given quantity of water falling from a higher to a lower level; each was capable of producing motive power. The kinetic theory had to wait for a *quantitative* experiment.

In 1842, R. J. Mayer (1814-1878) published a paper¹ in which, partly on philosophical grounds, he announced the equivalence of heat and energy, and from data on the specific heats of a gas he deduced a value for the mechanical equivalent of heat. Meanwhile, Joule (1818-1889), in England, unacquainted with Mayer's work, was carrying on a very careful series of experiments in which he converted the mechanical energy of a falling weight into heat by a paddle wheel revolving in water and thus determined that 778 foot-pounds of work would raise 1 pound of water 1°F. Joule announced his results at a meeting of the British Association for the Advancement of Science in 1847. The paper would have passed almost unnoticed, had it not been for William Thomson, later Lord Kelvin, who, grasping the real significance of the proposed theory, by his discussion made the paper the event of the meeting.

Quite independently of the work of Mayer and of Joule, Helmholtz (1821-1894) in 1847 read a paper before the Physical Society in Berlin on "Die Erhaltung der Kraft," in which, on the basis of the impossibility of perpetual-motion machines, he announced the law of the conservation of energy. The paper was rejected for publication by the editor of the *Annalen der Physik*! It was later published in pamphlet form.

The caloric theory could not withstand these attacks, and by 1850 the mechanical theory of heat and the doctrine of the conservation of energy were generally accepted. The science of heat then proceeded to grow apace. The second law of thermodynamics was announced by Clausius (1850) and, in another form, by Kelvin (1851), and in 1854 Kelvin proposed the thermodynamic scale of temperature. Thus was developed the highly successful "classical theory" of heat. We shall discuss later the facts that this theory, in turn, was unable to explain and the way in which the study of these facts led to the development of the quantum theory.

23. Light.—The revival of the wave theory of light, begun by Thomas Young (1773-1829), is one of the most important features of the history of the nineteenth century. Young pointed out that the dividing of a beam of light into a refracted ray at the interface between two mediums was to be expected from the wave theory but

had not been satisfactorily explained on the corpuscular theory. In 1801, he presented to the Royal Society a paper "On the Theory of Light and Colors," in which he proposed the principle of the interference of two wave trains as an explanation of Newton's rings and the colors of thin plates. From Newton's measurements of the thickness of the air layers necessary to produce the several colors, Young was enabled to compute wave lengths. In subsequent papers, he described the interference fringes which he had observed by placing hairs or silk threads in front of a narrow slit illuminated from the rear; he announced the change of phase on reflection; he explained diffraction bands by the principle of interference, and he showed that the spacing of these bands gave values of the wave length agreeing with those obtained from Newton's rings and that, therefore, both phenomena must be due to a common cause. Again, *quantitative* measurements became an indispensable link in the chain of reasoning.

But the dogmatic spirit in regard to scientific matters was not yet dead. Young's paper aroused a storm of protest, even of derision and abuse. He was attacked not by the church, as was Galileo, but by some of his scientific, or, more probably, pseudoscientific, contemporaries. His chief assailant was Henry Brougham, afterward Lord Chancellor of England, who "reviewed" Young's papers in the *Edinburgh Review*. The nature of Brougham's attack is indicated by the following quotation:

We wish to raise our feeble voice against innovations that can have no other effect than to check the progress of science and renew all those wild phantoms of the imagination which Bacon and Newton put to flight from her temple. We wish to recall philosophers to the strict and severe methods of investigation . . .

Although Young replied at length in a privately published pamphlet, it was a long time before public opinion was willing to receive his theories with an open mind.

In 1815, a Frenchman, Fresnel (1788-1827), rediscovered the phenomenon of interference, performing the famous experiment with the two mirrors. A few years later he developed a mathematical theory of such phenomena (1818-1819). He also explained the polarization of light by assuming that the light vibrations in the ether are *transverse* to the direction of propagation of the light rather than longitudinal. He did not know that this suggestion had already been made by Young in a letter to Arago written in 1817. Fresnel supported the explanation by showing experimentally that two plane-polarized beams of light cannot interfere at all if their planes of

polarization are perpendicular to each other. Phenomena of polarization had been known to Newton, the polarization of light by Iceland spar being discovered by Bartholinus in 1669. Newton had tried to fit these phenomena into the corpuscular theory by assuming a sort of structure in the corpuscles, but the explanation was not convincing, and polarization had remained an enigma to both theories of light. Even Fresnel's explanation seemed almost to demand the impossible, for it required the ether to be a solid, or at least to have such properties of a solid as are necessary for the transmission of transverse waves, i.e., the properties of rigidity and density. And yet the planets must move through this "solid" pervading all space, with no measurable changes in their periods of revolution!

Experimental evidence for the wave theory continued to accumulate. Finally, in 1850, Foucault performed a crucial experiment in its favor by showing, with his well-known revolving-mirror apparatus for measuring the velocity,¹ that light travels more slowly in water than in air, as it should on the wave theory, in order to account for the relative refractive index of these two mediums, whereas, on the corpuscular theory, the reverse should be the case.²

From 1850 until the end of the third period (1890), the wave theory held the field undisputed. The frequent assertions that the corpuscular theory was finally disposed of certainly seemed justified, particularly after the development of Maxwell's electromagnetic theory of light and its experimental verification. Yet the corpuscular theory was not actually dead. It was only sleeping.

Some important discoveries in light from 1800 to 1890 not previously mentioned are:

	DISCOVERER
Dark lines in the solar spectrum	Fraunhofer
Three-color theory of vision (1807)	Young
Rotary polarization of quartz (1811)	Arago
Polarization of scattered light (1813)	Arago
Rotary polarization by liquids (1815)	Biot
Light sensitivity of silver bromide (1826)	Balard
Change of conductivity of selenium on illumination (1837)	Knox
Doppler effect (1842)	Doppler
Foundation of spectral analysis (1859)	Kirchhoff and Bunsen

24. Electricity and Magnetism.—The history of electricity during the nineteenth century is so extensive that even a sketchy outline would fill a small volume. We shall, therefore, discuss little besides

¹ See EDSEER, "Light for Students."

² See any textbook on optics.

the fundamental discoveries of the opening decades and then the work of Faraday and that of Maxwell, which are so closely related to each other and to recent developments in physics that we can best present the part of electricity and magnetism in which we are particularly interested by giving an account of the contributions of these two men.

While the mathematical theory of electrostatics and of magnetism was being elaborated by Laplace, Green, and Poisson, fundamental discoveries were made in regard to electric *currents*. Galvani, in 1786, as a result of a chance observation that a frog's leg kicked convulsively when connected with the terminal of an electric machine, was led to an extensive study of "animal electricity." In the course of these experiments he observed that if the frog's leg was so suspended that the exposed nerves touched a metal plate, say silver, then, a contraction of the muscle occurred whenever the foot touched another metal, say iron. He even observed slight muscular contraction when both plates were of the same kind of metal. This led him to believe that the nerve was the source of electricity and that the metal served simply as conductor. Volta later found that potentials could also be produced using inorganic materials and, in 1800, he described the first battery for producing an electric current—the historically famous voltaic "pile," consisting of zinc and copper plates placed alternately and separated by blotting paper moistened with brine. He also described a battery consisting of cups containing brine or dilute acid connected by copper and zinc strips joined together.

Volta ascribed the effect to the contact of two dissimilar metals. We now know, however, that the current that can be caused in this manner to flow in a closed circuit is due to chemical action at the contacts of the metals with the electrolyte; the effect at the junction of the metals gives rise to potential differences of a different nature, the "contact difference of potential" or "Volta effect," which adds up to a total of zero in any closed circuit.

This new source of electricity was received with a great deal of interest. A few weeks after hearing of Volta's work, Nicholson and Carlisle accidentally discovered the decomposition of water by the electric current. Thinking to secure better contact between two wires forming part of the circuit, they had joined the ends of the wires by a drop of water. At once they observed the formation of a gas, which they recognized as hydrogen. This was the beginning of the study of electrolysis. During this same period the heating effect of the current and the arc light were discovered.

It was early suspected that there was some relation between electricity and magnetism, but the first significant discovery was made

in 1820 by Oersted, who found that a magnetic needle tends to set itself at right angles to a wire through which an electric current is flowing. Soon after, Biot and Savart discovered the law for the field of a long straight current, and, toward the end of 1820, Biot proposed, on the basis of a special experiment, the formula for the field due to a current element that is commonly used today and is often miscalled "Ampere's formula."¹ Soon after, the brilliant French physicist, Ampere (1775-1836), on the basis of just enough crucial experiments, showed that a closed current is equivalent in its magnetic effects to a magnetic shell. Then, reversing his line of thought, he suggested that magnetism itself might be due to currents circulating in the molecule. He also discovered the action of a magnetic field on a current. Thus, within five years of the first discovery, the foundations of electro-magnetism had been completely laid.

25. Michael Faraday.² (a) *Biographical Sketch*.—Michael Faraday was born in 1791, in a small village near London. He was the son of a blacksmith, James Faraday. Being required to assist his mother in providing for the family, he was engaged in 1804 as errand boy to a bookseller and stationer, and in the following year he was formally apprenticed to his employer to learn the art of bookbinding. During this apprenticeship, Faraday made good use of his spare time by reading some of the books that passed through the shop. He was particularly interested in works on science, and in connection with his reading he showed one of the important characteristics³ of the great investigator-to-be by performing such of the simple experiments described "as could be defrayed in their expense by a few pence per week."

Aside from his own reading, *Faraday's only scientific education consisted in a dozen lectures on natural philosophy by a Mr. Tatum and four lectures on chemistry by Sir Humphry Davy, in the winter of 1812.* Submitting the very careful and neatly written notes which he made of these lectures "as proof of his earnestness," he made bold to apply to Sir Humphry Davy for a position, however menial, at the Royal Institution of which Davy was then director. Davy was so pleased with the letter and the notes that in March, 1813, Faraday was

¹ I.e., $dH = \frac{1}{r^2} \int \sin \theta \, ds$. Cf. BIOT, J. B., "Precis elementaire de Physique experimentale," 1824. Ampere appears to have assumed that the force-action between two current elements must necessarily lie along a line drawn through the elements, which is not the case if Biot's formula is used for the magnetic field.

² See THOMPSON, SYLVANUS P., "Michael Faraday: His Life and Work."

³ In later life, he wrote: "I was never able to make a fact my own without seeing it."

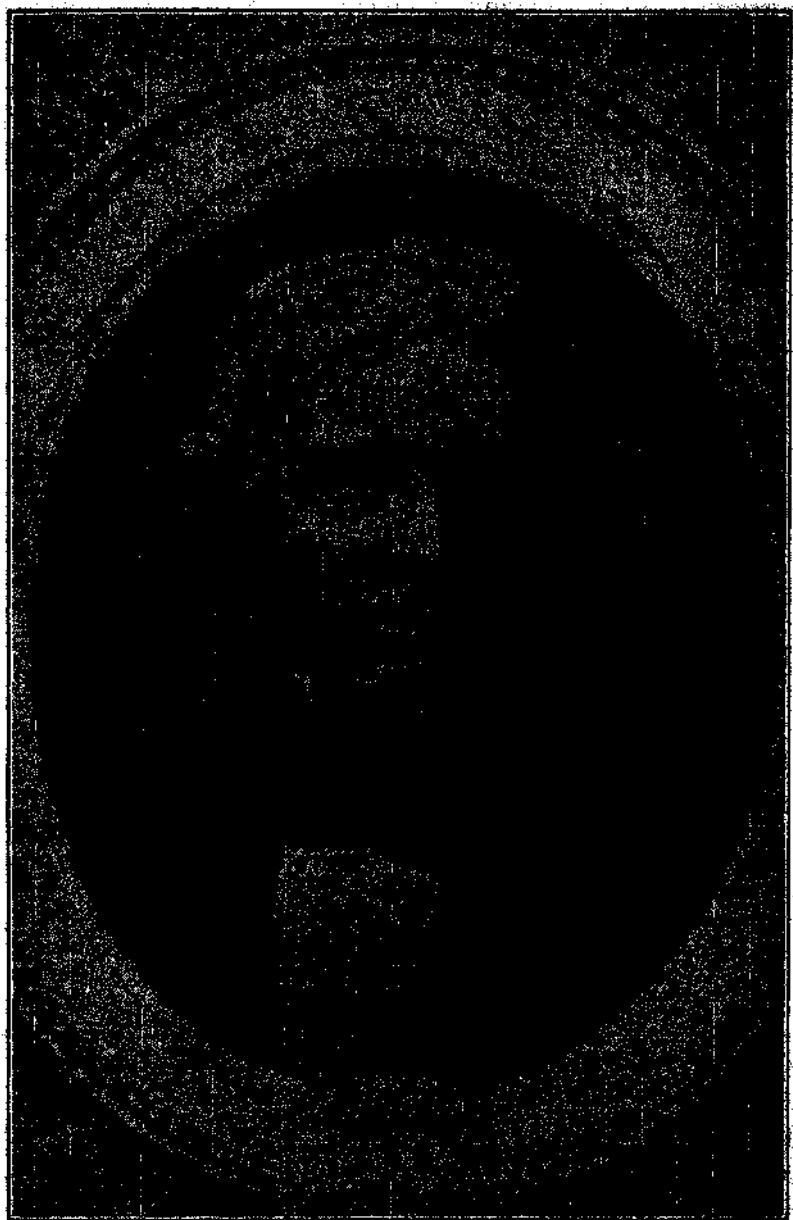


PLATE 3.—Faraday.

engaged as apparatus and lecture assistant at 25 shillings per week. In October, 1813, he accompanied Sir Humphry and Lady Davy on a trip to the Continent, which took them to many of the important scientific centers of Europe. Assistant though he was, Faraday impressed others because of his modesty, amiability, and intelligence; said one writer, "We admired Davy, we *loved* Faraday."

On returning to England, under Davy's encouragement, Faraday soon began original investigations, initially in chemistry. From 1816 to 1819 he published 37 papers. These were concerned with such subjects as the escape of gases through capillary tubes, the production of sound in tubes by flames, the combustion of the diamond, and the separation of manganese from iron. About 1820, he began his electrical researches. These, and others growing out of them, continued for nearly 40 years.

Almost his entire scientific life was spent at the Royal Institution. In 1825, he was made Director of the Laboratory. Declining offers of positions elsewhere, turning away professional occupations which might have made him wealthy, he gave to his science and to the institution he served a devotion seldom if ever equaled. The secret of his success, which brought him, during his lifetime, honors from all over the scientific world and which immortalized his name by the long list of scientific discoveries ascribed to him, is, perhaps, to be found in some excerpts from his many notes:

Aim at high things, but not presumptuously.

Endeavor to succeed—expect not to succeed.

It puzzles me greatly to know what makes the successful philosopher. Is it industry and perseverance with a moderate proportion of good sense and intelligence? Is not a modest assurance or earnestness a requisite? Do not many fail because they look rather to the renown to be acquired than to the pure acquisition of knowledge . . . ? I am sure I have seen many who would have been good and successful pursuers of science, and have gained themselves a high name, but that it was the name and the reward they were always looking forward to—the reward of the world's praise. In such there is always a shade of envy or regret over their minds and I cannot imagine a man making discoveries in science under these feelings.

The reader is urged to study carefully Faraday's life and works, particularly to read, as unexcelled examples of scientific expositions, portions of his "Experimental Researches in Electricity and Magnetism." We can mention here only a few of his most important discoveries.

(b) *The Principle of the Motor*.—Faraday had been interested in electromagnetism since April, 1821, when Wollaston attempted, at the

Royal Institution, to make a wire carrying an electric current revolve around its own axis when the pole of a magnet was brought near. The experiment was unsuccessful, but the phenomenon excited Faraday's interest, and he determined to make a study of it. First, he read what had been done by others and repeated many of their experiments. In the course of these experiments, he observed that, when the magnetic pole was brought near the wire, "the effort of the wire is always to pass off at right angles from the pole, indeed to go in a circle around it" ¹

The following day he wrote in his laboratory notebook:

Apparatus for revolution of wire and magnet. A deep basin with a bit of wax at bottom and then filled with mercury. A magnet stuck upright in wax so that pole [is] just above surface of mercury. Then piece of wire, floated by cork, at lower end dipping into mercury and above into silver cup.

On passing a current through the wire, it revolved *continuously* around the magnet. This was the first electric motor!

(c) *Electromagnetic Induction*.—Oersted's experiment and subsequent developments had clearly shown how "to produce magnetism by electricity." Faraday seems to have held it as one of the tenets of his scientific philosophy that every physical relation (of cause and effect) has its converse. If electricity can produce magnetism, then magnetism should produce electricity. His repeated attempts to accomplish this failed. For example, in 1825, he tried what seemed to be the obvious converse by looking for an electric current in a helix of wire coiled around a magnet. Later, he tried to find a current in a wire placed near another wire carrying current. Other scientists were looking for similar effects but without success. They were all looking for the production of a *steady* current.

But several times investigators were very near to the discovery of induced currents. In 1824, Arago observed the damping of the vibrations of a magnetic needle suspended over a copper plate. This observation was extended by causing the needle to revolve by revolving the copper plate underneath it, air disturbances being, of course, eliminated. It was shown that this "dragging" effect was greater, the greater the electrical conductivity of the spinning plate. Even the effect of radial slits in the copper disk, in reducing the dragging action on the magnet, was observed. Suggestive as these experiments were, however, the true explanation remained undiscovered.

In the summer of 1831 Faraday attacked the problem for a fifth time. This time, instead of placing a *permanent* magnet inside a

¹ Quotation from Faraday's laboratory notebook, Sept. 3, 1821.

helix, he procured a soft iron ring 6 inches in external diameter, on which he wound two coils of copper, *A* and *B*, "separated by twine and calico." To detect a possible current in coil *B*, he "connected its extremities by a copper wire passing to a distance and just over a magnetic needle." When coil *A* was connected to a battery, there was "a sensible effect on the needle. It *oscillated* and settled at last in *original position*. On breaking connection of side *A* with battery, again a disturbance of the needle." Slight as these momentary effects were, Faraday recognized their importance, although he too had been looking for a *continuous* effect. On Aug. 30, he writes, "May not these transient effects be connected with causes of difference between power of metals at rest and in motion in Arago's experiments?"

From this slender clue, Faraday proceeded rapidly to the discovery of the real effect. On the "third day" of his experiments, he wound a coil of wire around an iron cylinder and placed the cylinder so as to join the *N* pole of one permanent magnet with the *S* pole of another. The coil was connected to a galvanometer:

Every time the magnetic contact at *N* or *S* was mad? or broken there was a magnetic action at the indicating helix [*i.e.*, galvanometer]—the effect being, as in former cases, not permanent but a mere momentary push or pull.

On the fourth day, he showed that the presence of iron was not necessary; that the effect could be produced by the action of one helix on another. On the fifth day:

A cylindrical bar magnet . . . had one end just inserted into the end of the helix cylinder; then it was quickly thrust in the whole length and the galvanometer needle moved; then pulled out and again the needle moved, but in the opposite direction. The effect was repeated every time the magnet was put in or out, and therefore a wave of electricity was so produced from mere *approximation* of a magnet and not from its formation *in situ*.

At last! He had "converted magnetism into electricity." The essential requisite was *relative motion*, or a *change* of condition. On the ninth day, he produced a continuous current by turning a copper disk between the poles of a powerful electromagnet, the periphery of the disk being connected to its axis through an indicating galvanometer. This was the now well-known Faraday disk dynamo, the *very first* dynamoelectric machine.

Thus, after only a few days' work in his laboratory, following, however, years of patient and persistent experiment, Faraday had discovered a phenomenon for which the greatest scientists of his time **had sought in vain**—electromagnetic induction.

Following this discovery, Faraday devised and tried various electric machines to test and extend his newly discovered principle. One of these machines, consisting of a rotating rectangle of wire *with a commutator attached*, is the prototype of the modern dynamo. But his interest was always in pure science, for he writes:

I have rather, however, been desirous of discovering new facts and relations dependent on magnetoelectric induction, than of exalting the force of those already obtained; being assured that the latter would find their full development hereafter.

Being unacquainted with mathematical symbols and methods, Faraday always sought to explain his discoveries and to extend his researches by purely physical reasoning. To the mathematician, the law of magnetic attraction

$$F = \frac{m_1 m_2}{r^2}$$

may have been a sufficient explanation of the phenomenon. To Faraday, this gave a statement only of the *magnitude* of the magnetic forces; *it left the phenomenon itself quite unexplained*. Accordingly, he insisted that two magnetic poles, or two electric charges, could act on each other *only if the medium between the two played some important part in the phenomenon*. This insistence on the importance of the medium ultimately led him to the very fruitful concept of lines of force and of the "cutting" of these lines as essential to electromagnetic induction. At first qualitative, this concept was developed by Faraday into an essentially quantitative form, although it was first stated in mathematical language by F. Neumann in 1845. Commenting on Faraday's laws of electromagnetic induction, Maxwell wrote:

After nearly a half-century . . . , we may say that, though the practical applications of Faraday's discoveries have increased and are increasing in number and value every year, no exception to the statement of these laws as given by Faraday has been discovered, no new law has been added to them, and Faraday's original statement remains to this day the only one which asserts no more than can be verified by experiment, and the only one by which the theory of the phenomena can be expressed in a manner which is exactly and numerically accurate, and at the same time within the range of elementary methods of exposition.

(d) *The Laws of Electrolysis*.—Faraday next turned his attention to proving that "Electricity, whatever may be its source, is identical in its nature." He found, for example, that electricity from a friction machine would deflect a galvanometer and would cause chemical

decomposition just as would electricity produced by chemical action. This led him into the field of electrolysis. He found that many substances, such as certain chlorides and sulphates, are nonconductors when solid but are good conductors when melted, and that in the molten state they are decomposed by the passage of current. This showed that water was *not* essential to electrolysis. To clarify description of his experiments, he introduced the terms "electrode," "anode," "cathode," "ion," "anion," "cation," "electrolyte," "electrochemical equivalent," etc. A quantitative study of the phenomena resulted in his discovery of the laws of electrolysis that bear his name and which are the basis of all present-day work in that field.

Further, Faraday clearly recognized that a definite quantity of electricity is associated with each atom or ion in electrolysis. Had he been able to determine the number of atoms in unit mass of any substance, he would have anticipated, by 60 years, the determination of the fundamental charge e . For he says:

Equivalent weights of bodies are simply those quantities of them which contain equal quantities of electricity; . . . it being the *electricity* which determines the combining force. Or, if we adopt the atomic theory or phraseology, then the atoms of bodies which are equivalent to each other in their ordinary chemical action, have equal quantities of electricity naturally associated with them.

(e) *The Conservation of Energy*.—In connection with a proof of the fact that the electricity from the voltaic pile results from chemical action and not from mere contact of one substance with another, Faraday stated clearly the doctrine of the *conservation of energy several years before the statement of Helmholtz. In 1840, he wrote:

The contact theory assumes that a force which is able to overcome a powerful resistance . . . can arise out of nothing . . . This would indeed be a creation of power, and is like no other force in nature. We have many processes by which the form of the power is so changed that an apparent conversion of one into the other takes place . . . But in no case is there a pure creation or a production of power *without a corresponding exhaustion of something to supply it*.

(f) *The "Faraday" Effect*.—Reference has already been made to Faraday's abhorrence of the doctrine of "action at a distance." He believed that, if two electric charges attract each other, the medium **between** the two plays some important role. Presumably, therefore, **the medium between two such charges is in a different state than it would be if the charges were not present; and if so, such an altered state should be detectable by observing the alteration in some physical**

property of the medium. As early as 1822, Faraday had experimented with a beam of polarized light passing through a transparent solution carrying a current, to see whether the current caused any "depolarizing" action. Although he repeated the experiment several times in subsequent years, the results were all negative. In 1845, he returned to the problem, but still with negative results. He then tried solid dielectrics between plates of metal foil connected to a powerful electric machine to see whether, under electric strains, they would show any optical effects. No results!¹

Faraday then substituted a magnetic field for the electrostatic field to see whether the former would cause any depolarizing action on the beam of light. Various substances were tried but still with negative results. Finally, he placed in the magnetic field a very dense piece of lead glass, which he had made many years earlier. When the magnetic lines were parallel to the direction of the beam of polarized light, he observed that the plane of polarization was rotated. At last, he had found a relation between magnetism and light. This magnetic rotation is now known as the "Faraday effect." Again, his persistent search, maintained during 20 years of repeated failures, was rewarded by the discovery of an effect in the existence of which he had the most sublime confidence.

(g) *Miscellaneous*.—Among Faraday's other researches may be mentioned: numerous investigations in chemistry; the liquefaction of several gases formerly thought "permanent"; the diffusion of gases through solids; self-induction; certain fundamental properties of dielectrics; diamagnetism; distinction between anode and cathode in the electric discharge through gases at low pressure; vibration of plates; regelation of ice; alloys of steel; and optical glass.

Well may this simple, modest, self-taught philosopher be given a conspicuous place among the great benefactors of mankind.

26. Joseph Henry (1799-1878).—Any account of Faraday's work, however brief, should be accompanied by at least a mention of the researches of the American physicist, Joseph Henry, whose memory is honored by the name of the unit of inductance, the henry, which bears to electrokinetics a relation identical with that of the farad to electrostatics. Had Henry been able to experiment continuously, and with more resources, instead of only during a summer vacation of 1 month while teaching mathematics at Albany Academy, and then only with such apparatus as he could make with his own hands, he would undoubtedly have anticipated Faraday in the discovery of electro-

¹ Years later (1875) this effect was found by Kerr [*Phil. Mag.*, vol. 1, p. 337 (1875)].

magnetic induction, including the phenomena of self-induction. In all of his work, furthermore, he was greatly hampered by his isolation from the scientific atmosphere of Europe.

Henry was interested especially in the design and use of electromagnets. He constructed the first electric motor operating by an electromagnet, which rocked back and forth between two permanent magnets. He found that, for maximum tractive effect, the cells of the battery and also the "spools" of the electromagnet should be connected in series if the magnet were a long distance from the battery, but they should be connected in parallel if the wires joining the magnet to the battery were short. His work on electromagnets led *directly* to the commercial development of the telegraph.

27. James Clerk Maxwell (1831-1879).—It would be difficult to pick out two eminent scientists whose beginnings differed from each other more than did Maxwell's and Faraday's. Faraday came of very humble parentage; Maxwell, from a long line of distinguished ancestors. Faraday's early life was lived almost in poverty; Maxwell's family had abundant means. Faraday received only the most rudimentary education; Maxwell was given every advantage of school and university. They differed also in their aptitude for scientific work. Faraday was one of the greatest exponents of experimental science that the world has ever seen; whereas Maxwell, although likewise an able experimenter, is one of the greatest figures in the history of theoretical physics. And yet both made indispensable and mutually supplementary contributions to the classical theory of electromagnetics.

Maxwell was born in Edinburgh in 1831. At the age of ten he was sent to the Edinburgh Academy, where he was a friendly boy, though never quite amalgamating with the rest. But, however strange he sometimes seemed to his companions, he had three qualities which they could not fail to understand: agile strength of limb, imperturbable courage, and profound good nature.¹

When he left the Academy in 1847, he was "first in mathematics and in English and nearly first in Latin." Then, after 3 years in the University of Edinburgh, he entered Trinity College, Cambridge, from which he graduated in 1854 with high honors.

Maxwell early showed extraordinary interest in both theoretical and experimental research in physics. At the academy, he invented a means of drawing certain types of oval curves, and a few years later he published a paper on "The Theory of Rolling Curves" and another on "The Equilibrium of Elastic Solids"—all this before he was nineteen years old! During these same years he was also busy with

¹ GLAZEBROOK, "James Clerk Maxwell and Modern Physics."



PLATE 4.--Maxwell.

experiments of many sorts, especially in his little laboratory in a garret on the family estate at Glenlair, where he spent his vacations.

After 4 years at Aberdeen, he was Professor for 5 years at King's College, London (1860-1865); from here some of his most important papers were published, such as "Physical Lines of Force" (1862) and his greatest paper, "A Dynamical Theory of the Electromagnetic Field." After a retirement of several years, he was elected in 1870 to the newly founded professorship of experimental physics at Cambridge. In this capacity, he superintended the planning and equipment of the now famous Cavendish Laboratory, of which he was director until his untimely death in 1879.

A large proportion of Maxwell's papers, over 100 in number, may be grouped under three headings: color vision; molecular theory; and electromagnetic theory.

The work on color vision was undertaken to make a quantitative study of the physical facts pertinent to the theory of color sensations proposed by Thomas Young, according to which any luminous sensation is the result of exciting in the eye three primary sensations, red, green, and violet. For this purpose Maxwell invented a "color box," by means of which he could mix spectral colors.

Maxwell's work on molecular physics is very extensive. He discovered and, in part, established theoretically the law of the distribution of velocities among the molecules of a gas ("Maxwell's law"). He showed that when two gases are at the same temperature, the mean kinetic energy of translatory motion of their individual molecules is the same in both gases. From the kinetic theory of viscosity, he drew the surprising conclusion that the viscosity of a gas should be independent of its density so long as the mean free path is not too large, and he verified this conclusion by experiment. He brought to bear upon the whole subject mathematical methods "far in advance of anything previously attempted on the subject"; indeed, he is the co-founder with Clausius (1822-1888) of the kinetic theory of matter.

In the electromagnetic theory, Maxwell's great contributions were the "displacement currents" and the formulation of the general equations of the electromagnetic field, which led to the electromagnetic theory of light. In the preface to his treatise "Electricity and Magnetism," he makes the interesting remark:

Before I began the study of electricity I resolved to read no mathematics on the subject till I had first read through Faraday's "Experimental Researches on Electricity."

He became convinced that Faraday was right in regarding the dielectric as the true seat of electrical phenomena and in supposing that it acted by becoming electrically polarized, the positive ends of its molecules pointing on the whole with the field and the negative ends in the opposite direction. The term "dielectric," as used here, must be understood to include a tenuous medium or ether filling all space, even in what we call a vacuum. He drew the conclusion that when the polarization changes, this change must involve a displacement of electricity, and so there must exist in the dielectric, while the change is going on, a current having the same magnetic properties as the current in a conductor.

This assumption of displacement currents opened the way for the deduction of Maxwell's famous equations of the electromagnetic field. It is interesting, however, that he was first led to these equations through a mechanical analogy, *i.e.*, in studying the behavior of a *mechanical system* filling all space, which would be capable of causing the observed electrical and magnetic phenomena. He showed (1862) that his hypothetical medium would be capable of transmitting transverse vibrations with a speed equal to the ratio-of the electromagnetic to the electrostatic unit of charge. Although he did not take his model too seriously, he nevertheless remarks, that the ratio of the units

. . . agrees so exactly with the velocity of light calculated from the optical experiments of M. Fizeau, that we can scarcely avoid the inference that *light consists in the transverse undulations of the same medium, which is the cause of electric and magnetic phenomena.*

The theory was restated, without reference to any particular model, in his great paper of 1864, in which he says:

The theory which I propose may therefore be called a theory of the *Electromagnetic Field*, because it has to do with the space in the neighborhood of the electric or magnetic bodies, and it may be called a *Dynamical Theory*, because it assumes that in that space there is matter in motion by which the observed phenomena are produced.

In 1873, Maxwell published his "Treatise on Electricity and Magnetism," which ranks with Newton's "Principia" as one of the most important books in all science.

The subject of electromagnetic waves will be discussed further in the next chapter along with certain other results of electromagnetic theory that will be needed later. Here we shall add only a few words concerning the later history of the subject during the closing decades of the last century.

28. The Completion of Electromagnetic Theory.—The *physical ideas* underlying Maxwell's new theory were left none too clear by him. In his treatise, we find the assumption that all space is full of incompressible "electricity"; in a conductor this electricity can move freely (except for ohmic resistance), thus constituting an electric current, but in a dielectric "there is a force which we have called electric elasticity which acts against the electric displacement and forces the electricity back when the electromotive force is removed/' This is clear enough. But what is the origin of this "electromotive force?" And in what does electrification consist? According to Maxwell's theory a charged conductor has neither more nor less electricity on it or in its neighborhood than when it is not charged. This point in Maxwell's theory became still more obscure when in 1876 the American physicist, Rowland, showed experimentally that a moving charged conductor is surrounded by a magnetic field, the moving electrification evidently constituting a current.

The *mathematical theory*, on the other hand, was slowly developed by others, especially by H. A. Lorentz, and was shown to give a good account of all electric and magnetic phenomena and of the principal properties of light. In Germany, stimulated by Helmholtz, Hertz set out to search experimentally for the magnetic effects of Maxwell's displacement currents and in 1887 discovered waves that were undoubtedly of electrical nature. Later, it was shown that the speed of propagation of these waves is the same as that of light. Speculation as to the nature of the displacement currents in a vacuum then gradually died out, until today we speak only of electric and magnetic "fields" governed by the Maxwell-Lorentz equations. This development in electromagnetic theory illustrates a general tendency, notable during the last half-century but regretted by many, for the fundamentals of physics to become an abstract mathematical theory unsupported by underlying concrete ideas. Further examples are afforded by Einstein's principle of relativity, to be described in Chap. IV, and in the foundations of wave mechanics.

CHAPTER II

ELECTROMAGNETIC WAVES AND MOVING CHARGES

Before taking up the main theme of this book, we shall discuss, in this chapter, a few points in electromagnetic theory which it will be useful for the student to have in mind. Electromagnetic waves and the emission of radiation will be discussed, on the basis of Maxwell's theory, in sufficient detail to enable the student to appreciate more readily the conflict of this theory with the modern quantum theory. Then a few points concerning the fields of moving charges will be taken up.

Few derivations can be given here, for lack of space, but we shall begin by writing down Maxwell's equations.

29. Maxwell's Equations.—In an electromagnetic field there exists at each point, in general, both an electric and a magnetic field strength or field intensity. These quantities are vectors having direction as well as magnitude. To deal with them analytically, let us introduce a set of right-handed Cartesian axes. To save words, it will be convenient, hereafter, to say that any three mutually perpendicular vectors related to each other in direction like the directed lines Ox , Oy , Oz form a *right-handed {orthogonal} set*¹ of vectors.

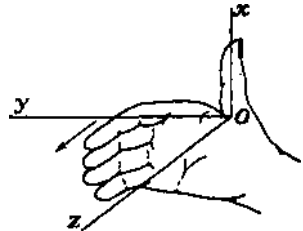


FIG. 1.—Illustrating the right-hand vector rule

Let us now represent the field vectors by their components in the directions of these axes. Let the components of the electric field strength E be denoted by E_x , E_y , E_z , respectively, and those of the magnetic field strength H by H_x , H_y , H_z . Then Maxwell's field equations for a region in which the dielectric constant is e and the magnetic permeability μ , but in which there is no electrical conductivity and no accumulation of electric charge, can be written thus:

$$\frac{dH_z}{dy} - \frac{dH_y}{dz} = \frac{e}{c} \frac{dE_x}{dt} \quad (1a)$$

$$\frac{dH_x}{dz} - \frac{dH_z}{dx} = \frac{e}{c} \frac{dE_y}{dt} \quad (1b)$$

¹ Vectors can be tested for this property with the right hand, as suggested in Fig. 1.

$$\frac{\partial H_y}{\partial x} - \frac{\partial H_x}{\partial y} = \frac{\epsilon}{c} \frac{\partial E_z}{\partial t}, \quad (1c)$$

$$\frac{\partial E_z}{\partial y} - \frac{\partial E_y}{\partial z} = -\frac{\mu}{c} \frac{\partial H_x}{\partial t}, \quad (1d)$$

$$\frac{\partial E_x}{\partial z} - \frac{\partial E_z}{\partial x} = -\frac{\mu}{c} \frac{\partial H_y}{\partial t}, \quad (1e)$$

$$\frac{\partial E_y}{\partial x} - \frac{\partial E_x}{\partial y} = -\frac{\mu}{c} \frac{\partial H_z}{\partial t}, \quad (1f)$$

$$\frac{\partial}{\partial x}(\mu H_x) + \frac{\partial}{\partial y}(\mu H_y) + \frac{\partial}{\partial z}(\mu H_z) = 0, \quad (1g)$$

$$\frac{\partial}{\partial x}(\epsilon E_x) + \frac{\partial}{\partial y}(\epsilon E_y) + \frac{\partial}{\partial z}(\epsilon E_z) = 0. \quad (1h)$$

We shall suppose that c.g.s. units are used and that E is in electrostatic units, whereas H is in electromagnetic units or oersteds; ϵ and μ are therefore, pure numbers, and equal to unity if the field is in a vacuum. The derivatives occurring in the equations are all partial derivatives, since $E_x, E_y, E_z, H_x, H_y, H_z$ may all vary, not only from point to point, but also from time to time, each component being thus a function of the four variables x, y, z, t . The constant c enters into these equations as the ratio of the electromagnetic to the electrostatic unit of charge; as is well known, however, this constant is also equal to the speed of light in vacuum.

The mathematical foundation of electromagnetic theory may then be completed by adding to the field equations the Lorentz force equation, which expresses the familiar value for the force exerted by the field upon a small test charge placed in it. Upon a stationary charge of q electrostatic units, there acts a force of qE dynes in the direction of E ; if the charge is moving with velocity v cm. sec.⁻¹, there is added to the force qE another force of magnitude $qvH \sin \phi/c$ dynes, ϕ being the angle between the directions of v and of H . This additional force acts perpendicularly both to v and to H and in such a direction that the three vectors v , vector component of H perpendicular to v , and the force itself form a right-handed set (see above and Fig. 1).

Of the six Maxwell equations, number (1h) is a mathematical equivalent of Gauss' theorem, familiar in electrostatics. Equation (10) represents the corresponding fact in the theory of magnetism, *i.e.*, that the tubes of the induction μH are closed, each tube having a uniform flux of induction across all of its cross sections. Equations (1d) to (1f) embody Faraday's principle that the electromotive force in a closed circuit, measured in electromagnetic units, equals the negative of the time rate of change of the flux of magnetic induction through the circuit. •

In equations (1a) to (1c), finally, the terms on the right, $e/c \, dE_x/dt$ and so on, represent the components of the density of the so-called "displacement current." In a conductor, as in the interior of a wire carrying a steady current, these terms are replaced by $4\pi j_x$, $4\pi j_y$, $4\pi j_z$, where j_x , j_y , j_z represent the components of the ordinary current density. Maxwell's assumption was, therefore, that, wherever an electric field is undergoing change, there exists something that is related to the magnetic field in the same way as are the ordinary currents in a conductor. The current density that is equivalent in its magnetic effects to the changing electric field has components j'_x , j'_y , j'_z given by the equations

$$j'_x = \frac{\epsilon}{4\pi c} \frac{\partial E_x}{\partial t}, \quad j'_y = \frac{\epsilon}{4\pi c} \frac{\partial E_y}{\partial t}, \quad j'_z = \frac{\epsilon}{4\pi c} \frac{\partial E_z}{\partial t}.$$

No reason can be given for this equivalence of a changing electric field to a set of electric currents; it constitutes an assumption that is to be tested, and so far has withstood the test, by comparison of deductions from it with the results of experiment.

With this interpretation of the right-hand members of Eqs. (1a) to (1c), these equations become mathematically equivalent to the familiar law that the work done in carrying a unit magnetic pole around a closed path is equal to 4π times the total current strength that links with the path.¹

30. Energy and Momentum in the Electromagnetic Field.—In the interaction of material bodies with electromagnetic fields, the bodies commonly gain or lose *energy*. It has been found possible to preserve the law of the conservation of energy by supposing a definite amount of energy to exist in the field as well as in material bodies. This is the usual procedure in physics. Whenever a new phenomenon is encountered, the attempt is made to invent a new form of energy so that the conservation law will hold; and so far it has always proved possible to do this.

The energy in the field is determined by calculating the work done in setting it up. In this way it can be proved quite generally,² and is shown for special cases in elementary textbooks, that energy can be supposed to be distributed throughout an electromagnetic field at the rate of

$$w = \frac{\epsilon E^2 + \mu H^2}{8\pi} \text{ ergs per unit volume,} \quad (2)$$

¹ For deductions of the equations, see, for example, Starling, "Electricity and Magnetism," Chap. XIII (or XIV); Page and Adams, "Principles of Electricity," Chap. XVI.¹

² Cf. SMYTHE, W. R., "Static and Dynamic Electricity," Sec. 13.04.

E being the electric intensity in electrostatic units, H the magnetic intensity in oersteds, ϵ the dielectric constant, and μ the permeability.¹

If the energy is assumed to be distributed in this manner, it must be supposed to move about when the field undergoes changes. From the theory, it can be shown that correct results are obtained if we suppose the energy to stream in the direction of a vector known as *Poynting's vector*. This is a vector drawn perpendicular to E and to H and in such a direction that a right-handed set of vectors is formed by the vector E , the vector component of H perpendicular to E , and Poynting's vector (cf. Sec. 29 and Fig. 2). Its magnitude is assumed to be

$$\Pi = \frac{c}{4\pi} EH \sin\theta \quad (3)$$

where θ denotes the angle between the directions of E and H , and c is the ratio of the electromagnetic to the electrostatic unit of charge, or the speed of light in vacuum.

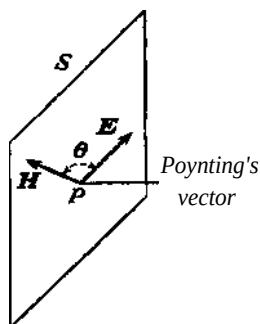


FIG. 2.

This last expression represents the rate at which the energy is flowing. If, at any point P in the field, a small plane surface is drawn parallel to E and to H and hence perpendicular to Poynting's vector (cf. Fig. 2), then the number of ergs of energy crossing each square centimeter of this area per second is given by the magnitude of Poynting's vector as just written. The student can readily verify in simple cases that the correct amount of energy is thereby delivered to the various parts of

electrical circuits; *e.g.*, the heat produced in a resistor by an electric current can be supposed to stream in from the surrounding space in the direction of Poynting's vector.

In mechanics, furthermore, the *principles of momentum* are second only to those of energy in importance. Now electromagnetic fields are capable of changing the momentum of material bodies. In all cases of practical importance it happens that this action does not alter the *total momentum* of all bodies concerned, because a force on one body is offset by a reaction on some other body. There are, however, a few phenomena in which this is not true; the most noteworthy is the phenomenon of light pressure, discussed below. But even in such

¹ This expression is restricted, of course, to free space or to matter that is isotropic and not ferromagnetic. In material bodies, furthermore, w represents free energy, in the thermodynamic sense, rather than stored energy; but this distinction is seldom important.

cases it can be shown¹ that the principles of momentum and of angular momentum can be preserved provided we assign momentum not only to matter but also to the electromagnetic field itself.

The momentum in the field is called *electromagnetic momentum*. It is a vector quantity, and it is assumed to have, at each point in the field, the direction of Poynting's vector at that point; furthermore, the amount of electromagnetic momentum in unit volume, in a region where the dielectric constant is ϵ and the permeability μ , is equal to Poynting's vector multiplied by $\epsilon\mu/c^2$.

Space cannot be taken here to deduce either of these principles concerning energy and momentum. It can easily be seen, however, that they can be interpreted in a concrete way that is of great interest in connection with modern ideas concerning mass and energy. We can, in fact, account for the transfer of energy by supposing that the energy in the field whose density is w is actually in motion with a certain velocity V in the direction of Poynting's vector Π . The necessary velocity is

$$V = \frac{\Pi}{w};$$

for, if we draw a right cylinder of height V and of unit cross section, with its length in the direction of Poynting's vector, this cylinder will contain wV or Π ergs of energy; and in a second this energy moves across the end of the cylinder, thus producing the rate of transfer required by Poynting's vector. If we then add the assumption that *each erg of the moving energy has a mass of $1/c^2$ or $1/(9 \times 10^{20})$ gram*, the energy in a cubic centimeter will have momentum equal to

$$(w/c^2)V = \Pi/c^2.$$

In free space, this is exactly the magnitude of the momentum in unit volume. Thus, the electromagnetic momentum in free space may be thought of as ordinary momentum possessed by the moving electromagnetic energy.

31. Electromagnetic Waves.—The field equations [(1a) to (1h)] admit of an enormous variety of different solutions. There is no general method for discovering useful solutions; in the end, one has to resort to guessing. Particular interest attaches to certain solutions that represent plane waves.

Suppose that

$$\mathbf{E}_y = f(x - vt), \quad \mathbf{E}_z = \mathbf{E}_x = 0, \quad (4)$$

¹ Cf. ABRAHAM, M., and R. BECKER, "Classical Electricity and Magnetism" Chap. XV; also LORENTZ, H. A., "Theory of Electrons" Secs. 19-25 (1916), a readable discussion.

$$H_z = \sqrt{\frac{\epsilon}{\mu}} f(x - vt), \quad H_x = H_y = 0, \quad v = \frac{c}{\sqrt{\epsilon\mu}}. \quad (5)$$

Here f stands for any differentiable function whatever. By substitution it is easily seen that all eight of the field equations are satisfied. Substitution in (16), for example, gives

$$-\sqrt{\frac{\epsilon}{\mu}} \frac{\partial f}{\partial x} = \frac{\epsilon}{c} \frac{\partial f}{\partial t}. \quad (6)$$

But

$$\frac{\partial f}{\partial x} = \frac{df}{d(x - vt)} \cdot \frac{\partial(x - vt)}{\partial x} = \frac{df}{d(x - vt)} \times 1, \quad (7)$$

because of the value assumed for v . Substituting these expressions for the derivatives in (6), we obtain an identity, thus showing that (6) is satisfied.

The physical characteristics of the field specified by this solution are simple and interesting. It represents waves of electric and magnetic field moving toward $+x$ with speed v . For, if a point is made to move toward $+x$ with a speed equal to v , at this point the quantity $x - vt$ always keeps the same value, since, in time At , x changes by $Ax = vAt$, and $x - vt$ changes, therefore, by

$$A(x - vt) = Ax - vAt = 0.$$

Hence, at this moving point, E_y and H_z always keep the same values. Furthermore, the waves are *plane*, since E_y has the same value at all points of any plane perpendicular to the direction of propagation, which is the z -axis; and so has H_z , also. Finally, the waves are *transverse*, for the vectors E and H are both perpendicular to the direction of propagation; and they are *plane-polarized*, with the electric vector in the xy plane.

As a special case, the function f might be sinusoidal in form. For example, with x' standing for any variable, we might take

$$f(x') = A \sin(2\pi x'/\lambda).$$

Then

$$f(x - vt) = A \sin 2\pi \left(\frac{x}{\lambda} - \frac{vt}{\lambda} \right),$$

and, if we also write ν for v/λ , Eqs. (4) and (5) become

$$E_y = A \sin 2\pi \left(\frac{x}{\lambda} - \nu t \right), \quad H_z = \sqrt{\frac{\epsilon}{\mu}} A \sin 2\pi \left(\frac{x}{\lambda} - \nu t \right). \quad (9a, b)$$

These equations represent sinusoidal waves of wave length λ and frequency ν , and of electric amplitude A .

Waves polarized in other planes, or traveling in other directions can be obtained by writing down equations like (4) and (5) in terms of rotated axes and then changing variables back to the old x, y, z . They all have the property that at any point, as in the example just given the electric field E and the magnetic field H are perpendicular both to each other and to the direction of propagation; the three vectors E , H , and direction of propagation form a right-handed set (cf. Sec. 29 and Fig. 3). The numerical ratio of H to E at any point is $\sqrt{\epsilon/\mu}$ when E is measured in electrostatic units and H in oersteds. Thus, in free space, $H = E$.

It should be remarked, however, that in other types of mediums, such as crystals or conductors, the relationships of E and H are somewhat different from those just described.

It is well known that electromagnetic waves of the sort we have been describing can be produced in the laboratory; and it was supposed until recently that nothing had to be changed in the description except the wave length in order to have an accurate description of ordinary light. The discoveries of the present century, however, as we shall see in later chapters, have made necessary a considerable modification in our conception of the electromagnetic field, so that the picture just described, while sufficiently correct for long electromagnetic waves, is not immediately applicable to ordinary light.

32. Energy of Electromagnetic Waves.—The most important property of electromagnetic waves is that they convey energy. In a plane wave, as we have seen, $H = \sqrt{\epsilon/\mu} E$, hence the electric and the magnetic parts of the energy density as written in (2) are equal, and we can also write, for a plane wave,

$$w = \frac{\epsilon E^2}{4\pi}. \quad (10)$$

This energy is carried along with the waves; and what is actually observable is usually not the energy density itself but the amount of energy that flows per second across unit area of a plane drawn perpendicular to the direction of propagation of the waves. This latter quantity is called the *intensity* of the train of waves; we shall denote it by I . The connection between I and w is easily found. Consider an elementary volume with faces having unit area and a thickness dx , placed perpendicular to the direction of propagation of the waves traveling as usual

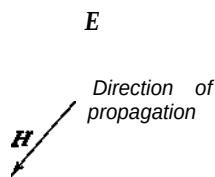


FIG- 3.—Relative directions

$$dt = \frac{dx}{v}$$

$$\epsilon = \mu = 1 \quad v = c,$$

$$I = cw = \frac{c}{4\pi} E^2.$$

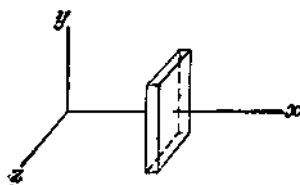


FIG. 4.

$$\theta = 90^\circ \quad \sqrt{\epsilon/\mu} E$$

$$\Pi = \frac{ev}{4\pi} E^2 = vw,$$

$$I = \frac{c}{8\pi} A^2.$$

$$\epsilon\mu/c^2$$

$$g = \frac{\epsilon\mu}{c^2} vw = \frac{\epsilon\mu I}{c^2} = \frac{w}{v}, \quad (15)$$

where w stands for the energy density in the waves and v for their velocity of propagation. Thus in free space

$$g = \frac{w}{c}, \quad (16)$$

and the total amount of momentum that is being carried by plane waves in association with total energy W is

$$G = \frac{W}{c}. \quad (17)$$

Experimentally, however, the electromagnetic momentum in the waves will manifest itself only in reactions upon material bodies when the momentum in the field is changed. Thus, if a beam of light is deflected, or perhaps absorbed, by a material body, a force will be observed to act on the body, just as a baseball player, when he destroys the momentum of a moving baseball by stopping it with his hand, feels a blow on the hand. The phenomena so produced are those of *radiation pressure*.

The magnitude of the pressure is easily calculated in any given case from the momentum brought up by the radiation. Since force equals rate of change of momentum, the pressure will equal the momentum delivered by the waves to unit area in a second. Since the waves move a distance v in a second, the momentum brought up will be that in a cylinder of unit cross section and length v , or vg c.g.s. units. If, then, the waves are entirely absorbed by the body, the radiation pressure on it will be

$$p = vg = w = \frac{nI}{c} \text{ dynes per cm.}^2, \quad (18)$$

by Eqs. (15) and (11), n denoting the refractive index of the medium or c/v .

That radiation falling on a body should exert a pressure upon it is by no means a concept peculiar to modern physics. Over 300 years ago, Kepler suggested that the curvature of comets' tails away from the sun might be due to radiation pressure—and modern astronomers believe that Kepler was right! Newton recognized that Kepler's suggestion was in harmony with the corpuscular theory of light (just as the pressure produced by a gas is due to the impact of its molecules). Experimentally, the subject was attacked as early as 1750, and at

intervals thereafter, but with inconclusive or conflicting results, due for the most part to the disturbing action of the gas surrounding the illuminated surface. Finally, however, the disturbances due to the gas were definitely eliminated, and radiation pressure was discovered almost simultaneously by Lebedev in Europe (1900) and by E. F. Nichols and G. F. Hull in America (1903).¹ The pressure, in the experiments of Nichols and Hull, was only about 7×10^{-5} dyne per square centimeter (which is several times the pressure due to strong sunlight), but they were able to measure this small pressure with sufficient precision to show that the observed values agreed within 1 percent with the theoretical values.

Radiation pressure is of particular interest in connection with the explanation of such atomic phenomena as the Compton effect.

34. Radiation Field of an Accelerated Point Charge.—We have seen how to obtain a solution of Maxwell's equations representing plane electromagnetic waves. It is natural to consider next how such waves can be generated or stopped. This problem was not solved by Maxwell; a complete treatment of it requires elaborate mathematical procedures. The emission of radiation by a point charge is of some interest in connection with the production of X-rays. Accordingly, the characteristics of such emission will be described.

(a) *A Uniformly Moving Point Charge.*—As a preliminar, the field around a *uniformly* moving charge will be described first. It can be shown from Maxwell's equations that, when any rigid set of charges is in uniform translational motion through empty space, the electric field is accompanied by a magnetic field whose intensity H at any point P has the magnitude

$$H = \frac{v}{c} E_{\perp}. \quad (19)$$

Here $E_{\perp}$ is the component of the electric intensity at P perpendicular to the direction of motion of the charges, v is their velocity, and c is the speed of light in vacuum. The direction of the magnetic field is perpendicular both to the direction of motion and to the electric field, and is such that the three vectors representing the direction of motion, the vector direction of $E_{\perp}$, and the vector direction of H form a right-handed orthogonal set (see Sec. 29). The magnetic field is sometimes imagined to be generated by transverse motion of the electric lines.

Consider now a point charge of q electrostatic units that is moving **at uniform** speed v along a straight line in vacuum. It can be shown **that the lines of** the electric field due to such a charge are straight

¹ Cf. NICHOLS and HULL, *Phys. Rev.*, vol. 17, pp. 29, 91 (1903).

and radial to the charge. Hence, if θ is the angle between a line drawn from the charge to any point P and the direction of motion of the charge, then at P we have $E_{\perp} = E \sin \theta$ and

$$H = \frac{v}{c} E \sin \theta, \quad (20)$$

where E is the intensity of the electric field at P . The lines of the magnetic field in this case are circles whose axis is the line of motion of the charge.

If the motion is slow enough so that v/c is small, then E has very nearly the same magnitude as for a stationary charge, so that $E = q/r^2$ where r is the distance of P from the charge. Then

$$H = \frac{qv}{cr^2} \sin \theta.$$

From this formula we see that the magnetic field around a slowly moving point charge is the same as that which the familiar Biot formula ascribes to a current element of magnitude $i ds = qv/c$ located at the position of the charge. If q is expressed in electromagnetic units, the equivalent current element is $i ds = qv$.

When v/c is not small, the formula for the intensity of the electric field is

$$E = \frac{\gamma}{[1 + (\gamma^2 - 1) \cos^2 \theta]^{\frac{3}{2}}} \frac{q}{r} \quad \gamma = \frac{1}{\sqrt{1 - \frac{v^2}{c^2}}}.$$

Equation (19) still holds, however, for the magnetic field. It is readily seen that, as v becomes nearly equal to c , both E and H become small except in directions lying close to the "equatorial" plane drawn through the charge and perpendicular to its motion; near this plane E and H have large values.

(b) *An Accelerated Point Charge.*—We may now consider the effects of an *acceleration* or change in the motion of the charge. The effect upon the field can be discovered easily, provided the following general principle is assumed. It can be shown from the field equations that the electromagnetic field due to any element of charge may be supposed to be emitted continually from that element and to be propagated out in all directions away from it at the speed of light.¹

This being assumed, the effect of acceleration can be found by elementary methods, and it will be instructive to do this in a simple case.

¹ Cf. ABRAHAM, M., and R. BECKER, "Electricity and Magnetism," p. 220 ("Electrodynamic Potentials"); SMYTHE, W. R., "Static and Dynamic Electricity," Sec. 13.25 ("Retarded Potentials").

Consider the special case of a point charge q , which for some time has been at rest at a point O in free space (Fig. 5) and now undergoes a uniform acceleration a during a very short time T , after which it continues to move uniformly with velocity

$$v = ar.$$

At the end of the period of acceleration, it will have moved to a slightly different point P_1 ; and, after an additional time t , much greater than T , it will have got to P_2 where

$$P_1P_2 = vt.$$

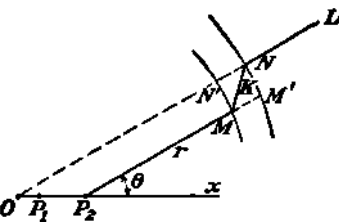


FIG. 5-Diagram illustrating the emission of a pulse of radiation.

Let us find the field at the end of the additional time t , when the charge is at P_2 .

Let ONL represent any line of force drawn from the charge while stationary at O ; let a point N lie on a sphere of radius $c(T + t)$ about O as center, c being the speed of light. Then the field emitted at the beginning of the period T has just had time to reach the surface of this sphere. The field *outside* must have been emitted earlier, and, therefore, according to the general principle just stated, it must be the original field associated with the stationary charge at O , not having received "news" of the acceleration of the charge. Similarly, *inside* a sphere of radius ct with center at P_1 , we have the field of the charge moving uniformly with speed v , since this field has had a time t during which to spread out. Between these two spheres is a shell within which the transition from the one type of field to the other occurs, the line of force running obliquely here from M to N . If we allow time to flow on, this transition shell moves outward with the speed of light. We shall show that it contains a radiation component of the field that carries a certain amount of energy off to infinity.

The part MN of the force line is straight, at least in the limit as $t \rightarrow 0$, because the acceleration is uniform.¹ Let us resolve the electric field at any point on MN into a radial component E_r , in the direction of P_2M , and a perpendicular component $E_\perp$. Then, at any point on

¹ That MN must be straight (in the limit) is easily seen, after reading the whole deduction, from the following consideration. The field at any point K on MN was emitted at a certain time t' after the charge left O . Had the acceleration ceased at this instant, the point M , as denoted in the text, would have been at K and the distance of M or K from ON , according to the expression found in the text for MN' , would have been $ar't \sin \theta$. Thus the distance of K from ON is proportional to t' . But so is KN , to the first order. Hence the line MN is straight.

MN ,

$$\frac{E_t}{E_r} = \frac{M'N}{MM'} = \frac{MN'}{MM'}.$$

But P_2M is (in the limit as $r \rightarrow 0$) parallel to ON' , since, according to Eq. (21) above, the change in the electric field due to the motion is of the second order in v , and since P_2M and ON' are different positions of the same force line (a line can be identified by its lying on a cone about Ox which encloses a given fraction of the electric flux of induction). Hence, if we neglect the distance OPI , which is of the second order in v ,

$$MN' = (OP_2) \sin \theta = P_1P_2 \sin \theta = vt \sin \theta = art \sin \theta.$$

Furthermore, MM' equals the difference in radii of the spheres if OPI is again neglected, whence,

$$MM' = cr.$$

Therefore, with the insertion of a minus sign because E_t has a component opposite to the acceleration a ,

$$\frac{E_t}{E_r} = -\frac{at}{c} \sin \theta.$$

But E_r must have the same value in the shell as outside in order to carry the flux of induction through the shell and so satisfy Gauss' theorem. Hence,

$$E_r = \frac{q}{r^2},$$

r being the radius of the shell. Thus,

$$E_t = -\frac{q}{r^2} \frac{at}{c} \sin \theta,$$

or, since $r = ct$, so that $t = r/c$,

$$E_t = -\frac{qa}{c^2 r} \sin \theta. \quad (22)$$

Thus, within the transition shell there is a tangential electric component E_t , which falls off, as r increases, only as the *first power* of r . This component lies in a plane drawn through the charge and containing the direction of its acceleration. As time goes on and $r \rightarrow \infty$, E_r will become entirely negligible in comparison with E_t . Thus, ultimately, we have a shell or pulse of field, traveling outward with the speed of light, in which the electric vector is *transverse*, just as in plane waves. In fact, any portion of the pulse ultimately approxi-

mates a plane wave. Nothing has been said about the magnetic field, which is not so easily treated; but it can scarcely be doubted that there must also exist in the pulse a transverse magnetic field, equal in strength to the electric field but perpendicular to it, since such a field exists in plane waves. The electromagnetic field that thus travels away to infinity from an accelerated charge is often called its *radiation field*.

By means of a relativity transformation, these results can readily be generalized so as to represent the radiation from a charge that is already in motion. It turns out that the radiation field is always numerically proportional to the *acceleration* of the charge, and in this field the electric intensity E , measured in electrostatic units, and the magnetic intensity H , measured in oersteds, are equal; furthermore, E , H , and the radius from the point of emission form a right-handed set of mutually perpendicular vectors (cf. Sec. 29). The complete formula for E is somewhat complicated, so we shall write it down only for the simple case in which the acceleration and the velocity lie along the same line, as in *linear motion*; in this case¹

$$E = - \frac{1}{(1 - \beta \cos \theta)^3} \frac{qa}{c^2 r} \sin \theta, \quad \beta = \frac{v}{c}, \quad (23a,b)$$

v being the instantaneous speed and a the instantaneous acceleration of the charge. The only effect of the velocity of the charge is represented by the first factor on the right in (23a), which serves to strengthen the radiation field in forward directions and to weaken it toward the rear of the moving charge.

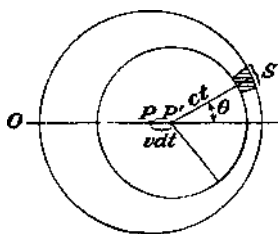


FIG. 6.—Diagram illustrating a pulse of radiation

35. Energy Radiated by an Accelerated Point Charge.—The most important feature introduced by the acceleration of a charge is the resulting radiation of energy.

We can readily calculate its magnitude for a charge moving along a line, from the formula just given. The results are of interest in connection with the theory of X-rays.

Let a charge, of magnitude q in electrostatic units, move with acceleration a , constant or variable, along a line Ox (Fig. 6). During an interval of time dt , let the charge move at speed v from P to P' , so that $PP' = v dt$. Then the radiation field emitted during this interval will be found, at an instant t seconds after the time at which the charge passed P' , between two spheres, one of radius ct about P'

¹ Cf. SMYTHE, *op. cit.*, Sec. 14.12.

as center, the other of radius $c(t + dt)$ about P as center. Let the time t be large enough so that in this region there is only the radiation field, in which $H = E$. Then there is in this field an energy density of magnitude (vacuum being assumed)

$$w = \frac{E^2 + H^2}{8\pi} = \frac{E^2}{4\pi}. \quad (24)$$

The distance between the two spheres, which would be $c dt$ if they had a common center, is changed by their relative displacement to the extent of the projection of PP' on the radius; thus, in a direction making an angle θ with the motion, the distance between the spheres is $c dt - v dt \cos \theta$ or $(c - v \cos \theta) dt$. Hence, between each pair of opposing unit areas on the spheres, there lies a volume equal to

$$(c - v \cos \theta) dt$$

times unity; and this volume contains energy of amount

$$dW_1 = w (c - v \cos \theta) dt \text{ ergs.}$$

Inserting the value of w from (24) and of E from (23a), we find

$$dW_1 = \frac{1}{4\pi} (c - v \cos \theta) \frac{q^2 a^2}{c^3 r^2} \frac{\sin^2 \theta}{(1 - \beta \cos \theta)^3} dt$$

or, dividing out c and using $\beta = v/c$ [Eq. (236)],

$$dW_1 = \frac{1}{4\pi} \frac{q^2 a^2}{c^3 r^2} \frac{\sin^2 \theta}{(1 - \beta \cos \theta)^3} dt. \quad (25)$$

If the velocity v is much below that of light, so that β is negligible, we can write this in the form

$$dW_1 = \frac{1}{4\pi} \frac{q^2 a^2}{c^3 r^2} \sin^2 \theta dt. \quad (26)$$

This energy dW_1 , lying between two unit areas on the spheres, eventually passes outward across a fixed unit area held so as to coincide momentarily with that on the outer sphere. The formula for dW_1 is applied to particular problems in different ways.

(a) *Instantaneous Rate of Radiation of Energy.*—The total rate at which energy is passing outward or is being radiated can be found by integrating over the outer sphere, whose radius is sensibly equal to r or to ct . We shall do this for the case of slow motion only. As an element of area on the sphere, take a ring of width $r d\theta$ and radius $r \sin \theta$, and hence of area $2\pi r^2 \sin \theta d\theta$, using as the axis of this ring the line from which θ is measured. The value of dW_1 is uniform over the ring, hence the energy that passes outward across it in time dt is

$$2\pi r^2 \sin \theta dW_1 d\theta = \frac{q^2 a^2}{2c^3} \sin^3 \theta d\theta dt$$

$$\frac{q^2 a^2}{2c^3} dt \int_0^\pi \sin^3 \theta d\theta.$$

$$\int_0^\pi \sin^3 \theta d\theta = \int_0^\pi (1 - \cos^2 \theta) \sin \theta d\theta = \left[-\cos \theta + \frac{1}{3} \cos^3 \theta \right]_0^\pi = \frac{4}{3}. \quad (27)$$

Hence, dividing by dt , we have for the energy passing outward over the sphere per second

$$\Omega = \frac{2}{3} \frac{q^2 a^2}{c^3}. \quad (28)$$

This may be regarded as the total rate at which the charge q is radiating energy at an instant when its acceleration has the value a . It is clear from the formula that, *according to classical theory, an accelerated charge must necessarily radiate energy.*

(b) *Stoppage of a Particle.*—Suppose a particle carrying a charge q and moving initially with speed v_0 is brought to rest without change in its direction of motion. In this case it is of interest to consider first the total radiation emitted by the charge during the entire stopping process, in a particular direction making an angle θ with the direction of motion. At a large distance r , the total energy W_1 crossing unit area of a plane perpendicular to the radius from the particle can be got by integrating dW_1 , as given in (25) or (26), with respect to the time. The result of the integration will depend upon the manner in which the acceleration a varies with the time; for simplicity, let it be uniform. Then, for the process in question

$$\int dt = \int_{v_0}^0 \frac{dv}{a} = -\frac{v_0}{a}.$$

Hence, if v_0 is small compared with the speed of light c , we have from (26)

$$W_1 = \int dW_1 = \frac{1}{4\pi} \frac{q^2 v_0 |a|}{c^3 r^2} \sin^2 \theta, \quad (29)$$

$|a|$ being the absolute value of the acceleration and equal to $-a$. If, on the other hand, v_0/c is not small, we have only to note that $dt = dv/a = c dp/a$, so that, writing $\beta_0 = v_0/c$, we have during the

$$\int \frac{dt}{(1 - \beta \cos \theta)^3} = \frac{c}{a} \int_{\beta_0}^0 \frac{d\beta}{(1 - \beta \cos \theta)^3} = \frac{c}{4a \cos \theta} \left[1 - \frac{1}{(1 - \beta_0 \cos \theta)^4} \right]$$

By noting that

$$\begin{aligned} 1 - (1 - \beta_0 \cos \theta)^4 &= [1 - (1 - \beta_0 \cos \theta)^2][1 + (1 - \beta_0 \cos \theta)^2] \\ &= (2\beta_0 \cos \theta - \beta_0^2 \cos^2 \theta)(2 - 2\beta_0 \cos \theta + \beta_0^2 \cos^2 \theta), \end{aligned}$$

the last expression can be got into a more useful form. The result of integrating (25) with respect to the time can then be written

$$W_1 = \frac{1}{4\pi} \frac{q^2 v_0 |a|}{c^3 r^2} \sin^2 \theta \frac{(1 - \frac{1}{2} \beta_0 \cos \theta)(1 - \beta_0 \cos \theta + \frac{1}{2} \beta_0^2 \cos^2 \theta)}{(1 - \beta_0 \cos \theta)^4}. \quad (30)$$

The *total* radiation emitted during the stopping process can then be found by integrating W_1 in turn over a sphere of radius r . We shall do this, however, for the case of slow motion only. As an element of area on the sphere, take the same ring as before. The total energy emitted is thus found to be, from (29),

$$W = 2\pi \int_0^\pi W_1 r^2 \sin \theta d\theta = \frac{q^2 v_0 |a|}{2c^3} \int_0^\pi \sin^3 \theta d\theta,$$

or, by (27),

$$W = \frac{2}{3} \frac{q^2 v_0 |a|}{c^3}. \quad (31)$$

We note that, the more rapidly the charge is brought to rest the larger is the amount of energy radiated.

(c) *Damped Harmonic Oscillator*.—A case of interest in comparison with the emission of radiation from atoms is that of a charged body bound elastically to an equilibrium position. If the body is set into vibration, it will gradually lose its energy by radiation because of the acceleration experienced at various parts of its path and will eventually come to rest. It should be remarked again, however, that correct results for atomic phenomena can be obtained only from quantum theory; results obtained from classical theory constitute at best interesting analogies.

Suppose for simplicity that the motion is one-dimensional. Then, if the damping by radiation is small, the displacement x of the body from its position of equilibrium will be given approximately by the usual equation of simple harmonic motion,

and its energy of vibration will be

$$W = \frac{1}{2} m (v_{\max})^2 = 2\pi^2 m \nu^2 A^2, \quad (32)$$

$V_{\max}$ or $2\pi\nu A$ being its maximum velocity and m its mass. The instantaneous rate of loss of energy is given by Eq. (28). In the present case the acceleration is

$$a = \frac{d^2x}{dt^2} = -4\pi^2\nu^2 A \sin 2\pi\nu(t + \delta).$$

The energy lost during a single period of vibration, lasting $1/\nu$ seconds, will be, therefore,

$$\begin{aligned} \int_{t_1}^{t_1+1/\nu} \frac{2q^2 a^2}{3c^3} dt &= \frac{32\pi^4\nu^4 q^2 A^2}{3c^3} \int_{t_1}^{t_1+1/\nu} \sin^2 2\pi\nu(t + \delta) dt \\ &= \frac{16\pi^4\nu^4 q^2 A^2}{3c^3}, \end{aligned} \quad (33)$$

since

$$\int_{t_1}^{t_1+1/\nu} \sin^2 2\pi\nu(t + \delta) dt = \int_{t_1}^{t_1+1/\nu} \left[\frac{1}{2} - \frac{1}{2} \cos 4\pi\nu(t + \delta) \right] dt = \frac{1}{2\nu}.$$

The average energy lost per second will then be the expression obtained in (33) multiplied by ν or $\frac{16\pi^4\nu^4 q^2 A^2}{3c^3}$. Hence, we can write for the energy of the oscillator, provided the loss in a single period is small, the approximate differential equation

$$\frac{dW}{dt} = -\frac{16\pi^4\nu^4 q^2 A^2}{3c^3} = -\frac{8\pi^2\nu^2 q^2}{3mc^3} W, \quad (34)$$

by Eq. (32). The solution of this equation is, if W_0 is the value of W at $t = 0$,

$$W = W_0 e^{-\alpha t}, \quad \alpha = \frac{8\pi^2\nu^2 q^2}{3mc^3}. \quad (35a,6)$$

The energy and amplitude of the vibrating charge thus decrease exponentially at a rate proportional to ν^2 .

36. Electromagnetic Mass.—It was pointed out by J. J. Thomson in 1881 that a body should appear to have a larger mass when charged than it has when uncharged. The effect is essentially one of self-induction; just as the process of starting a current in a wire calls into existence a momentary electromotive force opposing the current, so, similarly, setting a charge into motion produces, associated with the changing magnetic field around it, an electric field that acts on the charge so as to oppose its acceleration.

Nowadays, however, we prefer to view this phenomenon in terms of the electromagnetic momentum in the field. This momentum is easily calculated in the case of a uniformly charged sphere moving slowly (as compared with the speed of light). The electromagnetic field around such a sphere is the same as that of a point charge located at its center. At a point distant r from the center of the sphere, on a radius making an angle θ with the direction of motion, we have a radial electric field strength

$$E = \frac{q}{r^2},$$

q being the total charge on the surface of the sphere in electrostatic units, and also a magnetic field whose magnitude, according to Eq. (20) in Sec. 34, is

$$H = \frac{qv}{cr^2} \sin \theta.$$

Thus Poynting's vector, described in Sec. 30, will lie in a plane through the line of motion and will be perpendicular to the radius, as shown by the arrow marked P in Fig. 7; its magnitude, since E and H are perpendicular, will

$$\Pi = \frac{c}{4\pi} EH = \frac{q^2 v}{4\pi r^4} \sin \theta.$$

According to the rule stated in Sec. 30, therefore, there will be electromagnetic momentum in the region surrounding the sphere to the extent of Π/c^2 per unit volume, its direction being that of Poynting's vector. Inside the sphere there is no field and hence no momentum.

Because of the symmetry, however, the component of this momentum perpendicular to the motion will sum up to zero. Thus the resultant momentum in the whole field is parallel to the motion, and to find it we need sum up components in this direction only. The component of momentum parallel to the motion is, per unit volume,

$$\frac{\Pi}{c^2} \sin \theta = \frac{q^2 v}{4\pi c^2 r^4} \sin^2 \theta.$$

As an element of volume we may take a circular filament whose axis is the line of motion of the charge, with a cross section $r d\theta dr$ and a perimeter $2\pi r \sin \theta$ ($r \sin \theta$ being its radius), so that its volume is $2\pi r^2 \sin \theta d\theta dr$. The component of the momentum has the same value

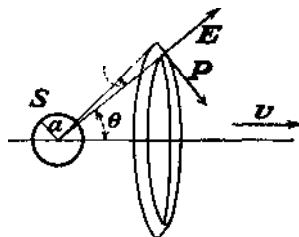


FIG. 7.—Diagram illustrating the calculation of the electromagnetic momentum around a moving charged sphere S .

at all points on this ring. Then the total electromagnetic momentum in the field will be

$$\begin{aligned} M &= \int \int \frac{q^2 v}{4\pi c^2 r^4} \sin^2 \theta \cdot 2\pi r^2 \sin \theta d\theta dr \\ &= \frac{q^2 v}{2c^2} \int_{-a}^a \frac{dr}{r^2} \int_0^\pi \sin^3 \theta d\theta, \end{aligned}$$

a being the radius of the sphere; or, from (27), and since $\int_a^\infty \frac{dr}{r^2} = \frac{1}{a}$,

$$M = \frac{2}{3} \frac{q^2 v}{ac^2}.$$

This is the same amount of momentum that the sphere would possess, according to the ordinary formula for momentum, if it had a mass m_q of magnitude

$$m_q = \frac{2}{3} \frac{q^2}{ac^2}; \quad (36)$$

for then $M = m_q v$. The sphere will behave dynamically as if it possessed a mass $m = m_0 + m_q$, where m_0 is its ordinary mass.

The formula just obtained for m_q is correct, however, only for indefinitely slow motion. In general, a more complicated formula is obtained. In case the sphere becomes contracted in the direction of motion in the ratio $(1 - v^2/c^2)^{1/2}$ in accordance with the requirement of the theory of relativity (*cf.* Sec. 61), it can be shown that the electromagnetic mass is

$$m_q = \frac{2}{3} \frac{q^2}{ac^2} \frac{1}{\sqrt{1 - v^2/c^2}}. \quad (37)$$

THE PHOTOELECTRIC AND THERMIONIC EFFECTS

An approach to the outstanding problem of the first quarter of the present century, viz., the nature of radiant energy, is most directly made by a study of the photoelectric effect. The term "photoelectric" might be applied with reason to a wide variety of phenomena involving the interaction between light and electricity, such as the change in the resistance of selenium under illumination or the rotation of the plane of polarization of a beam of light when passing through a medium placed in a strong electric field. This term is commonly restricted, however, to the discharge of negative electricity from bodies when illuminated by light of appropriate wave length. In this chapter, we shall give a brief discussion of this effect, emphasizing those features which proved especially difficult of explanation in terms of the wave theory of light. The discussion of the photoelectric effect naturally involves a consideration of the experiments that led to one of the most striking advances of the modern period, the discovery of the electron. Some of these experiments will be described.

Incidentally, we shall include a brief discussion of a closely related phenomenon, that of thermionic

37. Discovery of the Photoelectric Effect

The photoelectric effect was discovered by Heinrich Hertz,¹ quite accidentally, in 1887.

While carrying on experiments in which he demonstrated the existence of electromagnetic

waves, he observed that the spark between the terminals (*S*, Fig. 8) of his detecting circuit passed more rapidly when those terminals were illuminated by the light from the primary spark *P* than when an opaque screen, or even a piece of plate glass, was interposed between *P* and *S*. Because the primary spark was known to emit ultraviolet light, which would be filtered out by the glass, Hertz concluded that the effect was due to ultraviolet light falling on the secondary gap—a conclusion which he confirmed by screening *S* from *P* and allowing light from another source to fall on *S*. Further,

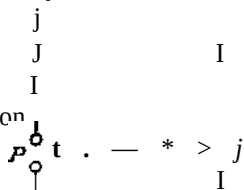


FIG. 8.

¹*Ann. d. Physik*, vol. 31, p. 983 (1887).

he found that the light must fall on the terminals themselves, which, for the effect to be observed, must be "smooth and clean." Hertz announced this discovery in a paper entitled "An Effect of Ultraviolet Light upon the Electric Discharge," and then returned to his investigations of electromagnetic waves.

Hertz's discovery at once attracted numerous investigators. Hallwachs¹ found that a freshly polished zinc plate, insulated and connected to an electroscope as an indicator, when charged *negatively* and illuminated by ultraviolet light, would lose its charge, but that there was no effect if the charge was *positive*. He concluded that in some unknown manner, when the plate was negatively charged and illuminated, negatively electrified particles were emitted from it. He even observed, by using an electrometer instead of an electroscope, that a *neutral* insulated plate, when illuminated, would acquire a small *positive* potential, i.e., would lose a *negative* charge. Not only did

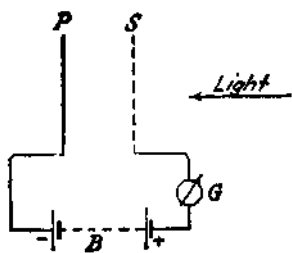


FIG. 9.

light permit the escape of negative electricity from a negatively electrified plate, it even caused the expulsion of negative electricity from a neutral plate.

Stoletow² devised an arrangement, shown diagrammatically in Fig. 9, for producing a continuous *photoelectric current*. *P* is a photoelectrically sensitive plate, say a polished zinc plate, connected to the negative terminal of a battery *B* of several cells. *S* is a wire grating or gauze connected to the positive terminal of the battery through a very sensitive galvanometer or electrometer *G*. When ultraviolet light falls upon *P*, a continuous current is observed in *G*, indicating that a negative charge is flowing from *P* to *S*. No current flows if the battery is reversed.

Elster and Geitel³ showed that there is a close relation between the contact potential series of metals and the photoelectric effect: the more electropositive the metal, the longer the wave length to which it would respond photoelectrically. The alkali metals, sodium, potassium, and rubidium, were found to be sensitive even to light of the visible spectrum.

38. A Problem.—The question then arose as to the mechanism by which negative electricity is transferred from the cathode to the anode. That the charge is carried by negatively electrified particles

¹ *Ann. d. Physik*, vol. 33, p. 301 (January, 1888).

² *J. de Physique*, vol. 9, p. 486 (1890).

³ *Ann. d. Physik*, vol. 38, pp. 40, 497 (1889).

was clearly indicated by experiments of Elster **and** Geitel, who showed that a transverse magnetic field diminishes the photoelectric current if the phenomenon takes place in a vacuum. But what are the particles?

Negative answers to this question were readily obtained. It was early found that the effect persisted even to the highest attainable vacuum and was quite independent of the "degree" of the vacuum after a certain low pressure had been reached. This seemed to indicate that the gas molecules themselves in the region between *P* and *S* could not be acting as carriers of the charge. The suggestion was made that, perhaps, under the influence of light, negatively charged particles of the cathode became detached and moved to the anode. This suggestion was rendered untenable by an experiment by P. Lenard,¹ in which a clean platinum wire acted as anode and a sodium amalgam as cathode, both being in an atmosphere of hydrogen. The photoelectric current was allowed to flow until about 3×10^{-6} coulomb had passed through the circuit. If the carriers of the charge were atoms of sodium, each atom could hardly be expected to carry a larger charge than it carries in electrolysis. Taking the electrochemical equivalent of sodium as 0.00024 gram per coulomb, there should have been deposited on the platinum wire at least 0.7×10^{-6} milligram of sodium, a quantity sufficient to be detectable by the well-known flame test. On removing the wire from the bulb, however, no trace of sodium could be detected.

If, then, the photoelectric current is carried neither by molecules of the gas surrounding the cathode nor by molecules of the cathode itself, what are the carriers? The answer to this question came through the convergence of a number of different lines of evidence, which finally culminated in the discovery of the electron by J. J. Thomson. Before continuing the discussion of the photoelectric effect we shall consider some of these other developments.²

¹ *Ann. d. Physik*, vol. 2, p. 359 (1900).

² Collectively these developments illustrate many characteristics of the growth of modern physics:

1. A number of seemingly unrelated lines of research frequently converge to provide an explanation of, or a theory for, a group of phenomena not hitherto understood.

2. The explanation or theory thus evolved is then found to bear directly on other branches of physics and often on other sciences.

3. Thus, the methods of physics are both (1) synthetic and (2) analytic—a fortunate circumstance, which makes it possible for the physicist to comprehend physics as a whole in spite of the vast increase, particularly in recent years, of factual knowledge.

4. These discoveries, of both fact and theory, are the *sine qua non* of applied physics and of much of industry—witness, as a single example, the wide use of the

39. Electricity in Matter.—A certain atomicity of electricity was suggested long ago by the laws of electrolysis. Faraday found that when the same quantity of electricity deposits univalent ions of different kinds, the amounts deposited, as measured by weight, are proportional to the *ionic weights* in the chemical sense. By the "ionic weight" is meant the ordinary atomic weight (oxygen = 16) if the ion consists of a single atom, or the sum of the atomic weights of the atoms composing a compound ion. Suppose an amount F of electricity deposits a gram-ion of a certain kind of univalent ion, *i.e.*, a quantity the mass of which in grams is equal to the ionic weight. If the ion is monatomic, the mass deposited will be a gram-atom, or a number of grams equal to the atomic weight. Then the same quantity of electricity would suffice to deposit a gram on any other kind of univalent ion. This quantity of electricity : an important fundamental constant in chemistry and physics and is called the *faraday*. The best modern value of it is¹

$$F = 96,488 \text{ coulombs.}$$

Now the number of ions in a gram-ion, a , atoms in a gram-atom (or of molecules in a gram-molecule) is in all cases the same; for, if equal numbers be taken of different ions, the masses will be proportional to the ionic weights of J . This number, which we shall denote by N_0 , is an important physical constant. It follows that all univalent ions carry the same numerical charge, equal to F/N_0 . The charge on a univalent ion thus constitutes a natural unit of charge, usually denoted by e . Multivalent ions carry multiples of e , divalent ions carry charges of $2e$, and so on. For this natural unit of charge Dr. G. Johnstone Stoney proposed in 1891 the name "electron."

If the number N_0 were known, we could at once calculate e from the equation

$$e \sim N_0^{-1}$$

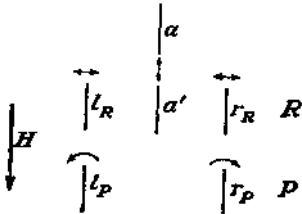
In Stoney's time, however, it was known only, from kinetic theory, various kinds of photo- and thermionic tubes.

5. This application of science to industry very frequently reacts to provide the research man with improved tools. Electron tubes, manufactured primarily for industrial purposes, are a boon to the physicist in his research laboratory.

6. In all these developments, scientists of all nations "collaborate," the more effectively because of the very fact that they are not formally organized, *let alone directed*, but are free as individuals to follow their respective interests.

An understanding of these generalizations is essential to an understanding of modern physics.

¹ BIRGE, *Rev. Modern Phys.* vol. 13, p. 233 (1941).



components IR (left, shorter wave length) and r_R (right, longer wave length) are plane-polarized with the electric vector perpendicular to the magnetic field.

If the pole pieces of the electromagnet be drilled through longitudinally so that one may view the source in a direction parallel to the magnetic field, only two components IP and r_P are seen, as shown at P . These two lines have the same respective wave lengths as the outside components in the previous case, but they are *circularly* polarized in opposite directions, as shown by the arrows above the lines. (In this latter case the magnetic field is directed toward the reader.)

In his original paper announcing his discovery, Zeeman discusses the phenomenon at first in terms of a mechanical ether, but finally he decides in favor of an explanation in terms of the electrical theory of matter, which had recently been developed in its complete form by Lorentz. It may be worth while to derive here the classical Zeeman-Lorentz theory of the effect, in spite of the fact that we now believe classical theory not to be applicable to atomic phenomena. A simple concrete picture is thus obtained which assists the memory, in contrast with the abstractness of the wave-mechanical theory. Furthermore, a motion such as Zeeman supposed to occur in the atom could actually be executed by an electron in a vacuum tube.

In order to have light of a fixed frequency emitted by electrical charges, according to classical theory, the charges must vibrate in simple harmonic motion. Let there be then, within an atom, a particle having a charge of e electromagnetic units and a mass of m grams (e being either positive or negative). When this particle is displaced a distance r from its normal position of equilibrium O , let it be acted upon by a restoring force of br dynes directed toward O . Then the *component* of the force parallel to any fixed line (e.g., AB in

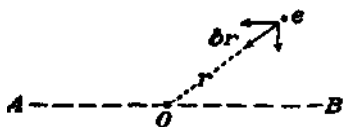


FIG. 11.—A particle vibrating harmonically in two dimensions.

Fig. 11) will also equal b times the component of the displacement *in this direction*. Hence, this component of the displacement will vary harmonically with a period

$$T = 2\pi \sqrt{\frac{m}{b}} \text{ seconds,} \quad (38)$$

according to the ordinary laws of simple harmonic motion. Any component perpendicular to AB will do the same but, perhaps, with different amplitude and phase. Thus, the general motion of the particle will be equivalent to three independent vibrations in mutually perpendicular directions, all executed with the same period T .

Now suppose the particle is vibrating in a single direction, say along the line AB in Fig. 11 or A_1B_1 in Fig. 12. Let a magnetic field of strength H oersteds be applied at right angles to the plane of the paper and directed away from the reader. According to the familiar elementary law, the charge will now experience an additional force f_H , at right angles both to the field and to its velocity, whose magnitude is

$$f_H = Hev \text{ dynes,}$$

e being the magnitude of the charge in electromagnetic units and v the velocity in centimeters per second. This force will deflect the charge sideways. If it is a positive charge and starts from A_1 toward

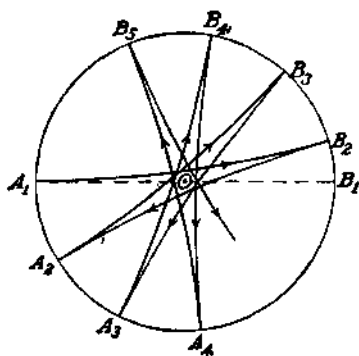


FIG. 12.—Effect of a magnetic field on a charged body vibrating initially with simple harmonic motion at right angles to the field.

B_1 (Fig. 12), it will arrive at some such point as B_2 ; starting back toward 0, it will be deflected again and arrive at A_2 , and so on. Thus, the field causes the path to rotate slowly, in a plane perpendicular to the field.

The effect of this rotation can be found most easily by first resolving the motion into two circular motions about 0. Suppose first that there is no magnetic field. Then the particle is capable of revolving around the circle shown in Fig. 12 with a steady speed v_0 or an angular velocity ω_0 and period of revolution T_0 given by

$$\frac{mv_0^2}{r} = m\omega_0^2 r = br, \quad \omega_0 = \sqrt{\frac{b}{m}} = \frac{2\pi}{T_0}. \quad (39a,6)$$

Suppose now we imagine a motion in which the particle revolves in this manner, and another in which the motion is similar but is hand and add the two resulting motions on the assumption that the particle starts from B_1 at the same time in both. Then components of displacement perpendicular to

$A1B1$ cancel out, and we have a linear harmonic vibration of period T along $A1B1$.

Now introduce a magnetic field, directed, as before, away from the reader in Fig. 12. In addition to the elastic force $6r$, there will then be the force Hev ; if the particle moves at constant speed around the circle in a counterclockwise direction and is positively charged, this added force will be constant and directed toward the center. A steady circular motion is, therefore, still possible, but the speed must be given not by (39a) but by

$$\frac{mv^2}{r} = br + Hev.$$

Putting $v = \omega r$, where ω is the new angular velocity, we have, after rearranging,

$$\omega_1^2 - H \frac{e}{m} \omega_1 = \omega_0^2,$$

for, by Eq. (396), $b/m = \omega_0^2$. Solving,¹

$$\omega_1 = \frac{He}{2m} + \sqrt{\omega_0^2 + \frac{H^2 e^2}{4m^2}}. \quad (40a)$$

If the motion is clockwise, instead of counterclockwise, the force due to the magnetic field acts in the opposite direction and we find¹

$$\omega_2 = -\frac{He}{2m} + \sqrt{\omega_0^2 + \frac{H^2 e^2}{4m^2}}. \quad (40b)$$

In all cases of interest, however, $He/2m$ is very small in comparison with ω_0 , so that the radicals in these equations may be replaced by ω_0 . Hence, if we change to frequencies by writing $\omega_0 = 2\pi\nu_0$, $\omega_1 = 2\pi\nu_1$, and $\omega_2 = 2\pi\nu_2$, we have, for a positive charge,

$$\nu_1 = \nu_0 + \frac{He}{4\pi m} \text{ (counterclockwise)}, \quad (41a)$$

$$\nu_2 = \nu_0 - \frac{He}{4\pi m} \text{ (clockwise)}. \quad (41b)$$

If two such circular motions with slightly different frequencies are superposed, it is easily seen that their resultant is a rotating linear vibration such as is illustrated in Fig. 12, and that the quantity $He/4\pi m$ gives the rate at which the direction of vibration is caused thus to precess about the direction of the magnetic field. This effect is known as the "Larmor precession." However the particle may be vibrating,

¹ The solution containing a negative radical is rejected because it makes ω_1 or ω_2 negative, implying rotation in a direction opposite to that assumed.

its component of motion *perpendicular* to the field can be resolved into two circular motions of the type described. Its component of motion *parallel* to the field, on the other hand, is unaffected and occurs at the undisturbed frequency ν_0 .

The characteristic features of the *emitted radiation* are easily inferred now if we recall that the electric intensity radiated by an accelerated charge is directed as nearly oppositely to the acceleration as it can be, subject to the condition that it must be perpendicular to the direction of propagation of the waves (see Sec. 34). If the emitted light is viewed with a spectroscope in a direction *perpendicular* to the magnetic field, as by an observer at P in Fig. 12, three lines will be seen. One, emitted by the component of vibration parallel to the field, will have a frequency ν_0 and will be polarized with the electric vector parallel to the applied magnetic field; the two others, emitted by the circular motions, will have frequencies ν_1 and ν_2 and will be polarized with the electric vector perpendicular to the magnetic field. If, on the other hand, the light is viewed in a direction *parallel* to the field, only the lines of frequency ν_1 and ν_2 can be seen [since in Eq. (22) $\sin \theta = 0$ for the third component of the vibration]; and they will be circularly polarized. These features agree exactly with Zeeman's observations.

For a *positive* particle, however, the slower rotation occurs in a clockwise direction as seen by an observer looking in the direction of the field [cf. Eq. (41b)], whereas Zeeman found counterclockwise rotation for the line of lower frequency (cf. Fig. 10, P, where the magnetic field is assumed to be directed *toward* the reader and the directions of rotation are therefore reversed). He concluded, therefore, that the charge on the radiating particles must be *negative*. The ratio e/m for the particles can be computed from the separation between the outer lines. Letting e stand hereafter for the numerical charge on a particle, we have from Eqs. (41a,b)

$$\Delta\nu = \nu_2 - \nu_1 = \frac{He}{2\pi m}, \quad \frac{e}{m} = \frac{2\pi}{H} \Delta\nu. \quad (42a,b)$$

Zeeman's first observation indicated that e/m must be "of the order of magnitude of 10^7 electromagnetic units." In a later experiment, working with much higher resolving power, he found $e/m = 1.6 \times 10^7$, as against the modern value of 1.76×10^7 .

Later experiments, however, have shown the Zeeman effect to be much more complicated than Zeeman at first assumed from his observations with apparatus of relatively low resolving power. In general, many more components are observed than are shown in Fig. 10. The

simple theory outlined above fails completely to explain these more complicated "patterns." The quantum theory of the Zeeman effect is discussed in Chap. VIII.

41. The Discovery of the Electron.—Previous to 1897, many studies had been made of that beautiful phenomenon, the discharge of electricity through rarefied gases. Let the discharge from an induction coil or an electrostatic machine pass between the negative terminal *C* (cathode) and the positive terminal *A* (anode) sealed into a glass tube (Fig. 13), which is being exhausted through the side tube *T*. At a very low pressure, there appears around the cathode a dark space, known as the "Crookes dark space," which, with further decrease in pressure, grows longer (*i.e.*, extends farther toward *D*), until finally it reaches the glass walls of the tube. The glass is then observed to glow, the color, greenish or bluish, depending on the kind of glass of which the tube is made. If screens pierced with holes *S1* and *S2* are intro-

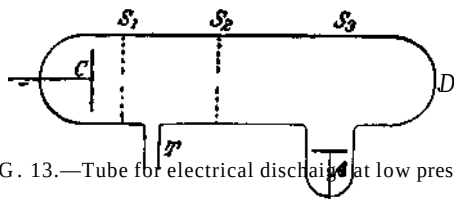


FIG. 13.—Tube for electrical discharge at low pressure.

duced, the glow is confined to a spot on the end of the tube, at *D*. The "something" which, under these conditions, seems to proceed from the cathode and to cause the phosphorescence of the glass was early called "cathode rays." In support of this view could be cited the usual facts listed in elementary texts of physics. The rays are deflected by a magnetic field; they are also deflected by an electrostatic field, the spot at *D* moving upward if a positively charged rod is brought up in the position *S3*. The concentration of the spot, in line with the holes in *S1* and *S2*, indicates that something is proceeding from *C* to 2); if the cathode is suitably curved, this something can be "focused" upon a piece of platinum foil within the tube, which is heated to incandescence if the rays are sufficiently intense. Finally, in 1895, Perrin caught the rays in an insulated chamber connected to an electroscope and proved that they carry a negative charge.

Final confirmation of the correctness of the view that cathode rays are moving negatively charged particles came from the classical experiments¹ of J. J. Thomson. His experiments are so fundamental in the

¹*Phil. Mag.*, vol. 44, p. 293 (October, 1897).

history of the electrical theory of matter as to warrant description. The highly evacuated glass tube [Fig. 14(a)] contains the cathode C and an anode A , which has a small rectangular slot in it through which the cathode rays may pass. B is a screen similar to, and electrically connected with, A . Cathode rays, accelerated from C toward A , after passing the slot in A move with uniform velocity and emerge from the slot in B as a small bundle of rectangular cross section, which causes a small, fluorescent patch at p_1 on the far side of the bulb. But when a potential difference V is maintained between the parallel

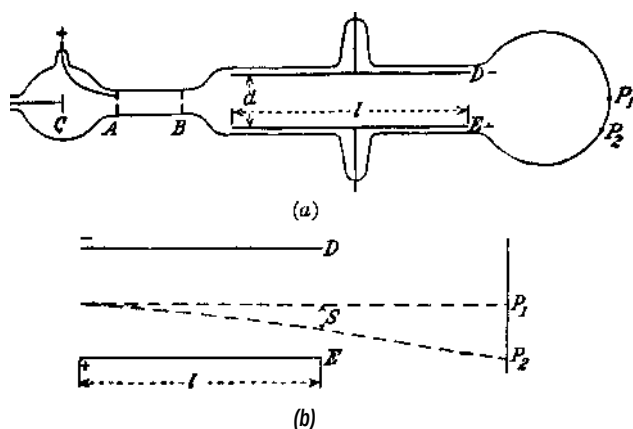


FIG. 14.- (a) Thomson's apparatus for determining the ratio e/m for electrons, (b) Deflection of an electron by the electrostatic field.

plates D and E , E being positive, the spot appears at P_2 , having been deflected downward by the electrostatic field. A pair of Helmholtz coils, not shown, whose diameters are equal to the length of the plates D and E , are placed, one in front and the other behind the tube, so as to produce a magnetic field perpendicular to the plane of the paper, the strength of which can be determined from the dimensions of the coils and the current through them. If the magnetic field is directed *toward* the reader, the spot P_1 is deflected *upward*.

Two experiments are now performed:

1. With a given electrostatic field between the plates, the strength of the magnetic field is adjusted to such a value as will cause the spot to return to the original undeviated position P_1 .
2. The magnetic field is then removed and the deflection P_1P_2 caused by the electrostatic field alone is measured.

From these two experiments, the ratio e/m may be determined as follows. In the first experiment, the electrostatic field has a magnitude $E = V/d$, if d denotes the distance between the plates, and it exerts a

force Ee upon the charge e . Let both E and e be measured in electromagnetic units. This force is just equal and opposite to the force Hev due to the magnetic field, H being the field in oersteds and v the velocity with which the particles emerge from B . The velocity is constant over the entire path from B to P_1 because there are no longitudinal forces acting on the particle. Hence $Hev = Ee$ and

$$v = \frac{E}{H}. \quad (43)$$

Thus the first experiment serves to measure the velocity of the particles as they emerge from B . In the second experiment, let the distance PIP_2 be measured; from this, the deflection S of the particles as they pass over the distance l between the plates D and E can be determined [cf. Fig. 14(6)]. This deflection results from a uniform acceleration

$$a = \frac{Ee}{m},$$

acting during a time l/v ; hence, by the laws of uniformly accelerated motion,

$$S = \frac{1}{2}E \frac{e}{m} \left(\frac{l}{v} \right)^2. \quad (44)$$

All quantities in this equation being known except e/m , the latter may be computed.

Thomson found that the value of e/m determined in this way was of the order of 10^7 and that it was independent of the kind of gas in the tube (air, H_2 , or CO_2) and, likewise, independent of the material of the electrodes (Al, Fe, or Pt). A later determination gave

$$\frac{e}{m} = 1.7 \times 10^7,$$

a value almost numerically identical with the value of e/m determined from the Zeeman effect for the particles taking part in light emission. (The modern value is 1.76×10^7 .) This value of e/m is very much larger than the value of e/m for hydrogen atoms in electrolysis. Its great magnitude might result either from a large value of e or a small value of m or both. It became a matter of much importance, therefore, to determine the charge e carried by these particles.

Now it seemed almost certain that the charge on the cathode particles must be of the same order as the charges on gaseous ions such as are produced in a gas by X-rays. The study of such ions had begun just a few years before Thomson's work. To measure their

charges, Townsend, working in Thomson's laboratory, utilized the clouds that form about the ions in saturated air. By observing the rate of fall of the cloud and applying Stokes' law for the free fall of spheres through a viscous medium, he was able to determine the size of the droplets; from a measurement of the total amount of water in the cloud, he could then calculate the number of droplets it had contained. He assumed that each droplet contained just 1 ion; hence, having measured also the total charge on the cloud, he was able to calculate the charge on a single ion. For this charge he obtained a value of about 3×10^{-10} electrostatic unit. A repetition of the measurements with some modifications by Thomson gave the value 6.5×10^{-10} electrostatic unit.¹

Thomson assumed that the charge on the gaseous ions was the same as that on his cathode particles. It followed, then, that the cathode particles must be previously unknown particles of extremely small mass, which he called "corpuscles" or "primordial atoms." For many years English writers stuck to the name "corpuscle" for these particles, using the word "electron" in Stoney's original sense to denote the amount of charge carried by a corpuscle or a univalent ion; but others, including Lorentz, called the corpuscles themselves "electrons," and this usage ultimately became well established. It was generally assumed that electrons are a constituent part of all atoms and are responsible for the emission of light by them, thus accounting for the fact that the ratio e/m had been found to be the same for the vibrating particles causing the Zeeman effect as for the cathode rays. Thus was made possible the explanation of a number of more or less diverse phenomena on the basis of a single concept.

42. Electronic Magnitudes.—A much more reliable method of measuring ionic charges was developed by Millikan in 1909. He found that tiny droplets of oil in ionized air, viewed with a microscope, would frequently pick up charges and could then be held suspended, or accelerated upward or downward, by applying a suitable electric field. When uncharged, the droplets fall at a slow uniform rate, their weight being balanced by the drag due to the viscosity of the surrounding air. By observing their rate of fall, Millikan was able to determine the size and weight of the droplets. According to Stokes' law, a sphere of radius a moving at a steady slow speed v through a fluid whose coefficient of viscosity is η experiences a resisting force

$$F = 6\pi\eta av_0. \quad (45)$$

¹ For a good account of work on the electronic charge see Millikan, "Electrons (+ and —), Protons, Photons, Neutrons, and Cosmic Rays," 1935.

If we equate this force to the weight of a droplet, which is $\frac{4}{3}\pi a^3 \rho g$ in terms of the density ρ of the oil, we have

$$\frac{4}{3}\pi a^3 \rho g = 6\pi \eta a v_0. \quad (46)$$

Now suppose the drop picks up a charge of e units and that a vertical electric field E is present, e and E both being measured in electrostatic units. Then a force eE is added to the weight and instead of (46) we have

$$\frac{4}{3}\pi a^3 \rho g + eE = 6\pi \eta a v_1,$$

v_1 being the new velocity of steady fall. From these equations we find

$$eE = 6\pi \eta a (v_1 - v_0);$$

and, inserting in this equation the value of a given by (46) and solving for e , we find

$$e = 6\pi \eta^{3/2} (v_1 - v_0) \left(\frac{9v_0}{2\rho g} \right)^{1/2} \frac{1}{E}. \quad (47)$$

All quantities in this last expression being known, e can be calculated.

Millikan found that the charges so calculated from his observations were all multiples of a smallest charge. The latter he assumed to be the charge on the electron, an assumption which is difficult to test directly but seems to be well supported by indirect evidence. For the electronic charge he found $e = 4.774 \times 10^{-10}$ electrostatic unit.

This latter value of e was accepted for many years. Its accuracy was first called into question by Backlin,¹ who, from values of X-ray wave lengths measured with a ruled grating found $e = 4.793 \times 10^{-10}$ e.s.u., a value 0.4 percent higher than that reported by Millikan. It now appears that the discrepancy was due to Millikan's having used too low a value of the viscosity of air, which appears as 77 in Eq. (47) above. Recent measurements give a value of 77 about 0.5 percent higher than Millikan's,² and this higher value leads to an "oil-drop value" of e of $(4.8036 \pm 0.0048) \times 10^{-10}$ e.s.u. Dunnington concludes that the best "ruled-grating" value of e is $(4.8025 \pm 0.0004) \times 10^{-10}$ e.s.u. The best modern value of the ratio e/m , which has been measured recently by different methods with high precision, appears to be 1.7592×10^7 . From e and e/m , the value of the electronic mass m can then be calculated.⁴

We thus obtain for these three important constants and for No.

¹ BACKLIN, E., "Absolute Wellenbestimmungen der Röntgenstrahlen," Dissertation, Upsala, 1928.

² Cf. DUNNINGTON, *Rev. Modern Phys.*, vol. 11, p. 71 (1939).

³ Cf. BIRGE, *Phys. Rev.*, vol. 13, p. 233 (1941); DUNNINGTON, *he. cit*

⁴ This value of m is the so-called "rest mass" of the electron, i.e., its mass when at rest or when moving with velocities negligibly small compared with the velocity of light. See Sec. 63.

representing Avogadro's number or the number of atoms in a gram-atom, calculated as F/e where F is the value of the faraday (Sec. 39):

$$\begin{aligned} e &= 4.803 \times 10^{-10} \text{ electrostatic unit,} \\ \frac{e}{m} &= 1.7592 \times 10^7 \text{ electromagnetic units,} \\ m &= 0.9107 \times 10^{-27} \text{ gram,} \\ N_0 &= 6.023 \times 10^{23}. \end{aligned}$$

It is often of interest, also, to know how the electron compares in mass with a hydrogen atom. The faraday, mentioned in Sec. 39, represents the charge carried by 1 gram-atom or 1.0081 grams of hydrogen; its value is 9,648.8 electromagnetic units. Thus for a hydrogen atom of mass M_H , which carries the same charge as the electron when ionized, $e/M_H = 9,648.8/1.0081$ or

$$\frac{e}{M_H} = 9.571 \times 10^3 \text{ electromagnetic units.}$$

Dividing this number into the value of e/m for the electron, we have, e canceling,

$$\frac{M_H}{m} = 1,837$$

as the ratio of the two masses. (Actually, the number obtained is 1,838, but 1,837 is probably more nearly correct; see Sec. 97.)

43. Photoelectrons.—The discovery of the electron at once suggested the hypothesis that the photoelectric effect is due to the liberation, from the illuminated metal plate, of electrons which, under the influence of the electric field, pass from cathode to anode, thereby causing the photoelectric current. This hypothesis was confirmed by Lenard,¹ who showed that the photoelectric discharge is deflected in a magnetic field exactly as are cathode rays. By measuring the deflection of the "photoelectric rays" in a known magnetic field, he found a value of e/m of (about) 1.2×10^7 , in qualitative agreement with Thomson's value of e/m for electrons.²

Lenard's method of determining e/m for photoelectrons involves basic principles which, with ever increasing refinement, have been widely applied in¹¹ charged-particle physics. His apparatus is shown diagrammatically in Fig. 15. A glass tube, which could be exhausted to the highest attainable vacuum through the side tube T, contained an *aluminum cathode C, which could be illuminated by ultraviolet light* from a spark S, *the light passing through the quartz plate Q. The*

¹ *Ann. d. Physik*, vol. 2, p. 359 (1900).

² Later, Alberti [*Ann. d. Physik*, vol. 39, p. 1138 (1912)] found for photoelectrons $e/m = 1.765 \times 10^7$ e.m.u. gram⁻¹.

cathode C could be charged to any potential, positive or negative. A screen A , with a small hole at its center and connected to earth, served as anode. P_1 and P_2 were small metal electrodes connected to electrometers. When C was illuminated and charged to a negative potential of several volts, photoelectrons were liberated and accelerated toward the anode A . A few electrons passed through the hole in the center of A and proceeded thereafter at uniform velocity to the elec-

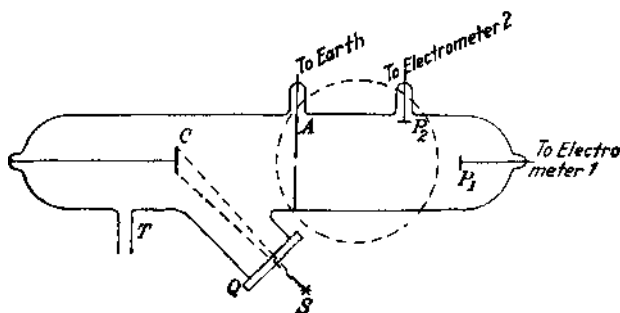


FIG. 15.—Lenard's apparatus for determining the ratio e/m for photoelectrons.

trode P_1 , their reception there being indicated by the electrometer 1. But if, by means of a pair of Helmholtz coils (represented by the dotted circle), a magnetic field directed toward the reader was produced in the region between A and P_1 , the electrons were deflected upward in a circular path and, with a sufficient field strength, struck the electrode P_2 .

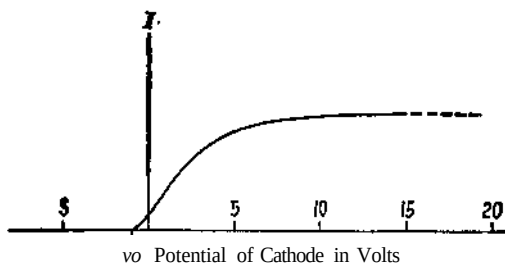


FIG. 16.—Variation of photoelectric current with cathode potential.

Lenard first investigated the relation between the current reaching the anode and the potential V applied to C , A being assumed to be always at zero potential. There was no photoelectric current when V was several volts positive. But, when V was dropped to about 2 volts *positive*, a small current was observed. This indicated that the photoelectrons were not simply *freed* from the cathode but that some of them at least *were ejected with sufficient velocity to enable them to overcome the retarding potential* of 2 volts. The current increased when V

was reduced to zero and increased still more rapidly as V was made negative but attained a "saturation" value after V had reached some 15 or 20 volts negative. These data are shown diagrammatically in Fig. 16. I is the photoelectric current and V_0 is the positive potential that was required to prevent the escape of electrons.

The determination of e/m was made essentially as follows. Let a negative potential V , large compared with V_0 , be applied to the cathode C, all potentials being in electromagnetic units. The photoelectron, on reaching the anode, will have a kinetic energy given approximately by

$$Ve = \frac{1}{2} mv^2,$$

where m and e are the mass and the charge in electromagnetic units of the electron, respectively, and v is its velocity on reaching A (Fig. 15). Assuming that after leaving A the electron moves in a uniform magnetic field, the circular path that it follows is determined by the equation

$$Hev = \frac{mv^2}{R},$$

where H is the strength of the magnetic field just necessary to cause the electron to reach P₂, and R is the radius of the corresponding circular path, determined from the geometry of the apparatus. From the last two equations we have

$$\frac{e}{m} = \frac{2V}{H^2 R^2},$$

from which e/m may be computed.

44. Relation between Photoelectric Current and Intensity of Illumination of the Cathode.—The experiments of Elster and Geitel, Lenard, and Ladenburg seemed to show that, as long as there is no change in the spectral quality of the light causing the emission of electrons, the photoelectric current is directly proportional to the intensity of illumination on the emitting surface. Subsequent experiments¹ confirmed the law of proportionality and showed that it holds rigorously over a very wide range² of intensities—in the experiments of Elster and Geitel, over a range of 50,000,000 to 1. This law of proportionality was, therefore, well established, and it is one of the most important laws of the photoelectric effect.

45. Energy Distribution of Photoelectrons.—In his experiments, Lenard showed, by observing the retarding potential against which these electrons could move, that for a given emitter their kinetic

¹ RICHTMYER, *Phys. Rev.*, vol. 29, pp. 71, 404 (1909).

² ELSTER and GEITEL, *Phys. Zeits.*, vol. 14, p. 741 (1913); vol. 15, p. 610 (1914); vol. 17, p. 268 (1916). KUNZ and STEBBINS, *Phys. Rev.*, vol. 7, p. 62 (1916).

energy did not exceed a definite maximum, given by V_0e , where V_0 is the maximum retarding potential for which a photoelectric current was observed and e is the electronic charge. From Lenard's time to the present, this question of the energy of photoelectrons has been the subject of many investigations.

We mention the experiments of Richardson and Compton,¹ who, by introducing important corrections and techniques, cleared up a number of previous discrepancies and gave impetus to the quantitative verification of Einstein's photoelectric equation (see next section). The apparatus used by Richardson and Compton is sketched in Fig. 17. The photoelectric emitter C , a strip 1 X 5 mm. of the metal under

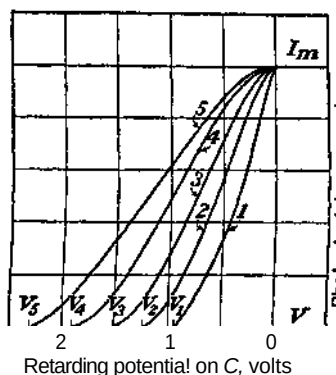
investigation, was placed at the center of a spherical glass bulb B some 7.5 cm. in diameter and silvered on its inner surface. The emitter C could be charged to any desired potential, read by voltmeter V . The bulb could be evacuated through the side tube T . Monochromatic light L passed through the quartz window W and fell on C . The silver coating on the inside of B , serving as anode A , was connected to the electrometer Q by which the photoelectric current was measured.

FIG. 17.—Apparatus of Richardson and Compton for observations on the energy of photoelectrons.

The large spherical anode, with the small emitter at its center, served two important purposes. First, since the electric field around C was nearly radial, it was possible to measure the energy distribution of the photoelectrons irrespective of direction of emission and thus to determine what has come to be known as the "total energy/" in contradistinction to the "normal energy," which is measured when cathode and anode are a pair of parallel plates, as in Fig. 9. Second, the impact of the photoelectrons on the anode caused the diffuse emission from the latter of a certain number of secondary electrons. In the case of parallel plates, a considerable proportion of these return to the cathode, and the observed current is, therefore, not the true photoelectric current. Only a few of these secondaries, however, reach the small cathode C at the center of the large anode, and the error in the observed current is negligible. Richardson and Compton made careful correction for the contact difference of potential between the cathode C and the silver anode, an important correction, for the maximum retarding potentials—1 or 2 volts—were of the same order of magnitude as the contact potentials.

¹ RICHARDSON and COMPTON, *Phil. Mag.*, vol. 24, p. 575 (1912).

The relation between photoelectric current and retarding potential obtained by Richardson and Compton is shown for aluminum as emitter in Fig. 18. When the aluminum surface was illuminated by ultra-violet light of constant intensity and of wave length 0.000020 cm., no photoelectric current was observed when the potential of *C* was greater than 2.3 volts—*V_h* curve 5. As the retarding potential decreased from this value to zero, the current rose to a maximum *I_m* beyond which there was no increase for negative potentials on the cathode. For longer wave lengths, similar curves—4,3,2,1—were obtained, but with the important difference¹ that the critical retarding potentials—*V₄*, *V₃*, *V₂*, *V₁*—became progressively less as the wave length increased. Curves of the same shape were obtained for other intensities of illumination, the current increasing in all cases in proportion to the intensity of the light.



Curve	Wave length, cm.	Critical potential, volts
(1)	0.0000313	<i>V₁</i> - 0.90
(2)	0.0000275	<i>v₂</i> - 1.30
(3)	0.0000254	<i>V₃</i> - 1.50
(4)	0.000023	<i>V₄</i> - 1.90
(5)	0.000020	<i>V_t</i> - 2.30

FIG. 18.—Variation of photoelectric current with retarding potential.

These observations are readily interpreted on the assumption that, for a given emitter, light of wave length *X* causes the emission of photoelectrons with initial velocities *v* varying from zero up to a maximum

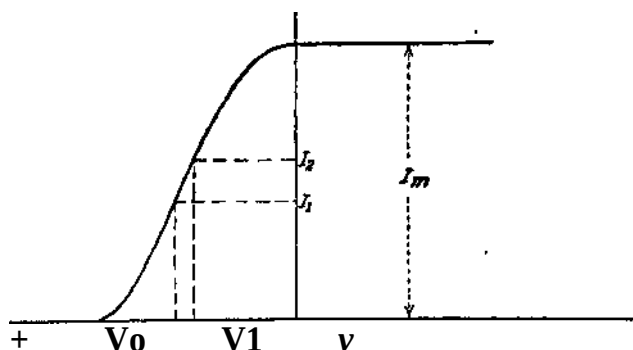


FIG. 19.—Typical variation of photoelectric current with retarding potential.

v_m determined by *X*. The typical variation of photoelectric current with retarding potential *V* is shown again in Fig. 19. Since no

¹ The intensity of these several monochromatic beams was adjusted to give the same value of *I_m* in each case.

photoelectric current was observed for retarding potentials greater than V_0 , v_m is at once given by

$$V_0 e = \frac{1}{2} m v_m^2$$

e being the numerical electronic charge. When the potential V applied to the emitter is zero or negative, all electrons freed by the light reach the anode and the photoelectric current has its maximum value I_m . When $V_0 > V > 0$, only those electrons will reach the anode for which

$$\frac{1}{2} m v^2 > V e.$$

Let the photoelectric currents be I_1 and I_2 when the potentials are V_1 and V_2 , respectively. Then $(I_2 - I_1)/e$ photoelectrons per second leave the emitter with velocities between V_1 and V_2 defined by

$$V_1 e = \frac{1}{2} m v_1^2 = E_1, \quad V_2 e = \frac{1}{2} m v_2^2 = E_2,$$

E_1 and E_2 being the corresponding kinetic energies. Thus the slope of the curve in Fig. 19 at any point is proportional to the number of photoelectrons possessing energy E corresponding to the value of V at that point. A curve plotted between these slopes and V is the energy distribution curve for photoelectrons; its general form is shown in Fig. 20.

46. Relation between the Velocities of Photoelectrons and the Frequency of the Light.—A very simple linear relation has been found to exist between the maximum energy of photoelectrons and the frequency ν of the light that causes their emission. As was shown by Millikan, if a curve is plotted between $V_0 e$ and ν , a straight line results which has an intercept ν_0 on the frequency axis (Fig. 21). The experimental meaning of this intercept is that light of frequency less than ν_0 cannot cause the emission of photoelectrons from the metal concerned. The quantity ν_0 is characteristic of the emitting electrode, but the slope of the curve is the same for all electrodes. The equation of the curve may be written

$$V_0 e = \frac{1}{2} m v_m^2 = h(\nu - \nu_0) = h\nu - h\nu_0$$

$$\frac{1}{2} m v_m^2 = h\nu - \omega_0,$$

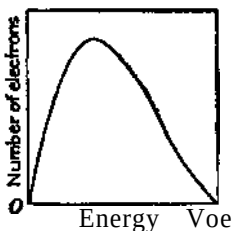


FIG. 20.—Distribution of photoelectrons as to energy.

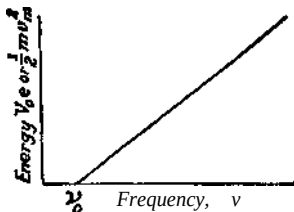


FIG. 21.—Variation of maximum energy of photoelectrons with frequency of the exciting radiation.

where h is the slope of the curve and w_0 is written for $h\nu_0$. If, in Fig. 21 and Eq. (48), the ordinates are expressed in ergs and the abscissas in frequencies (sec.^{-1}), the numerical value of h , as found by Millikan, is 6.56×10^{-27} erg sec. (A better value is 6.61×10^{-27} .) This constant h , called "Planck's constant" for reasons which will appear in Chap. V, is one of the fundamental constants of nature and has played an extraordinary part in modern physics. *The product $h\nu$ is called a quantum of energy corresponding to light of frequency ν .*

Equation (48) has had an interesting history and is one of the most fundamental equations, or laws, of modern physics. It was first proposed on theoretical grounds by Einstein, in 1905, as a result of the extension to the photoelectric process of the concept, previously developed by Planck (see Chap. V), that interchanges of energy between radiation and matter take place in energy quanta $h\nu$, where ν is the frequency of the radiation absorbed or emitted and h is a constant. Einstein arrived at the equation¹ by assuming that the whole quantum $h\nu$ of radiant energy was absorbed by an electron but that a part w_0 of it was expended by the electron in escaping from the emitter. He had at his disposal only qualitative data to show that his equation gave results of at least the right order of magnitude. The equation received final and complete experimental verification as a result of the precise experiments of Millikan.² Subsequent work extended the validity of the equation to the X-ray region, where frequencies are involved that are several thousand times the frequencies of visible light. Einstein's photoelectric equation played an enormous part in the development of the modern quantum theory. But in spite of its generality and of the many successful applications that have been made of it in recent theories in physics, the equation is, as we shall see presently, based on a concept of radiation—the concept of "light quanta"—completely at variance with the most fundamental postulates and conclusions of the classical electromagnetic theory of radiation.

47. Other Properties of Photoelectric Emission.—It is not our purpose to give a complete account of the properties of the photoelectric effect, but a few other features of interest may be mentioned in passing.

The critical or "threshold" frequency ν_0 has been found to vary considerably with the state of the surface, as does the photoelectric current in general. Usually ν_0 lies in the ultraviolet, but for the alkali metals and for barium and strontium it lies in the visible region;

¹ EINSTEIN, *Ann. d. Physik*, vol. 17, p. 132 (1905).

² MILLIKAN, *Phys. Rev.*, vol. 7, p. 355 (1916).

for potassium ν_0 lies in the red, and for cesium even in the infrared. Photoelectric cells used for the measurement of light commonly contain either an alkali metal or barium.

' *Variation of temperature* usually has little or no effect upon the photoelectric current, so long as the temperature does not exceed several hundred degrees centigrade and so long as no change occurs in the crystal structure or in the physical state of the metal. The alkalis, however, form an exception to this rule.

If *polarized light* is used, differences are commonly found as the plane of polarization is rotated, except, of course, at normal incidence. The effect is complicated, and its cause is in doubt. Especially interesting is the "selective effect" in the alkali metals. Over certain ranges of wave length, the photoelectric current from these metals is much greater when the electric vector in the light has a component perpendicular to the surface than when it is parallel to the surface; in the case of certain sodium-potassium alloys, the ratio of the currents in the two cases may be 10:1 or 20:1 or even more.

For further discussion of the photoelectric effect the student is referred to the treatise of Hughes and DuBridge.¹

We shall return here to those fundamental problems presented by the photoelectric effect which have a bearing on quantum theory.

48. Source of the Photoelectric Energy.—Whence comes the energy which the photoelectrons are observed to carry away from the surface of the emitter? Does it come from energy previously stored in the atom, or by direct absorption from the incident light? And, in the latter case, what is the mechanism by which the energy is absorbed?

The experimental facts are such as to suggest that the energy is obtained from the incident light. This presumption is strongly supported by the fact that (1) the photoelectric emission varies in strict proportion to the intensity of the light and (2), where the frequency of the light is given an increment $\Delta\nu$, the maximum kinetic energy of the photoelectrons is increased by an amount $h \cdot \Delta\nu$ that is *independent of the nature of the emitting substance*. But how can we picture a mechanism for the absorption of the radiant energy that will function in accordance with these facts, and also in such a way that (3) the *maximum kinetic energy* of the photoelectrons is *rigorously independent of the intensity of the light*?

The difficulties in the way of postulating such a mechanism on the basis of the classical wave theory of light can best be shown by some numerical data.

¹ HUGHES, A. L., and L. A. DUBRIDGE, "Photoelectric Phenomena," 1932.

A photoelectric current of 1×10^{12} e.m.u. per square centimeter is easily obtained from a sodium surface *in vacuo* with very moderate intensity of illumination. This current corresponds to the emission of about 6.25×10^7 electrons per square centimeter per second. These electrons come from the surface layers of atoms. We can easily compute the order of magnitude of the number of atoms involved. Calling Avogadro's number N_0 , the atomic weight of the emitter A , and its density ρ , the number of atoms per unit volume n is given by

$N_0 = 6.02 \times 10^{23}$; and, for sodium, $A = 23.0$ and $\rho = 0.97$ gram per cubic centimeter. Therefore, $n = 25.4 \times 10^{21}$ atoms per cubic centimeter. Assuming the atoms in a centimeter cube of sodium to be arranged in regular rows, columns, and layers,¹ we find that there are $\sqrt[3]{25.4 \times 10^{21}} = 2.94 \times 10^7$ atoms along each edge, or $(2.94 \times 10^7)^2 = 8.64 \times 10^{14}$ atoms in a layer 1 cm.² in area and 1 atom deep. Assuming, for the sake of illustration, that the photoelectrons are supplied by, say, the first 10 layers of atoms, we find that

$$8.64 \times 10^{15} \text{ atoms}$$

furnish, under the conditions of illumination above mentioned, 6.25×10^7 electrons per second. That is, 1 atom in

$$\frac{8.64 \times 10^{15}}{6.25 \times 10^7} = 1.38 \times 10^8,$$

on the average, furnishes 1 photoelectron per second. But what is it that determines *which one* of these hundred million atoms shall furnish an electron in any particular second? According to the wave theory of light, radiant energy should be distributed continuously and uniformly over the wave front. The atoms are presumably all alike. *On the basis of the classical theory of radiation*, we are unable to put forward any reasonable hypothesis that should result in singling out the one atom in a hundred million.

Furthermore, it is hard to understand how the energy that is actually carried off by a photoelectron comes to be concentrated in one spot. For example, an illumination of 0.1 meter-candle, or perhaps 10 ergs per cm.² per sec, produces an easily measurable photoelectric current from sodium. Not over 0.05 erg of this, however, falls within the spectral range that is effective for sodium, which is

¹ The crystal structure of sodium is in reality a body-centered cube. See "International Critical Tables," vol. I, p. 340.

sensitive only in the blue and near-ultraviolet part of the spectrum. Furthermore, sodium has a high reflecting power, and only *absorbed* light can possibly be used in expelling electrons. We conclude that not over $0.001 \text{ erg per cm.}^2 \text{ per sec.}$ is available for the photoelectric process.

According to the electromagnetic theory, this absorbed energy should be equally distributed among the 8.6×10^{15} atoms which form the (10, arbitrarily assumed) surface layers, so that there is $1.1 \times 10^{-19} \text{ erg per sec.}$ for each atom. Now a quantum of violet light corresponding to $\lambda = 0.00004 \text{ cm.}$ is

$$h\nu = 6.61 \times 10^{-27} \times 0.75 \times 10^{15} = 5 \times 10^{-12} \text{ erg.}$$

Thus, if an atom can absorb only the energy that falls on it (so to speak), the number of seconds required for it to obtain enough energy to expel a photoelectron is

$$\frac{5 \times 10^{-12}}{1.1 \times 10^{-19}} = 4.5 \times 10^7 \text{ seconds.}$$

This is over 500 days!

This calculation may not be quite correct, however. Theoretically, it can be shown that an electron vibrating in *resonance* with the light can absorb as much energy as falls on a considerable fraction of a *square wave length*. There are difficulties about assuming photoelectric emission to be a resonance phenomenon, however; and it is hard to see why there should be a resonator actively absorbing in one atom while all of the neighboring atoms are quiescent. If, however, we brush aside these objections, we may conclude that, in the present case, an atom might possibly absorb the energy falling on 10^{-10} cm.^2 , or $10^{-10} \times 10^{-3} = 10^{-13} \text{ erg per sec.}$ The time required to absorb a quantum would then be only $5 \times 10^{-12} / 10^{-13} = 50 \text{ seconds.}$ Even this, however, is too long a time to agree with experiment. No time lag has ever been observed between the beginning of illumination and the starting of the photoelectric current; precise measurements have shown that, if there is such a time lag, it is less than 3 billionths of a second!

The classical electromagnetic theory of light thus encounters enormous difficulties in attempting to explain the observed facts in regard to photoelectric emission.

49. The Photoelectric Effect and the Corpuscular Theory of Light.

If we could return to the corpuscular theory and regard light as a rain of "corpuscles" or quanta of energy, all of the difficulties raised in the preceding section would disappear. If we could assume (*a*) that each

quantum, or, in modern terms, "photon," possesses energy $h\nu$ and (6) that a collision between a photon and an atom may result in the absorption of the whole photon and the emission of a photoelectron with the initial energy $h\nu$, we could at once explain why only one atom in many millions expels an electron in any particular second and also why there is no time lag in the photoelectric process.

Such a theory would predict, also, that the photoelectric current should be proportional to light intensity, *i.e.*, to the number of photons striking, as is experimentally observed. Furthermore, we should have a ready explanation of the observed fact that the initial velocity of the photoelectrons expelled by monochromatic light is independent of the intensity of the light; for each photon would act independently upon the emitting substance.

The difficulties with such a radical theory of light, however, are many. For one thing, if we regard light as a "shower" of photons, what can possibly be the meaning of *frequency*? There is nothing periodic about a falling raindrop, for example. As a matter of fact, in order to find the frequency of a beam of light, we *measure* (a) the velocity c of the light and (b) its wave length λ *on the assumption that light consists of waves*, and then we *compute* the frequency as

$$\nu = \frac{c}{\lambda}$$

Thus we have to rely on the wave theory of light to give us the energy value $h\nu$ of a photon! And there still remains the phenomenon of interference, which, since its discovery by Young in 1802, has defied explanation on any other basis than by assuming light to be a wave phenomenon. However, the experimental facts of photoelectricity are equally as cogent as the phenomena of interference, and these *cannot be explained on the basis of the classical wave theory of light.*

Thus in 1920, say, the physicist could sum up the situation about like this: On one side of a seemingly impenetrable barrier, or fence, is to be found a group of phenomena—such as interference, polarization, smaller velocity of light in optically denser bodies, indeed, the whole electromagnetic theory and its ramifications—according to which we should say, without the slightest doubt, that light *must consist of waves*. On the other side of the fence is to be found another group—the photoelectric effect, and other phenomena which we shall consider in subsequent chapters—according to which we should say, again without the slightest hesitation, if we did not know what was on the first side of the fence, that light *must be corpuscular*. The situation thus created was perhaps the most puzzling one that has ever arisen in the

whole history of physics. We cannot resolve the mystery at this point. We shall return to the question after we have become familiar with the quantum theory of atomic structure and of spectra.

The remainder of the present chapter will be devoted to another aspect of the photoelectric problem, and to a different but related phenomenon involving electrons. Granted that the photoelectrons receive their energy in some manner from the incident light, do these electrons come out of the atoms, or do they come from among the so-called "free" electrons in the metal? Before attempting to answer this question, we shall discuss the spontaneous emission of electrons from hot bodies.

60. Thermionic Emission.—It has been known for 200 years that air in the neighborhood of hot solids has the power of conducting electricity. In a systematic investigation begun about 1880, Elster and Geitel showed that, in general, at a red heat, charged bodies lose a positive charge more readily than they do a negative one, whereas at a white heat a negative charge is more readily conducted away; a few substances, however, lose a negative charge most readily at all temperatures. About the same time, Edison noticed that a current can flow in an evacuated bulb from a glowing carbon filament to another filament when the hot filament is charged negatively but not when it is positive.

The ions that carry the electricity in such cases were studied by McClelland and by J. J. Thomson. In 1899, Thomson measured e/m , the ratio of charge to mass, for the *negative* ions, using the method described in Sec. 41. He found this ratio to have about the same large value as for cathode rays and concluded that these negative ions were electrons. Later he showed that, for the positive ions, e/m was in all cases of the same order of magnitude as for electrolytic ions and concluded that the positive ions were positively charged atoms or molecules.

This emission of electrons from hot bodies furnishes the basis for the thermionic vacuum tubes that have come into such wide use. No general treatment of thermionic emission will be attempted here¹; we shall discuss only a few aspects of the phenomenon that are important in connection with the theory of the emission of electrons from conductors.

The thermionic current from a given emitter is found to increase very rapidly with increasing temperature. It depends also upon geometrical factors, which determine space-charge effects, and upon the potential of the emitter relative to its surroundings. From a given emitter operating at a given temperature, however, the thermionic

¹ Cf. REIMANN, A. L., "Thermionic Emission," 1934.

current **cannot** be made to exceed a **certain limiting value**. **Special** interest attaches to this maximum possible thermionic current. Let I denote the maximum current obtainable per square centimeter of the surface from a given emitter operating at absolute temperature T . By means of thermodynamic reasoning, it can be shown that I should vary with temperature, approximately, at least, according to Richardson's equation

$$I = AT^2 e^{-\frac{e\varphi}{kT}}. \quad (49)$$

The proof is rather elaborate and will not be given here;¹ the equation itself can be understood without going through its deduction. The Napierian base of logarithms is denoted here by e to distinguish it from the numerical electronic charge e ; k is the Boltzmann constant (1.381×10^{-16} erg per degree centigrade), and φ and A are constants depending on the emitting substance and on the state of its surface.

The Richardson equation, as stated in (49) above, is found to represent the experimental facts very well. The variation of the exponential factor in the equation is so rapid, however, that it is not possible to tell whether the factor T^2 is really correct. An argument from quantum theory indicates that the constant A ought, perhaps, to be a universal constant, equal to 60 amperes per square centimeter for all substances; but experimentally it is found to vary over a considerable range from one substance to another.

The quantity $e\varphi$ in the Richardson equation represents an energy. For this reason, φ , known as the "work function," is commonly measured in volts, representing the change of potential through which the electron must pass in order to gain or lose $e\varphi$ ergs of energy. In the theoretical derivation of the equation, $e\varphi$ enters as representing the "heat of emission" of the thermions at the absolute zero of temperature, or, in other words, it is the amount of energy required to extract a thermion at that temperature. The value of φ can be calculated from observations of J by plotting $\log(I/T^2)$ against $1/T$; according to Richardson's equation

$$\log \frac{I}{T^2} = \log A - \frac{e\varphi}{k} \frac{1}{T}$$

so that a straight line is obtained whose slope is $-e\varphi/k$. Values of φ so obtained, when converted into volts, range from 2 to 6 volts for various metals.

¹ See RICHARDSON, O. W., "Electron Theory of Matter," Chap. XVIII, 1916; REIMANN, *loc. cit.*

51. Relation between Thermionic and Photoelectric Constants.—It

is obvious that the photoelectric effect and the thermionic emission of electrons are in some respects similar phenomena. They seem to differ mainly in the manner in which the electron is given the additional energy to enable it to escape through the surface of the emitter against the potential barrier. In the thermionic case, the electron acquires this energy from the thermal agitation incident to the high temperature of the emitter; in the photoelectric case, by the absorption of light. In the case of thermionic emission, it is natural to suppose that the heat of emission $e\phi$ represents work done by the electron in escaping from the emitter, just as the larger part of the heat of evaporation of a liquid represents work done by the evaporating molecules against the attraction of the molecules left behind in the liquid. If that is so, $e\phi$ ought to be the same as the quantity w_0 or $h\nu_0$ in the Einstein photoelectric equation, which was assumed to represent work done by the photoelectron. Thus, the thermionic work function ought to be related to the photoelectric threshold, the ν_0 of Fig. 21, by the equation

$$\phi = \frac{w_0}{e} = \frac{h\nu_0}{e},$$

ϕ

ϕ

Metal	$h\nu_0/e$, volts	volts
Platinum..	6.30	6.27
Rhodium	4.57	4.58
Tantalum.	4.05	4.07
Tungsten	4.58	4.52
Palladium. .	4.96	4.99
Silver . . .	4.73	4.08
Gold	4.82	4.42

The agreement between the values of the work function determined photoelectrically and those determined thermionically seems good enough to warrant the hypothesis that the photoelectrons and the thermionic electrons have a common origin.

¹ Quoted from HUGHES and DUBRIDGE, *loc. cit.*

52. Velocities of Emission of Thermions.—The velocities with which thermions are emitted from the surface of the emitter can be measured in the same way in which the velocities of photoelectrons are measured (Sec. 45). The velocities of the thermions are found to be very much smaller than those of photoelectrons and to be distributed according to Maxwell's law, just as if they were molecules in a gas at the temperature of the emitter.

To see clearly what this latter statement means, suppose that the electrons emerge into a space entirely surrounded by the emitting substance, all at the same temperature T ; *i.e.*, consider emission into a cavity in the emitter. As electrons accumulate in the cavity, they will begin to return to the emitter, and an equilibrium condition will come into existence in which as many electrons return to the emitter per second as issue from it. The cavity is then filled with an *electron gas* in equilibrium at temperature T . The density of this gas is so low that it will obey the laws of perfect gases. The rate at which molecules, *i.e.*, electrons, return to the emitter from such a gas can be calculated from kinetic theory in terms of the density and temperature of the gas. If the assumption is made that all of the electrons enter the emitter as they strike its surface, none of them being reflected, then the rate of return must equal the rate of emission as given by Richardson's equation. In this way, the density and pressure of the electron gas corresponding to any temperature can be calculated. In order to preserve equilibrium, the emitted electrons must be distributed in velocity in the same manner as are the returning electrons, except for a reversal of all velocities. Hence, their distribution in velocity can also be calculated, from Maxwell's law for molecular velocities in a gas. It follows from kinetic theory that the mean kinetic energy of the electrons as they come from the emitter must be $2kT$; this amounts only to 0.22 electron-volt at 1000°C. (An *electron-volt* is the work done on an electron as it passes from one point to another between which there is a potential difference of 1 volt. It equals

$$4.803 \times \frac{10^{-10}}{300} = 1.601 \times 10^{-12} \text{ erg.})$$

As a concluding remark concerning thermionic emission, it may be mentioned that if an accelerating electric field is applied to the emitter, the maximum thermionic current is somewhat increased. As a matter of fact, very intense fields, of the order of 10^8 volts per cm., are able to cause small currents of electrons to flow from metals even at room temperature. These are sometimes called "field currents."

53. Theories of Electrons in Metals.—An explanation of all the features of photoelectric and thermionic emission can be given only on the basis of a theory of the electrical constitution of matter. At this point, therefore, the electron theory of metals will be very briefly discussed.

Soon after the discovery of the electron, it was suggested that electrical conductivity in metals may be due to the existence, in the interatomic spaces, of "free" electrons, which, drifting under the influence of an applied electric field or potential gradient, cause the transport of charge, the actual direction of motion of the electrons being opposite to the usual convention concerning the *positive* direction of current. This theory was developed into the well-known *classical* theory of metallic conduction by Drude, Lorentz,¹ and others. They assumed that the free electrons move about and collide with each other just as do the molecules of a gas, having the usual maxwellian distribution of velocities proper to a gas of electrons at the temperature of the metal. They showed that such free electrons would account, also, for the high *thermal* conductivity of metals; because of their light mass, according to classical theory, the electrons would have much higher velocities of thermal agitation than do the atoms, and for this reason they would be, in proportion to their numbers, much more effective in the conduction of heat.

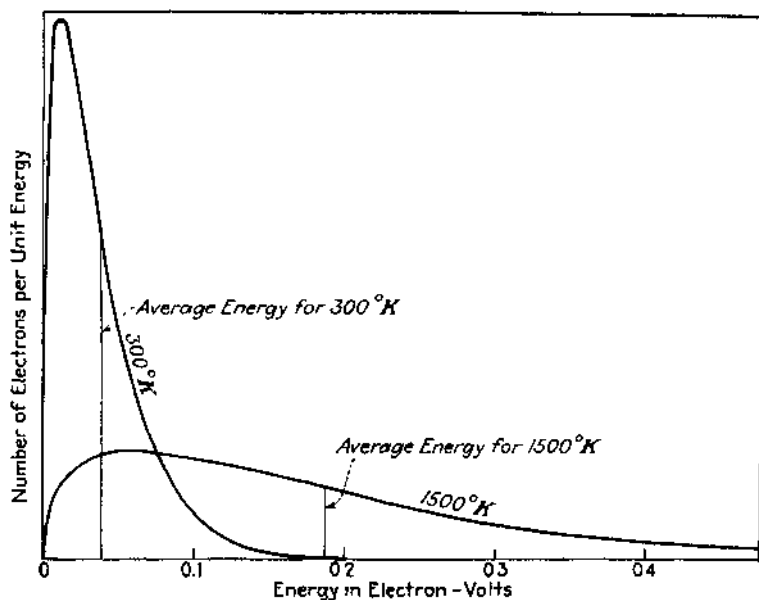
The theory had some success in accounting for the properties of the electrical and thermal conductivities of metals. Most of the formulas that were obtained, however, contained as constants the number of free electrons per unit volume and also the mean free path of the electrons among the atoms, quantities that were not independently known. Furthermore, a serious difficulty was encountered in connection with specific heats. The specific heat of a metal is fully accounted for by the kinetic and potential energy which the atoms themselves should possess, according to the classical principle of the equipartition of energy (see Sec. 79). It was necessary to suppose, therefore, that for some reason the heat energy of the free electrons was very small, or at least independent of temperature. Several lines of evidence indicated, however, that the number of the free electrons should be of the order of magnitude of the number of atoms; and, according to the principle of equipartition, a free electron should, have the same average kinetic energy as an atom. Thus, it was hard to understand why the specific heats of good conductors were not considerably larger than the experimental values.

This last difficulty was removed completely in a new theory

¹LORENTZ, H. A., "Theory of Electrons."

introduced by Sommerfeld in 1928. His theory may be regarded as a first step toward the modern wave-mechanical theory of metals. The latter theory has been developed to a certain point during the last few years, but it remains incomplete and rather complicated. Its conclusions are reproduced in the Sommerfeld theory with sufficient completeness and accuracy for the application to the more important features of photoelectric and thermionic emission, hence the discussion will be limited here to the Sommerfeld theory.

In this theory the concept of a gas of free electrons within the metals is retained but this gas is treated in terms of wave mechanics. Because



of their light mass and their high density in the metal, the electrons form a degenerate gas of the "Fermi-Dirac type," being so named after the two physicists who first worked out the theory of such a gas. The electrons have a distribution of velocities very different from the classical one as represented by Maxwell's law.

According to classical theory, the electrons should all have zero energy of translation at absolute zero. As the temperature rises, their energies of translation should increase and at $T^{\circ}\text{K.}$ should have an average value per electron of $3/2kT$, the distribution of energy among the electrons being the familiar Maxwell distribution curve shown in Fig. 22 for temperatures of 300. and 1600°K. A few electrons

have high energy, but the majority are grouped around the mean value $3/2kT$, which at 1500°K . is equivalent to some 0.2 electron-volt.

For comparison, the distribution of energy given by the Fermi-Dirac theory is shown in Fig. 23 for the electrons in platinum at three temperatures: absolute zero, 300, and 1500°K . This distribution is calculated from wave-mechanical theory, but it would lead us too far afield to explain how it is derived. It is to be observed that the Fermi-Dirac distribution does not give zero energy for all free electrons at absolute zero, as is required by the classical theory. Instead, the graph for absolute zero starts from the origin and rises according

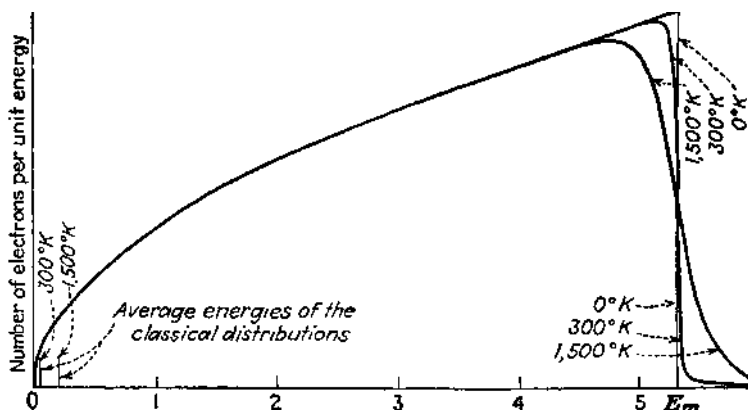


FIG. 23.—Distribution of energy among the conduction electrons of platinum at three different temperatures, according to the Fermi-Dirac-Sommerfeld theory.

to an equation of the type $y = a\sqrt{x}$ until a maximum energy E_m is reached, at which point the graph drops suddenly to zero. No electrons possess energy greater than E_m . According to the theory, the value of E_m is given¹ by

$$E_m = \left(\frac{h^2}{2m}\right)\left(\frac{3n}{8\pi}\right)^{2/3},$$

where h is Planck's constant, m is the mass of the electron, and n is the number of electrons per unit volume. If it be assumed that the number of free electrons per unit volume is of the same magnitude as the number of atoms per unit volume, then E_m for platinum (6.6×10^{22} electrons per cubic centimeter) is approximately 6 electron-volts. The theory shows that the mean value E of the energy per electron at absolute zero is

$$E = 3/5 E_m = 3.6 \text{ electron-volts for platinum.}$$

¹ Cf. KBNNABD, E. H., "Kinetic Theory of Gases," Secs. 238, 240, 1938.

This is to be compared with the corresponding mean values given by the classical theory, *e.g.*, 0.2 electron-volt for 1500°K. quoted above.

These results are astonishing to any one who is used to thinking in classical terms, for they mean that even at absolute zero the free electrons in a metal have enormous energies. Their energy may be thought of as kinetic in nature, but the significance of the term "kinetic" is less clear in wave mechanics than in classical theory.

The *rate of increase* of the energy with temperature, on the other hand, is much less than it is according to classical theory. Referring again to the distributions in energy for platinum at 300°K. and at 1500°K. as shown in Fig. 23, we note that only in the neighborhood of E_m is there any significant change in distribution with temperature. As the temperature rises, the sharp discontinuity for absolute zero at E_m becomes "rounded off" more and more and an increasing number of electrons come to possess energy greater than E_m . A few electrons even have energies up to several times E_m ; the curve has a long tail (extending, theoretically, up to all energies), and in this tail it can be shown that the distribution of energies follows approximately Maxwell's law, just as in the classical theory of gases.

From the curves it is obvious, furthermore, that the *total energy* of the free electrons changes very little with rise of temperature. Hence the free electrons will contribute little to the *change* in energy of the metal as its temperature rises, *i.e.*, to its specific heat. Thus the difficulty noted above in regard to specific heats disappears.

54. Origin of Photoelectrons and Thermions.—Either the classical or the Sommerfeld theory furnishes at once an explanation of *thermionic* emission. Presumably, the free electrons are retained in the metal by forces of attraction. There must be as many positive units of charge on the atoms as there are free electrons, and the electrons will be held by the attraction of these positive charges. If, for any reason, an electron escapes from the metal, as it moves away, it will be drawn back by an attraction that may be regarded as arising from its "electrical image" in the surface of the metal, this image being caused by the repulsion of the other electrons by the escaping one and the consequent exposure of positive charge. Thus, in escaping, the electron will do a certain amount of work against this attraction, which will be denoted by $\mathfrak{I}$. This work represents an increase in the potential energy of the electron. In the metal there will always be a few free electrons with energies exceeding $\mathfrak{I}$. These electrons will be able to escape as thermions. As the temperature rises, the number of electrons with energies greater than $\mathfrak{I}$ will increase rapidly; hence

the thermionic current will increase rapidly with rising temperature, as is observed to be the case.

In emerging from the metal with kinetic energies proper to thermal equilibrium at the temperature of the metal, the electrons will undergo a change in energy, actually a gain, for which we may write $e\phi$, e being the numerical electronic charge. Energy equal to $e\phi$ must be furnished to them from an external source. The relation between the energy $e\phi$ and the work Ω differs on the two theories.

According to the classical theory, the average kinetic energy possessed by the free electrons in the metal is the same as the average kinetic energy possessed by electrons in an electron gas surrounding the metal and in thermal equilibrium with it. Hence, on the classical theory, the thermions lose kinetic energy on their way out only because of the work that they do against the attractive forces, which result in a conversion of kinetic into potential energy, and accordingly $e\phi = \Omega$.

In the Sommerfeld theory, on the other hand, the energy of the free electrons in the metal is much greater than that of an electron gas at the same temperature. This excess energy of the free electrons furnishes a large part of the work Ω . The thermions come effectively from levels close to the maximum energy E_m hence their average net gain in energy is only

$$e\phi = \Omega - E_m. \quad (51)$$

This equation becomes strictly correct as the temperature approaches absolute zero. Hence ϕ as defined here represents the same quantity that is denoted by ϕ in Richardson's equation (49).

For platinum ϕ is 6.3 volts and E_m/e is at least 6 volts; hence Q/e is at least 12.3 volts. For the metals in general, Ω/e is 5 to 20 volts higher than the ordinary "work function" ϕ . These high values of ϕ are qualitatively confirmed by observations on the index of refraction of electron waves (c/. Sec. 108 in Chap. VII).¹

The relation between these various quantities is shown diagrammatically in Fig. 24. The heavy curve shows the way in which the potential energy of an electron varies as the electron, coming from the left in the figure, passes through the surface of the metal into the space outside. The potential energy rises by a total amount Ω . The ordinate of any point above this line may represent the energy of an electron; the height of the point above the heavy line will then represent its kinetic energy. The horizontal lines at the level marked E_m and below indicate the Fermi energy levels (actually much more filiosely spaced) at 0°K.; at any other temperature, there are a few

¹ See also THOMSON, G. P., "The Wave Mechanics of Free Electrons," 1930.

electrons with energy above E_m . The additional energy possessed by these few electrons constitutes energy of thermal agitation; its relation to ϕ and ω_0 is a fine detail that will not be considered in our simplified description of the theory.

Origin of Photoelectrons.—There was long a difference of opinion as to whether photoelectrons come from among the free electrons or out of the interior of the atoms. On either hypothesis, it is easy to explain the fact that all velocities occur up to a maximum value v_m given by the Einstein equation

$$\frac{1}{2} m v_m^2 = h\nu - \omega_0.$$

Light is known to penetrate a distance equal to many atomic diameters into the metal; hence, photoelectrons should originate at various depths. In order to reach the surface, the electron would have to pass by (or through) many atoms, and collisions would be probable,

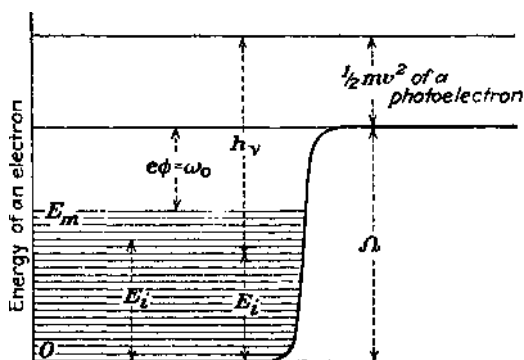


FIG. 24.—Energy of electrons in a metal. An electron with extra energy $e < p$ can just escape as a thermion. A photoelectron absorbs $h\nu$ and retains $\frac{1}{2} m v^2$ as kinetic energy.

at each of which the electron might lose some of the energy $h\nu$ initially given to it by the light. Some electrons would lose almost all of their initial energy in this way and would emerge with velocity almost zero. Others would lose less energy. Those which start very close to the surface would lose the least energy and would emerge with the greatest velocities; their loss of energy is represented in the equation by the constant term ω_0 .

If, now, the photoelectrons come from among the free electrons, they will have an initial energy E_i due to thermal agitation and, after absorbing $h\nu$ from the light, will start off with energy $h\nu + E_i$. The photoelectrons produced near the surface should thus emerge with energy

$$\frac{1}{2} m v^2 = h\nu + E_i - \omega_0.$$

Since E_i has no upper limit, there should exist, theoretically, no maximum velocity for a given frequency ν , and no sharp threshold ν_0 , the curves tailing off instead of plunging sharply into the axis. According to classical theory the value of E_1 is small; according to the classical principle of the equipartition of energy, or Eq. (100) in Sec. 79, the average value of E_i is

$$\bar{E}_i = \frac{3}{2}kT,$$

where T is the absolute temperature and k is the Boltzmann constant (1.381×10^{-16} erg deg.⁻¹). At 300°K. E_1 is equivalent to about 0.04 electron-volt; at 1500°K., to about 0.20 electron-volt. Practically, therefore, a fairly definite maximum velocity v_m and threshold frequency ν_0 would exist. But, since E_i is proportional to T , the apparent values of v_m and VQ should vary somewhat with temperature. Most of the experimental evidence indicates, however, that a variation of the right order of magnitude does not exist, at least not in all cases. Formerly, this was commonly accepted as evidence that the photoelectrons come from the interior of the atom and not from among the free electrons, since there was no reason for believing that the *internal energy of atoms* depends on temperature over the range covered by the experiments.

The assumption that the photoelectrons come out of the atoms was supported by observations on the photoelectric effect with X-rays, the frequencies of which are several thousand times those of ordinary light. In these experiments (see Sec. 184) there is very direct evidence not only that photoelectrons come from atoms but that they come from different "depths" within atoms and, in order to leave the atom, must expend quantities of energy which can be directly measured and which can be checked by independent means.

Further, the discovery of the Compton effect with X-rays (see Sec. 186) and its explanation emphasized the importance of another objection to the theory that the photoelectrons come from the free electrons. It would be impossible for a truly free electron to absorb a quantum of radiation, for the laws of the conservation of energy and of momentum could not simultaneously hold. If an electron, free and at rest, acquires kinetic energy $\frac{1}{2}mv^2$ by the absorption of a quantum $h\nu$, we should have, from the laws of conservation of energy and of momentum,¹

$$h\nu = \frac{1}{2}mv^2, \quad \frac{h\nu}{c} = mv,$$

where c is the velocity of light, m the mass of the electron, and v

4, ¹ See Sec. 186(a) for the momentum of a photon.

its velocity after absorbing the quantum. These equations can hold simultaneously only if $v = 2c$ —an impossibility. But if the electron absorbs light while in an atom, the momentum of the rest of the atom may change, and the above laws of mechanics may hold simultaneously.

If, however, photoelectrons ejected by visible and ultraviolet light originate in the atom, a new difficulty arises. The electron will then have to do work in order to escape from the atom into the interatomic space, and this work will be included in w_0 . Thus, w_0 would necessarily exceed the work of escape $e\phi$ of a thermion, which starts presumably in the interatomic space. The difference ought to be of the order of the work required to ionize an atom, which, at least for atoms in the vapor state, amounts always to several electron-volts. Yet it is found experimentally that w_0 and $e\phi$ differ (at least usually) by less than 1 electron-volt, often by less than 0.1 electron-volt.

Thus, the classical theory was not able to furnish a satisfactory answer to the question as to the origin of the photoelectrons. The new theory, on the other hand, removes all objections to supposing that photoelectrons liberated by visible or ultraviolet light (in contrast with most of those liberated by X-rays) come from the so-called "free" electrons. For, according to the new theory, these electrons are not really free; they are simply bound to the metal as a whole rather than to an individual atom. They differ from the electrons in the interior of an atom chiefly in that they are subject to forces exerted by several atoms at once. The difficulty regarding conservation of momentum, therefore, does not arise, since a reaction upon the metal itself can occur whenever a photoelectron is ejected.

The situation is improved, also, in regard to the variation with temperature. Even on the Sommerfeld theory there ought, theoretically, to be no true limit to the maximum energy of ejection, and hence no frequency threshold; for a few free electrons are present with all energies. The curves in Figs. 18 and 19 ought, therefore, to be rounded off at the end, so as to meet the axis asymptotically instead of plunging into it at a finite angle (except at $T = 0$). This effect would be much smaller, however, than the effect required by classical theory, for the number of electrons in the maxwellian tail of the Fermi-Dirac distribution is extremely small. Furthermore, there is experimental evidence that a slight rounding off of the curves at this point does exist.¹

At the beginning, the Sommerfeld theory constitutes a more adequate theory of metals, which can be the basis of wave mechanics. No attempt can

be made here to describe the progress that has been made toward the development of a complete theory; the student must be referred to other books.¹

¹See, **for** example, WILSON, A. H., "Theory of Metals," 1936; SEITZ, F., "Modern Theory of Solids," **1940**; also nonmathematical articles by Seitz and **Johnson** in the *J. Applied Phys.*, vol. 2, p. 8, and by Shockley in the *Bell System Tech. J.*, vol. 17, p. 645 (1939).

CHAPTER IV

THE THEORY OF RELATIVITY

The greatest revolution in physical thought during the twentieth century undoubtedly has been that which resulted from the introduction of the quantum theory. This, however, was not the first notable innovation in modern physical theory. In its own way, the new form of relativity put forward by Einstein in 1905 made an equally complete break with classical ways of thought. Relativity touches all branches of atomic physics here and there, but in itself it occupies a detached position, more fundamental than that of other parts of physical theory. For this reason it seems appropriate at this point to devote a chapter to a discussion of the theory of relativity. Any student who does not wish to interrupt the study of atomic physics at the present juncture can omit the present chapter without appreciable embarrassment in reading subsequent chapters.

65. Newtonian Relativity.—The subject of speculation since ancient times, it was early recognized that all motion involves displacement *relative* to something or other; but ideas have varied in regard to the entity relative to which the displacement occurs. In his treatise on mechanics, Newton says that "absolute motion is the translation of a body from one absolute place to another absolute place."¹ But he does not explain what he means by "absolute place." He states explicitly that the physicist can detect translatory motion only in the form of motion *relative to other material bodies*.

-Motion involves, also, the passage of time. Until recently, time was regarded as something essentially distinct from space or from the behavior of material bodies. Newton says, "Absolute, true, and mathematical time, of itself, and by its own nature, flows uniformly on, without regard to anything external."² Thus there was supposed to be a single time scale valid everywhere. Until 1905, this view seems to have been accepted universally.

The kind of relativity embodied in these views has been called "Newtonian relativity.*/ For purposes of comparison, it will be worth while to formulate it in mathematical terms.

. Let us represent position by means of Cartesian coordinates x, y, z

¹ Cf. MACH, E., "Science of Mechanics," 4th ed., p. 226, 1919.

² *Ibid.*, p. 222.

and let t denote the time. Then a set of four numbers representing values of x , y , z , and t specifies the position and the time at which an *event* of some sort occurs. The event might, for instance, consist in a projectile's passing a certain point in space at a certain time, or it might consist in an electron's leaving a filament or entering a Geiger counter. All physical phenomena, in so far as they involve positions and times, can be resolved into sequences of events.

In order that given values of our x , y , z may fix the location of an event, we must have some material reference body from which distances can be measured. Similarly, in order to define a time, we must have available some reference process, such as the rotation of the earth or the swinging of a pendulum, in order that times may be specified by stating the stage to which the reference process has advanced. These material means of fixing positions and times, together with the methods adopted for using them, are said to constitute a *space-time frame of reference*.

Now suppose we have two different frames of reference, each in uniform translatory motion relative to the other. One frame might

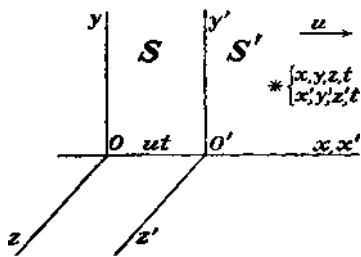


FIG. 25.—Two frames of reference in uniform relative translatory motion.

be attached to the earth, while the other is located on an airplane flying overhead. Let us call the two frames S and S' ; and let the velocity of S' relative to S be u . Let coordinates and times of any event as obtained when the frame S is used be denoted by x , y , z , t ; and let those obtained for the same event when S' is used be denoted by x' , y' , z' , t' . In order to make the relation between these variables as simple as possible, let us choose our axes so that the x - and x' -axes are both parallel to the velocity u and, in fact, slide along each other; and let the y - and y' -axes be parallel to y and y' , respectively (Fig. 25). Let us also agree to count time from the instant at which the two origins of coordinates, O and O' , momentarily coincide. Then the coordinates of O' , the origin of S' as measured in the first frame S , will be $x = ut$, $y = 0$, $z = 0$; and, if any event occurs at a position and time specified by certain values of x , y , z , t , then, according to Newton, its position and time as measured using S' will be represented by x' , y' , z' , t' , where

$$\begin{aligned}x' &= x - ut, & t' &= t, \\y' &= y, \\z' &= z.\end{aligned}$$

These equations may be called the *equations of transformation* for Newtonian relativity. They enable us to pass from coordinates and times of events referred to one frame of reference over to coordinates and times referred to another, when all relations are assumed to be Newtonians. The equation, $t' = t$, which may seem superfluous but is included with an eye to the future, expresses the fact that in the Newtonian theory there is supposed to be a single time scale valid for all frames of reference.

56. Relativity and the Propagation of Light.—With the adoption of the wave theory of light, a new element, unknown to Newton, was brought into the problem of motion. For, if light consists of "waves" in an ether, these waves should move with a definite velocity *relative to the ether*, and their velocity *relative to material bodies* should change when the motion of these bodies through the ether changes. Analogous statements made about waves in material mediums are certainly true. Water waves of a given length, for example, move with a certain speed over the water; as seen from a moving ship, they move past the ship faster when the ship is moving over the water against the waves than when the ship is moving in the same direction as the waves themselves. Now, in accordance with Huygens' principle, many optical phenomena depend in a very simple way upon the velocity

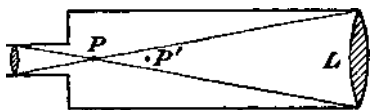


FIG. 26.—Effect of motion of a telescope through the ether.

of light relative to material objects.

It might be anticipated, therefore, that such phenomena would be influenced by a motion of the optical apparatus through the ether. The velocities that can be given to a material body in the laboratory are extremely small as compared with the velocity of light; but this is less true of the velocity of the earth in its orbital motion about the sun, which is one ten-thousandth of the velocity of light in free space. An influence of the earth's orbital motion upon optical phenomena might therefore be expected.

An interesting case to consider is the formation of images by the object lens of a telescope. Suppose a light wave from a star enters the telescope sketched in Fig. 26; when the telescope is at rest in the ether, let the wave come to a focus so as to form a star image on the cross hairs at P . Then, if the telescope, instead of being at rest, is moving toward the star, the wave might be expected to focus on the same point P in the ether as before. While the wave is passing from the lens to this point, however, the telescope will move forward, carrying the cross hair to some other point P' . The eyepiece would

therefore have to be drawn out farther in order to focus on the image of the star. Thus the focal length of the telescope would appear to the observer to be increased. Similarly, if the telescope were moving in the same direction as the light, its effective focal length would be shortened. If it were moving at right angles to its axis, on the other hand, there would be only an apparent lateral displacement of the star, constituting the familiar phenomenon of aberration.

Thus, as an astronomical telescope is pointed at stars in different directions, its apparent focal length might be expected to vary slightly because of the earth's orbital motion. But *no such effect has ever been observed*.

A simple explanation of this fact was suggested long ago by Fresnel. He suggested that moving transparent bodies, such as a lens, may partially drag the light waves along with them. In the case just considered, if the lens L were to drag the light with it in its motion (toward the right in Fig. 26), the part of the wave that goes through the center of the lens would spend a longer time in the lens, and hence would be retarded more than it would be if there were no motion. Consequently the wave would emerge from the lens more concave in shape, and would focus on a point nearer the lens. If the amount of the drag were just right, there might be no effect at all on the apparent focal length of the telescope, the star image falling on the cross hair however the telescope might be moving. Fresnel showed¹ that all effects on phenomena of refraction would be prevented if it were a law of optics that any moving transparent medium of refractive index n changes the velocity of light waves in such a way as to add vectorially to their velocity in the stationary medium, which is c/n , the fraction

of the velocity of the medium. Actually Fizeau showed experimentally in 1851, by an interference method, that a moving column of water *does* drag the light waves with it to the exact extent required by Fresnel's formula!

The cause of this effect on the light waves was supposed by Fresnel to lie in an actual partial dragging of the ether itself along with the moving medium, analogous to the dragging of the water by a moving fish net. When Lorentz worked out his electromagnetic theory, however, about 1895, on the assumption that the ether remains always

¹ For an excellent discussion of relativity see M. Born, "Einstein's Theory of Relativity," translated by H. L. Brose, Methuen & Co., Ltd., London, 1924.

at rest, he found that the theory led automatically to the Fresnel formula for the drag of the light waves by moving matter.

If the Fresnel drag of the light waves is assumed, it can be shown that there can never be any effect on ordinary optical phenomena that is of the *first order* in the velocity of the apparatus through the ether. Thus up to this point no difficulty arises from the assumption that light waves are propagated through an ether.

There might, however, be *second-order* effects. Since the square of the earth's orbital velocity is *only* a hundred-millionth of the square of the velocity of light, such effects will be difficult to detect. In seeking for a sufficiently delicate means of observation for this purpose, Michelson was led to invent his interferometer. With this instrument, in conjunction with Morley, he performed in 1887 a famous experiment in which a second-order effect could surely have been detected if it had been present.

57. The Michelson-Morley Experiment.—The interferometer arrangement that was used in this experiment is sketched in Fig. 27.

A beam of light from a lamp S falls upon a half-silvered glass plate P placed at 45° to the beam, which divides each wave into two parts. One partial wave, reflected from P , travels off sideways to a mirror M_1 , by which it is reflected back to P ; part of it is transmitted through P and enters the telescope T . The other part of the original wave, transmitted at once through P , travels ahead and is reflected back by a second mirror M_2 ; upon returning to P , it is partially reflected into the telescope on top of the first part of the wave, with which it forms an interference pattern.

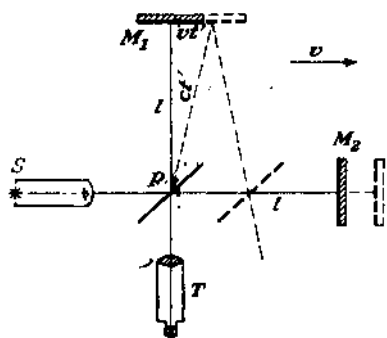


FIG. 27.—Diagram illustrating the Michelson-Morley experiment.

Let both mirrors be at the same distance from the plate P . Then, if the apparatus is at rest in the ether, the two waves take the same time to return to P and meet in phase both there and in the telescope. Suppose, however, that the apparatus is moving with speed v through the ether in the direction of the initial beam of light. Then, if the initial wave strikes the plate P when it has the position shown in the figure, the paths of the waves and the subsequent positions of reflection from mirrors and plate will be as shown by the dotted lines. The necessary change in the direction of the transverse beam is produced automatically, as an aberration effect, through the operation of

Huygens' principle. But now the times taken by the two waves on their journeys are no longer equal. The wave moving longitudinally toward M_2 has a velocity *relative to the apparatus* of $c - v$ on the outgoing trip, c being the speed of light through the ether, and $c + v$ on the return trip. Hence the time taken by this wave to get back to the plate is

$$t_2 = \frac{l}{c - v} + \frac{l}{c + v} = \frac{2cl}{c^2 - v^2},$$

l being the distance from the plate to either mirror. If we keep small terms only as far as v^2/c^2 , we can write, by the binomial theorem,

$$\frac{1}{c^2 - v^2} = (c^2 - v^2)^{-1} = c^{-2} + v^2 c^{-4}.$$

Hence

$$t_2 = 2l \left(\frac{1}{c} + \frac{v^2}{c^3} \right).$$

The wave moving transversely, on the other hand, travels along the hypotenuse of a right-angled triangle, one side of which is of length l , and therefore it has farther to go. Let it take a time t' to go from P to M_1 , traveling a distance ct' through the ether. Then in the same time the mirror M_1 advances a distance vt' . Hence

$$c^2 t'^2 = v^2 t'^2 + l^2, \quad t' = \frac{l}{(c^2 - v^2)^{1/2}}.$$

But, using the binomial theorem again,

$$(c^2 - v^2)^{-1/2} = c^{-1} + \frac{v^2 c^{-3}}{2},$$

as far as terms in v^2/c^2 . Hence this wave returns to P after a time

$$h = 2t' = 2l \left(\frac{1}{c} + \frac{1}{2} \frac{v^2}{c^3} \right).$$

Thus the two waves interfere in the telescope with a phase difference of $h - t_1 = lv^2/c^3$ seconds or $c(t_2 - t_1)/\lambda = lv^2/(c^2\lambda)$ wave lengths, λ being the wave length of the light. The fringe pattern is accordingly shifted by the motion through

$$\frac{v^2 l}{c^2 \lambda}$$

fringes.¹

¹ For the formation of the fringes in the interferometer see C. R. Mann, "Manual of Advanced Optics."

In performing the experiment, the whole apparatus, floated on mercury, was rotated repeatedly through 90° . Since the two light paths are caused to exchange roles by such a rotation, it should cause the fringe pattern to shift back and forth by twice the amount just calculated or by

$$N = 2 \frac{v^2 l}{c^2 \lambda}$$

fringe widths. By reflecting the beam back and forth several times, the effective length l was brought up to 11 meters. Even then, with a wave length of about 5.9×10^{-5} cm., if we insert for v the whole orbital velocity of the earth so that $v/c = 10^{-4}$, we find only

$$N = 2 \times 10^{-8} \times \frac{11 \times 10^3}{5.9 \times 10^{-5}} = 0.37,$$

or about one-third of a fringe. Michelson and Morley were sure, however, that they could detect a shift of a hundredth of a fringe. *Such shifts as were observed amounted only to a small fraction of the theoretical value, and were not consistent.* Thus the result of the experiment was negative; the expected effects of motion through the ether did not appear".

It might happen, to be sure, that at a given moment the earth by accident had no resultant component of velocity parallel to the surface of the earth; for upon its orbital motion there would probably be superposed a motion of the entire solar system through the ether. But, if this happened at a certain time, then 6 months later the earth's orbital velocity about the sun would be reversed, and its velocity through the ether should then be twice its orbital velocity. Michelson and Morley made observations not only at various times of day but also at different seasons of the year, always with the same negative result.¹

From the theoretical point of view, Michelson and Morley's failure to detect the anticipated motion of the earth through the ether was one of the most important experimental results ever obtained. This result was very hard to bring into harmony with current theories of light and matter. The theoretical calculation rests on a peculiarly simple foundation, for the only properties of light that are made use of are the constancy of its velocity in space and Huygens' principle. Neither of these principles can be given up if we are to retain the idea

¹ MICHELSON and MORLEY, *Silliman J.*, vol. 34, pp. 333, 427 (1887); MORLEY and MICHELSON, *Phil. Mag.*, vol. 24, p. 449 (1887). Cf. MILLER, *Rev. Modern Phys.*, vol. 5, p. 203 (1933).

of waves in a mechanical ether. Only three possible explanations of the negative result seem to offer themselves.

1. Perhaps the student will have wondered why we cannot simply assume that the earth drags the ether with it, much as a moving baseball carries along the air next to it. On this assumption there would never be any motion of the earth through the ether at all, and no difficulties could arise. One objection to this explanation is that the ether next to the earth would then be in motion relative to the ether farther away; and this relative motion between different parts of the ether would cause a deflection of the light rays coming from the stars, just as wind is observed to deflect sound waves. This deflection would alter the amount of the stellar aberration.¹ It has been found very difficult to devise a plausible type of motion for the ether which would give a value of the aberration agreeing with observation and yet at the same time preserve the negative result of the Michelson-Morley experiment. A second objection arises from the fact that, as has already been stated, experiments show that a transparent object of laboratory size does not drag the light waves with the full velocity of the moving matter, as it necessarily would do if it *completely* dragged the ether along with it; and the observed *partial* drag is fully accounted for by current electromagnetic theory, which is based on a stationary ether.

2. As an alternative, the negative result would at once be accounted for if we could assume that light projected from a moving source has added to its own natural velocity the velocity of the source, just as the velocity of a projectile fired from a moving ship is equal to the vector sum of its velocity of projection from the gun and the velocity of the ship. If this were true, the negative result obtained by Michelson and Morley would at once be explained, for the light from their lamp would have always the same constant velocity relative to the lamp and to the interferometer. Such an assumption, however, is in gross conflict with the wave theory of light. It is of the essence of waves that they have a definite velocity *relative to the medium which transmits them*, just as sound waves have a definite velocity relative to the air. Furthermore, there is strong *experimental evidence against* the assumption in question, which we have not space to describe.

3. The third possible explanation of Michelson and Morley's negative result, put forward independently by Fitzgerald and by Lorentz, is that motion through the ether might in some manner cause the material composing the interferometer to *shorten in a direction*

¹ Stellar aberration is the apparent change in direction of a star due to its orbital motion. See any text on astronomy.

parallel to the motion. It is easily seen that such a contraction in the ratio $\sqrt{1 - v^2/c^2}$ would serve to equalize the light paths and thus to prevent a displacement of the fringes. Lorentz endeavored to make the occurrence of such a contraction plausible on the basis of his electrical theory of matter, but he was never quite able to show that it *must* occur. This explanation came to be the favored one, but it was never felt to be very satisfactory because of its *ad hoc*¹ nature.

58. The New Relativity of Einstein.—From the situation just described one easily gains the impression that there exists something like a conspiracy in nature to prevent us from detecting motion through the ether. A similar situation can be shown to exist in the field of electromagnetism as well as in optics. A number of electrical or magnetic experiments can be invented which, at first sight, offer promise of revealing motion through the ether; but always there occurs some other effect which just cancels the effect sought. Sometimes it is actually the Lorentz-Fitzgerald contraction just described or the forces associated with this contraction by which the cancellation is effected.

In reflecting upon this extraordinary situation, Einstein arrived in 1905 at a radically new idea.² He proposed the view that, for some reason not yet known, motion through an ether filling empty space is a meaningless concept; only motion relative to material bodies has physical significance. He then set about investigating how this "assumption could be made to harmonize with the known laws of optics. Possibilities of conflict arise in any argument involving the velocity of light. Consider, for example, a frame of reference S' (say on the earth) moving relative to another frame S (say on the sun), and suppose that S' carries a source of light. Then light from this source must move with the same velocity relative to S as light from a source on S itself, since, as explained above, we cannot suppose that the velocity of light is influenced by motion of its source. But this light must also move with the same velocity *relative to S'* ; otherwise the laws of nature would not be the same on S' as on S , and no reason could be assigned for their being different. Yet it seems quite impossible that light should move with the same velocity relative to *each* of two frames that are *moving relatively to each other!*

Einstein accordingly put the laws of the propagation of light

¹ An *ad hoc* hypothesis is one that is proposed in order to explain a single fact as distinguished from a hypothesis that explains simultaneously several distinct facts. "*Ad hoc*" translated from the Latin means 'to this' or 'for this.'

² EINSTEIN, *Ann. d. Physik*, vol. 17, p. 891 (1905).

in the forefront of the discussion. He presented his new theory as based upon two propositions, which are known as the postulates of the *special* or *restricted theory of relativity*. They may be stated as follows:

1. *The laws of physical phenomena are the same when stated in terms of either of two inertial frames of reference (and can involve no reference to motion through an ether.)*
2. *The velocity of light is independent of the motion of its source.*

By an "inertial" frame is meant one in which the law of inertia is true, *i.e.*, relative to which a body that is free from external influence and at rest remains at rest. An inertial frame is thus what is usually called an "unaccelerated" one; and the relative motion of two inertial frames can only consist of a uniform translation. Thus, in the special theory of relativity, we compare statements of physical laws expressed alternatively in terms of two frames of reference which are in *uniform translation* relative to each other.

Of these two "postulates," the second is believed to represent a rather simple experimental fact, whereas the first is a generalization from a wide range of physical experience. There is no implication that the first postulate, which contains the new principle of relativity, is in any way self-evident; like the assumptions made in all physical theories, it is intended as a hypothesis to be tested by comparing deductions from it with experimental observation.

59. Simultaneity and Time Order.—Einstein found the key by which these two postulates could be brought into harmony in a modification of our ideas concerning time. He showed that it is necessary to give up the Newtonian conception of absolute time. Newton had undoubtedly supposed that it is always possible to say which of two events precedes the other, or that they occur simultaneously. But how can this be done if the two events occur in widely separated locations?

In practice, times at two different places are compared by reference to a clock at each place. It is necessary to set these clocks into synchronism; and in modern practice this is done by means of radio signals, which are a form of light signal. For precision, however, it is necessary to correct for the time required for light to travel from one place to the other; and making this correction requires a knowledge of the velocity of light *in one direction*. Ordinary measurements of the velocity of light furnish only its *average velocity in two opposite directions*. We could measure its velocity in one direction if we had

our clocks already synchronized, of course; but this thought leads into a logical circle.

As an alternative, we might think of returning to the old-fashioned practice of navigators; we might carry a chronometer from one place to the other and set both clocks by comparison with this chronometer. But how could we prove experimentally that the chronometer runs at a constant rate while moving in various directions? Every method that can be devised to prove this, in general, or, to measure the velocity of light in one direction, turns out to rest upon some fresh assumption that cannot be tested in advance. Finally, one might think of hurling a ball from one location to the other with indefinitely great speed, so that no correction for its time of flight would be necessary. This would, in fact, do the trick. But, if all masses increase with velocity as the mass of the electron is known to do, a ball could not possibly be projected with a speed exceeding that of light. From the standpoint of existing knowledge, it is entirely possible, and it is a consequence of Einstein's new theory of relativity, that no signal can be transmitted faster than a light signal.

It follows, then, that the time order of two events may in some cases be entirely indeterminate; the idea of a relative order in time is inapplicable to such a pair of events. This happens in all situations such that a light signal, leaving the location of one event at the instant of its occurrence, cannot reach the location of the second event before the occurrence of this event.

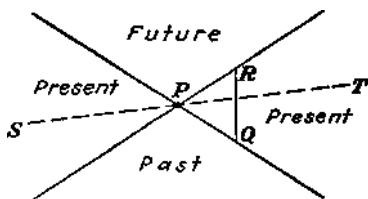


FIG- 28.--Diagram illustrating the classification of space-time relations.

The ambiguity in time just described can be regarded as a modification in the conception of the temporal *present*. This view can easily be illustrated diagrammatically if we consider only events that happen on the x -axis. Let us plot values of x and t as determined by an observer at an instant t_0 when he himself is at P (Fig. 28). Through P we can draw two straight lines representing the progress of light signals that pass the observer just when he is at P , one signal moving toward $+x$ and the other toward $-x$; i.e., any point on one of these lines represents the passage of one of these signals past a certain position on the z -axis at a certain time t . Then the area above these two lines represents events that lie in the *future* for the observer. These events have values of t so greatly exceeding U that the observer still has an opportunity to influence them causally. Below the two lines lies the observer's *past*, consisting of events which may have,

influenced him causally, and of which he may already have acquired knowledge at the instant t_0 by means of light signals or other means of communication.

There is also a *third* region, however, *between* the lines. This region constitutes the observer's *physical present*. He cannot influence events in this region—there would not be time to get any signal to the location of one of them before it occurred; nor, for the converse reason, can he have any knowledge of them at the time t_0 . *At each point of space these events belonging to the observer's present extend over a finite interval of time*, such as that represented by the line QR in the figure, the interval increasing with distance from P . For an observer in New York this interval amounts to $1/60$ of a second for events in San Francisco; within this $1/60$ sec. (the time required for light to make the double journey), he cannot pick out any unique instant in San Francisco which is simultaneous with the instant *to* in New York. Here we have an essential difference between Einstein's relativity and Newton's; for the latter reduced the physical present to a single line, such as ST . The observer himself might, of course, set up a system of space-time measurements, as we do on the earth, in which only the events lying along some such line as ST would be regarded as simultaneous with P . But, as we shall see presently, the choice of this line is largely arbitrary, depending upon the frame of reference that is used; it may be in fact any line through P that lies entirely within the regions labeled "present." Thus the simultaneity of events is not an absolute relationship but a relative one, depending upon the choice of a frame of reference.

60. The Lorentz Transformation.—If the postulates of Einstein's relativity are accepted, it becomes important to inquire whether or not the accepted laws of physics are all in harmony with the theory. Furthermore, on the basis of this theory, it may be possible to predict new phenomena, which can be looked for experimentally. In any case, like all physical theories, the theory of relativity must stand or fall according to whether or not the deductions that can be drawn from it agree with the experimental facts.

In order to make deductions from relativity, we must compare the descriptions of some phenomenon in terms of two inertial frames which are moving relatively to each other. It is important to know, first of all, how measurements of space and time compare. In discussing these, for the sake of vividness, we shall often speak of an *observer* who is supposed to make the measurements that are referred to a particular frame. This must not be understood to imply, however, that relativity has any closer connection than the rest of physics with human psychology.

The ambiguity in regard to times at different places, which was discussed in the last section, is commonly removed, in using any given frame of reference, by setting up the time scales in such a way that the velocity of light in vacuum measures the same in any one direction as it does in the opposite direction. The velocity of light in vacuum then becomes the same in all directions and therefore is a universal constant. Such an adjustment of the time scales can be supposed to be effected experimentally, as it actually is in common practice, by sending light signals (*i.e.*, radio signals) back and forth and correcting for time of propagation on the assumption of a constant velocity of light.

Let us use the same units with all frames of reference; to make sure of doing this, we may select such units of length and time that certain natural constants have the same values for both frames. For example, we may make the wave lengths of spectral lines and the velocity of light measure the same in all frames. When this has been done, the equations expressing all physical laws must take the same form, according to the first postulate of relativity.

Now let S and S' denote two inertial frames, of which S' has a uniform translational velocity u relative to S . Let axes be chosen again so that the x -axes and the xy planes in the two frames slide along each other (*cf.* Fig. 25); and, at the origins, let us count time from the instant at which these points momentarily coincide. Let t be the time at which some event happens at a point x, y, z , as measured by the observer on frame S , and let x', y', z', t' be the coordinates and time of the same event as measured by the observer on S' . Then it can be shown that the postulates of relativity require that, c being the velocity of light in vacuum,

$$\begin{aligned} x' &= \gamma(x - ut) & y' &= y & \gamma &= \frac{1}{\sqrt{1 - \frac{u^2}{c^2}}} & (52) \\ t' &= \gamma\left(t - \frac{ux}{c^2}\right) & z' &= z \end{aligned}$$

These equations, known as the "Lorentz transformation," were discovered by Lorentz in the course of his study of moving matter in an electromagnetic field; but he regarded one frame as at rest in the ether and attached an immediate physical meaning only to measurements made with this frame. The new principle embodied in the theory of relativity is that all inertial frames are to be treated on an equal footing. The equations can be derived from the requirement that the velocity of light shall be the same in both frames, together with considerations as to the isotropy of space; the deduction **will not** be

given here,¹ but a proof will be given in Sec. 62 that the Lorentz equations actually do preserve the constancy of the velocity of light.

By solving Eqs. (52) for x, y, z, t , we obtain the inverse transformation:

$$\begin{aligned}x &= \gamma(x' + ut') & y &= y' \\t &= \gamma\left(t' + \frac{ux'}{c^2}\right) & z &= z'.\end{aligned}\tag{53}$$

By means of these equations of transformation, positions and times measured in one frame of reference can at once be translated into positions and times of the same events as measured in the other frame. From the equations, we can easily confirm the statement, made previously, that, according to the new relativity, events which happen at the same place at different times, as viewed from one frame, may be seen from another frame to happen at different places as well. Similarly, a difference in spatial position with respect to one frame may correspond to a difference in both space and time with respect to another. Thus a space difference can be converted partly into a time difference, or vice versa, merely by changing the frame of reference that is being used. For this reason it has become customary to speak of space and time as aspects of a four-dimensional continuum known as "space time."

61. Contractions in Space and Time.—There are two famous cases in which the Lorentz transformation leads immediately to results of special interest.

Consider a body which, when at rest in S , has a length L_0 in the direction of the x -axis. Let it be set moving relative to S at such speed that it is at rest in S' . Its length as measured in S' will now be L_0 ; for its length is determined by certain natural laws and hence must have a certain fixed value in any frame in which the body is at rest. Let us see how the length now measures in S , relative to which the body is moving with speed u . Before doing this, we must consider what we mean by the *length* of a *moving* object. It seems reasonable to define the length as the distance between two points fixed in S which are occupied by the ends of the object simultaneously, i.e., at the same time t . If the coordinates of these points are x_1 and x_2 , the length is then

$$L = x_2 - x_1.$$

By the same definition, since the body is at rest in S' , its ends have fixed coordinates x'_1, x'_2 such that

$$L_0 = x'_2 - x'_1.$$

¹ Cf. SILBERSTEIN, L., "Theory of Relativity," 1924.

If we substitute in this last equation values of x'_2 and x_1 calculated from the first one of Eqs. (52) for a given value of t , we obtain

$$L_0 = \gamma(x_2 - x_1).$$

Therefore

$$L = \frac{1}{\gamma} L_0 = L_0 \sqrt{1 - \frac{u^2}{c^2}}. \quad (54)$$

From this equation we may draw two distinct conclusions. In the first place, any body measures shorter in terms of a frame relative to which it is moving with speed u than it does as measured in a frame relative to which it is at rest, the ratio of shortening being $\sqrt{1 - u^2/c^2}$. This is a relation between *measurements* referred to different frames.

In the second place, relative to a single frame, any physical body set into motion with speed u shortens in the direction of its motion, as was postulated by Fitzgerald and Lorentz, in the ratio $\sqrt{1 - u^2/c^2}$. In one sense, the contraction is perhaps not a "real" one, since, in a frame in which it is at rest the body measures the same as before; but, as far as effects on surrounding bodies are concerned, the contraction is as real as if it were due to a drop in temperature. For example, a row of similar crystals placed in contact, and then accelerated equally so as to preserve the spacing of their centers, would separate because of the contraction under discussion and would allow light to pass through between them. Perhaps Ave might say that we have here a sort of *kinematical perspective*, analogous in a way to the ordinary experience that an object appears to change in size as it recedes into the distance.

Unfortunately, these effects are too small to observe on the laboratory scale. It can be said, however, that the negative result of the Michelson-Morley experiment, when considered by an observer relative to whose frame of reference the apparatus is in motion, is due to the contraction in question.

There is a somewhat similar effect in *time*. Consider any good "clock," by which is meant a physical system containing a periodic motion or process which can be used to mark off equal intervals of time; to make it a "good" one, let the period of this process be always the same when measured in a frame in which the system is momentarily at rest. According to the first postulate a vibrating atom will constitute such a clock. Let the clock be at rest in s' , and let an interval equal to its period as observed in this frame be

$$t'_2 - t'_1 = T_0.$$

Then x' is constant at the clock; hence the corresponding interval and period in S will be, from one of Eqs. (53),

$$T = t_2 - t_1 = \gamma(t'_2 - t'_1) = \gamma T_0.$$

This result has again a double meaning. Its most interesting meaning is that the effects of one physical system on another relative to which the first is moving at velocity u are modified just as if all natural processes on the moving system were slowed down in the ratio $\sqrt{1 - u^2/c^2}$. Let ν_0 be the frequency of some natural process on the moving system, as measured in a frame of reference that is moving with this system and in which the system is, therefore, at rest. Then the frequency of this same process, as measured in another frame relative to which the system is moving with velocity u , is

$$\nu = \nu_0 \sqrt{1 - \frac{u^2}{c^2}}. \quad (55)$$

It follows, for example, that spectral lines, when observed with a spectroscope relative to which the atoms emitting the lines are in motion, should show a slight displacement toward the red, as compared with the lines from stationary atoms, this displacement being superposed upon the ordinary Doppler effect. This effect appears actually to have been observed by Ives, working with canal rays in hydrogen.¹

62. The Transformation of Velocities.—The Lorentz transformation leads to important formulas for the transformation of velocities from one frame of reference to another. These formulas are easily obtained by noting that, if any moving entity has a velocity v relative to S as measured by the observer who uses S , or v' relative to S' when measured by the S' observer, with corresponding Cartesian components v_x, v_y, v_z and v'_x, v'_y, v'_z , then

$$v_x = \frac{dx}{dt}, \quad v_y = \frac{dy}{dt}, \quad v_z = \frac{dz}{dt}; \quad v'_x = \frac{dx'}{dt'}, \quad v'_y = \frac{dy'}{dt'}, \quad v'_z = \frac{dz'}{dt'}.$$

On the other hand, from (52)

$$\begin{aligned} dx' &= \gamma(dx - u dt) & dy' &= dy \\ dt' &= \gamma\left(dt - \frac{u}{c^2} dx\right) & dz' &= dz. \end{aligned}$$

From these expressions, one finds by substitution that

$$\begin{aligned} v'_x &= \frac{v_x - u}{1 - uv_x/c^2}, & v'_y &= \frac{v_y}{\gamma(1 - uv_x/c^2)}, \\ & & v'_z &= \frac{v_z}{\gamma(1 - uv_x/c^2)}. \end{aligned} \quad (56a, 6, c)$$

¹ IVES and STILWELL, *J. O. S. A.*, vol. 28, 215 (1938).

As an example illustrating the mode of deduction,

$$v'_x = \frac{dx'}{dt'} = \frac{\gamma(dx - u dt)}{\gamma(dt - u dx/c^2)} = \frac{dx/dt - u}{1 - u dx/c^2 dt} = \frac{v_x - u}{1 - uv_x/c^2}.$$

By means of these equations we can finally verify that the Lorentz transformation makes the speed of light in vacuum always equal to c . For from (56a,b,c) we find, for the square of the velocity v' ,

$$v'^2 = v_x'^2 + v_y'^2 + v_z'^2 = \frac{1}{(1 - uv_x/c^2)^2} \left[(v_x - u)^2 + \left(1 - \frac{u^2}{c^2}\right) (v_y^2 + v_z^2) \right].$$

If now $v = c$, so that $v_y^2 + v_z^2 = c^2 - v_x^2$, the expression in brackets becomes

$$c^2 \left(1 - \frac{2uv_x}{c^2} + \frac{u^2 v_x^2}{c^4} \right) = c^2 \left(1 - \frac{uv_x^2}{c^2} \right),$$

so that $v'^2 = c^2$ and $v' = c$. In a similar way it can be shown that $v = c$ if $v' = c$.

If, on the other hand, $v < c$, then so is v' , and vice versa, as can readily be shown.

The equations just given for the transformation of velocities as measured in *different frames* should not be confused with the ordinary rules for the composition of two velocities measured in the *same frame*. The latter rules are, of course, still valid. To take a numerical example, let two electrons, ejected from a filament stationary in S , move off with equal speeds of magnitude $0.9c$, one going toward $-x$ and the other toward $+x$. Then their speed relative to each other, still measured in S , is $1.8c$, by the usual rule. This exceeds c . But, if we make $u = -0.9c$, so that the frame S' keeps up with the electron going toward $-x$, by (56a) the velocity of the second electron relative to the first, *measured now in S'* , is

$$\frac{[0.9c - (-0.9c)]}{(1 + 0.9^2 c^2/c^2)} = \frac{1.8c}{1.81},$$

which is a little less than c .

If in Eqs. (56a,b,c) we let $c \rightarrow \infty$, so that $\gamma \rightarrow 1$, the equations approach the form that is familiar in Newtonian kinematics:

$$v'_x = v_x - u, \quad v'_y = v_y, \quad v'_z = v_z.$$

At the same time the equations of the Lorentz transformation, Eqs. (52), pass over into the Newtonian ones as given in Sec. 55. Thus the Newtonian relations constitute an approximation that is valid as long as we are dealing with velocities much below that of light.

63. Relativistic Mechanics. The Variation of Mass.—The laws of mechanics as they left the hands of Newton are found upon examination to be not quite in harmony with the new theory of relativity. The necessary corrections to them were discovered originally in studying the motion of charged particles in electromagnetic fields. The proper relativistic corrections can also be inferred without difficulty from a study of ordinary mechanical phenomena.

For this purpose, we select for study a phenomenon the outcome of which can be inferred from considerations of symmetry, so that it cannot be in doubt. Let two exactly similar elastic balls approach each other along parallel lines and at equal speeds relative to a frame of reference S' , and let them collide with each other elastically. Then,

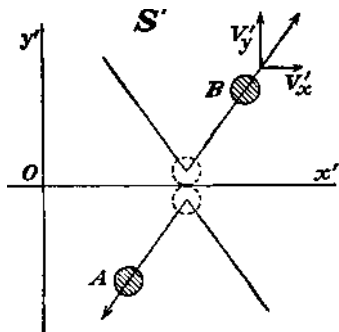


FIG. 29.—Diagram of a symmetrical collision of two similar balls.

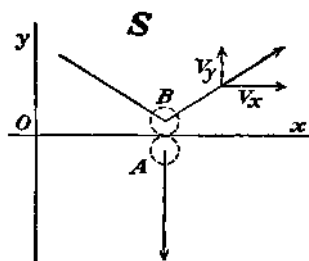


FIG. 30.—The same collision viewed from another frame of reference.

by the law of conservation of energy, their total kinetic energy must be the same after the impact as before; and, because of the symmetry, this requires that the balls rebound with their speeds unaltered and with their directions of motion deflected through the same angle (Fig. 29). In describing the collision, let Cartesian axes be drawn so that all lines of motion make equal angles with the x -axis and are parallel to the xy plane. Let the x' -component of velocity of the two balls relative to S' be $-v'_x$ and $+v'_x$, respectively, where $v_x > 0$.

Now let us describe this same collision in terms of another frame of reference, S , which is moving toward $-x$ relative to S' at velocity $-u$, where $u = v_x$, and let the axes in S be drawn in the usual manner. Then one ball, which we may call A , as viewed from frame S , moves up in a direction parallel to the y -axis and rebounds with a simple reversal of its motion, whereas the other, which we may call B , rebounds obliquely (Fig. 30).

In terms of either frame, the x -component of the velocity of both

balls remains constant; the y -component, on the other hand, becomes exactly reversed by the collision. As measured in S' , let the y -component of the velocity of A before the collision be v'_y ; then, because of the assumed symmetry, the corresponding component of B is $-v_y$. According to Newtonian mechanics, these components would also have the same values when measured in the S frame. The requirements of relativity, however, make this impossible. In terms of S , let the y -velocity of A before the collision be w_y , that of B , $-v_y$. Then, by applying (566) to the y -velocity first of A and then of B , we obtain the two equations

$$v'_y = \frac{w_y}{\gamma}, \quad v'_y = \frac{v_y}{\gamma(1 - uv_x/c^2)},$$

v_x being the x -component of the velocity of B referred to S . Dividing these two equations, we find that

$$\frac{v_y}{w_y} = 1 - \frac{uv_x}{c^2}. \quad (57)$$

Thus $v_y < w_y$, so that, as viewed from S , *ball B suffers a smaller change in its y -component of velocity than does A.*

This result makes it clear that the mechanics of Newton as it stands is not in harmony with the new theory of relativity. It needs some sort of correction. The form of this correction cannot be inferred from the theory of relativity alone; but a hint as to its probable form is furnished by the theory of the electromagnetic field. In Sec. 33 it was found possible to preserve the principle of the conservation of momentum, but the momentum of a *charged* body was not found to be exactly proportional to its velocity, as it is in Newtonian mechanics. The *inertial mass* of a charged body, denned as the ratio of its momentum to its velocity, was found to *increase with the velocity*, because of the electromagnetic momentum in the field. This suggests that in revising the laws of mechanics we should endeavor to preserve the principle of the conservation of momentum but should allow the inertial mass to vary.

Let us assume, therefore, that in the collision just described the two balls undergo equal and opposite changes in their y -components of *momentum*, denning momentum in the usual way as the product of inertial mass and velocity. Then, if m_A , m_B are the respective inertial masses of the two balls relative to frame S , we have $2m_A w_y = 2m_B v_y$; hence

$$\frac{m_B}{m_A} = \frac{w_y}{v_y} = \left(1 - \frac{uv_x}{c^2}\right)^{-1}$$

by (57). But $u = v_x$ and v_x represents the x -component of the velocity of B referred to S' . Hence by (56)

$$u = v'_x = \frac{v_x - u}{1 - uv_x/c^2}.$$

Multiplying this equation through by $v_x (1 - uv_x/c^2)$, we obtain

$$uv_x - \frac{u^2 v_x^2}{c^2} = v_x^2 - uv_x^2,$$

therefore

$$1 - \frac{v_x^2}{c^2} = 1 - 2\frac{uv_x}{c^2} + \frac{u^2 v_x^2}{c^4} = \left(1 - \frac{uv_x}{c^2}\right)^2.$$

Hence

$$\frac{m_B}{m_A} = \left(1 - \frac{v_x^2}{c^2}\right)^{-\frac{1}{2}}. \quad (58)$$

To see what this result implies, let $w_y \rightarrow 0$, so that also $v_y \rightarrow 0$. Then m_A approaches a limit m_0 , which is known as the *rest mass* or the mass of the ball for its initial motion out of a state of rest. The other ball B , on the other hand, moves in the limit parallel to x with speed $v = v_x$; and by (58) its mass m_B approaches a limiting value m given by

$$m = \frac{m_0}{\sqrt{1 - \frac{v^2}{c^2}}}. \quad (59)$$

In this equation, m_0 represents also the mass of B when it is at rest, the two balls being by hypothesis just alike.

Thus the theory of relativity requires that the inertial mass of a body shall vary with its velocity v in proportion to $(1 - v^2/c^2)^{-1/2}$. If this is assumed, it can be shown that the principle of the conservation of momentum can be preserved for all types of elastic collision. The mass of the electron has actually been found by experiment to vary with velocity in the manner required by the theory of relativity. Unfortunately the increase in mass is so small at velocities attainable in the laboratory that it has not been detected for any other body. Thus Newtonian mechanics represents an approximation to the correct mechanics which is close enough for all ordinary purposes.

The momentum of a body of mass m moving with speed v can then be written in either of two forms, thus:

$$p = mv = \frac{m_0 v}{(1 - v^2/c^2)^{1/2}}. \quad (60)$$

64. Force and Kinetic Energy.—All other mechanical terms and principles need now to be scrutinized in order to see whether or not they require changes.

The force on a body may be defined as the *rate of change of its momentum*. This definition is certainly valid in Newtonian mechanics. If the mass varies, however, and if we retain this definition of force, we are compelled to abandon the familiar $F = ma$; for in linear motion we have then

$$F = \frac{d}{dt}(mv) = m \frac{dv}{dt} + v \frac{dm}{dt} = ma + v \frac{dm}{dt}.$$

Because of this complication, it is usually best to attack problems in relativistic mechanics in terms of momentum rather than of acceleration.

The definition of work as *force times distance* and of *energy* as *stored work* can be retained without change. This procedure is in harmony with electromagnetic theory. The new expression for the *kinetic energy* T of a moving body can then be found in the usual way by calculating the work done in setting it into motion. This gives

$$T = \int_{v=0}^{v-v} F ds = \int_{v=0}^{v-v} \frac{d}{dt}(mv) \frac{ds}{dt} dt = \int v \frac{d}{dt}(mv) dt = \int v d(mv),$$

since $ds/dt = v$, the instantaneous velocity. Inserting the value of the mass m from Eq. (59), we have then

$$T = \int v d(mv) = \int v d \frac{m_0 v}{(1 - v^2/c^2)^{1/2}} = m_0 \int v \left[\frac{1}{(1 - v^2/c^2)^{1/2}} + \frac{v^2/c^2}{(1 - v^2/c^2)^{3/2}} \right] dv = m_0 \int_0^v \frac{v dv}{(1 - v^2/c^2)^{3/2}} = m_0 c^2 \left[\frac{1}{(1 - v^2/c^2)^{1/2}} \right]_0^v.$$

Thus we find, for the *kinetic energy* of a body of which the rest mass is m_0 , when moving with speed v ,

$$T = m_0 c^2 \left(\frac{1}{\sqrt{1 - v^2/c^2}} - 1 \right). \quad (61)$$

We can also expand in powers of v , obtaining

$$\left(1 - \frac{v^2}{c^2} \right)^{-1/2} = 1 + \frac{1}{2} \frac{v^2}{c^2} + \frac{3}{8} \frac{v^4}{c^4} \cdots,$$

so that

$$T = \frac{1}{2} m_0 v^2 + \frac{3}{8} m_0 \frac{v^4}{c^2} \cdots$$

Thus, if $v \ll c$, T reduces approximately to the ordinary value for the kinetic energy, $\frac{1}{2} m_0 v^2$. Under the same circumstances, the momen-

turn can also be written, as usual, m_0v . In general, Newtonian mechanics constitutes an approximate form of mechanical theory that is valid for any motion which is slow as compared with the speed of light.

65. A Relation between Mass and Energy.—Combining Eqs. (59) and (61), we can write

$$T = (m - m_0)c^2. \quad (62)$$

Thus the kinetic energy of a moving body equals c^2 times its gain in mass due to the motion. This relation suggests that we may think of the increase in energy as the actual cause of the increase in mass. As we have seen, the momentum present in an electromagnetic field can easily be accounted for in terms of a similar idea. It is then an attractive hypothesis to suppose that even the rest mass m_0 is due to the presence of an internal store of energy of amount m_0c^2 . This may be called the *rest energy* of the body.

The total energy of a moving body would then be $m_0c^2 + T$ or

$$W = mc^2 = \frac{m_0c^2}{\sqrt{1 - v^2/c^2}}; \quad (63)$$

and we can write for its inertial mass

$$m = \frac{W}{c^2},$$

and for its momentum

$$p = mv = \frac{Wv}{c^2}.$$

For some purposes, it is useful also to have a relation between W and p that does not involve v . From (63) and (65)

$$W^2 = m_0^2c^4 + p^2c^2. \quad (66)$$

The foregoing relations suggest that inertial mass may be a property of *energy* rather than of matter as such, each erg of energy possessing, or having associated with it, $1/c^2$ gram of mass. The law of the conservation of mass then becomes merely another aspect of the law of the conservation of energy.

These ideas have proved very useful in dealing with nuclear phenomena, which will be described in a subsequent chapter. It appears to be possible, for instance, for a γ -ray photon, impinging upon a nucleus, to be converted into an electron and positron. In such a case, the energy of the photon reappears in part as the kinetic energy of the particles but largely as their rest-mass energy $2m_0c^2$.

Some physicists prefer to speak of a conversion of energy into mass in the case just cited; they would say that neither energy by itself nor mass by itself is conserved, but only the sum total of mass times c^2 and energy. In this usage the term "mass" is restricted to "rest mass." The view stated above seems to be preferable, but the difference is scarcely more than a matter of words.

The association of mass and energy is not limited to kinetic energy or to the rest mass. It can be shown that relativity requires mass, or at least momentum, to be associated even with potential energy. This is easily shown in special cases; it may be worth while to give an example.

Suppose two equal masses, moving with equal and opposite velocities along the x -axis as seen by an S observer, collide with each other, a spring acting as a buffer between them; just as they come to rest, let a lock snap shut and hold them thereafter combined into a single mass. In this process, the initial kinetic energy of the masses is converted by the collision into potential energy of the spring. Let us view this collision from a second frame of reference, S' , that is moving with velocity u parallel to x . Then, if the velocities of the two bodies before the collision, as measured in S , are $v_1 = v_{x1} = v$, $v_2 = v_{x2} = -v$, the same velocities as measured in S' are, by Eq. (50a),

$$v'_1 = v'_{x1} = \frac{v - u}{1 - \frac{uv}{c^2}} = -u + \left(1 - \frac{u^2}{c^2}\right) \frac{v}{1 - \frac{uv}{c^2}},$$

$$v'_2 = v'_{x2} = \frac{-v - u}{1 + \frac{uv}{c^2}} = -u - \left(1 - \frac{u^2}{c^2}\right) \frac{v}{1 + \frac{uv}{c^2}}.$$

Therefore,

$$\frac{v'_1}{\left(1 - \frac{v'^2_1}{c^2}\right)^{1/2}} = -\frac{u}{\left(1 - \frac{v'^2_1}{c^2}\right)^{1/2}} + \left(1 - \frac{u^2}{c^2}\right) \frac{v}{\left(1 - \frac{u^2 + v^2}{c^2} + \frac{u^2 v^2}{c^4}\right)^{1/2}},$$

the last term being obtained in this form after inserting for v'_1 under the radical the first expression given for v'_1 above. Similarly,

$$\frac{v'_2}{\left(1 - \frac{v'^2_2}{c^2}\right)^{1/2}} = -\frac{u}{\left(1 - \frac{v'^2_2}{c^2}\right)^{1/2}} - \left(1 - \frac{u^2}{c^2}\right) \frac{v}{\left(1 - \frac{u^2 + v^2}{c^2} + \frac{u^2 v^2}{c^4}\right)^{1/2}}.$$

$$\frac{m_0 v'_1}{\sqrt{1 - v'^2_1/c^2}} + \frac{m_0 v'_2}{\sqrt{1 - v'^2_2/c^2}} = -m_0 u \left(\frac{1}{\sqrt{1 - v'^2_1/c^2}} + \frac{1}{\sqrt{1 - v'^2_2/c^2}} \right). \quad (67)$$

On the other hand, if we suppose the rest mass of the combined body after collision to be merely the sum of the rest masses of the separate bodies, or $2m_0$, the total momentum *after* collision is

$$\frac{-2m_0u}{\sqrt{1 - u^2/c^2}}$$

by Eq. (60), since the velocity of the combined body is then $-u$. This is *not equal* to the momentum *before* collision, as given by (67); this can be seen very easily in case v_1' and v_2' are both either greater or less than u .

Thus conservation of momentum fails if only the rest masses are taken into account. But now suppose we include in the mass of the combined body the mass that is to be associated, in the manner described above, with that part of the kinetic energy relative to S' which has been converted by the collision into potential energy of the spring. Then we have a total amount of mass which is proportional to the total energy; and, in consequence of the conservation of energy, this total mass is a constant. The total mass after the collision is, therefore, the same as it was before the collision or

$$\frac{m_0}{\sqrt{1 - v_1'^2/c^2}} + \frac{m_0}{\sqrt{1 - v_2'^2/c^2}}.$$

If we multiply this value of the total mass by the common velocity of the bodies after the collision, which is $-u$, we obtain for the total momentum *after* collision exactly the same expression as that given in Eq. (67) for the total momentum *before* collision. The principle of the conservation of momentum thus holds here if, and only if, we assume that the potential energy in the spring makes its proper contribution to the mass and momentum of the system.

66. Relativity and Electromagnetism.—Contrary to the situation in mechanics, a review of the laws of the electromagnetic field shows that these particular laws are in harmony with relativity just as they stand. This might have been expected in view of the fact that the theory of relativity actually developed out of experiments in that part of the field of electromagnetism which is called "optics."

The distinction between the *electric* field and the *magnetic* field becomes, however, from the new point of view, in part a *relative* one, depending upon the frame of reference that is being employed. This conclusion, at first sight surprising, can be reached by means of elementary considerations. For example, if there is a set of electric charges at rest relative to the S frame, they will produce, as determined **with the use of this frame, only an electrostatic field. But to an**

observer using an S' frame in motion relative to S , these same charges will constitute current elements and will be surrounded by a magnetic field as well. The general formulas for the transformation of electromagnetic fields from one frame of reference to another will not be deduced here, but they are easily stated.

As measured by the S observer, let the components of the electric field strength, in electrostatic units, be E_x, E_y, E_z , and let those of the magnetic field strength, in electromagnetic units, be H_x, H_y, H_z ; let the corresponding quantities as measured by the S' observer be indicated by the same symbols with the addition of primes. As usual, let S' have a velocity u relative to S , parallel to the x -axis, and let $\gamma = 1/\sqrt{1 - u^2/c^2}$, c being the velocity of light. Then the equations of transformation for the electromagnetic vectors are

$$\begin{aligned} E'_x &= E_x, & H'_x &= H_x, & \gamma &= \frac{1}{\sqrt{1 - u^2/c^2}}. \\ E'_y &= \gamma \left(E_y - \frac{u}{c} H_z \right), & H'_y &= \gamma \left(H_y + \frac{u}{c} E_z \right), \\ E'_z &= \gamma \left(E_z + \frac{u}{c} H_y \right), & H'_z &= \gamma \left(H_z - \frac{u}{c} E_y \right). \end{aligned}$$

Isolated charges measure the same in all frames; this must be so, since the charge on an electron must be a universal constant, and the

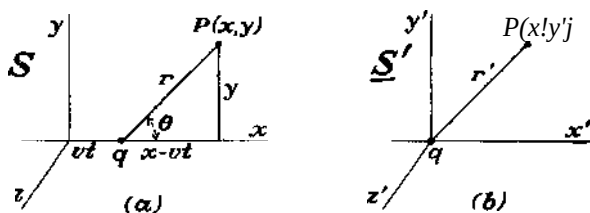


FIG. 31.—Diagram of a point charge moving relative to S , stationary relative to S'

number of electrons in a body, like the number of distinct objects in any given group, must be the same in all frames of reference. Furthermore, the usual rules for the *force* on a moving or stationary charge still hold; they furnish, in fact, the definition of what is meant by E and H .

The formulas just written may be used to ascertain the effect of uniform motion upon the field of a point charge and thereby to deduce certain formulas that were stated without proof in Chap. II.

Let a charge q be moving with speed v , and take the x -axis through the charge and in the direction of its motion; draw the xy plane through any point P in the field (Fig. 31). First let us use an S'

frame of reference in which the charge is at rest [Fig. 31(6)]. In this frame the field is purely electrostatic and is already known; if we take the origin at the charge and let x', y', z' be the coordinates of P referred to S' , the components of the field at P are, by Coulomb's law,

$$E'_x = \frac{qx}{r'^3}, \quad E'_y = \frac{qy}{r'^3}, \quad E'_z = 0, \quad r'^2 = x'^2 + y'^2. \quad (68a, b, c, d)$$

We now change to an S frame, relative to which S' has a velocity parallel to x of magnitude u . Relative to S , the charge has a velocity $v = u$. Then, from the second set of the above formulas for the transformation of fields, since $H'_x = H'_y = H'_z = 0$, we find, changing u to v ,

$$H_x = 0, \quad H_y = -\frac{v}{c}E_z, \quad H_z = \frac{v}{c}E_y. \quad (69a, b, c)$$

Also, inserting these values of H_y and H_z in the first set of the above formulas, with u replaced by v ,

$$E'_x = E_x, \quad E'_y = \gamma \left(1 - \frac{v^2}{c^2} \right) E_y = \frac{1}{\gamma} E_y,$$

Thus, at P , $E_z = 0$, since $E'_z = 0$; and hence, by (69c), $H_y = 0$. Equation (69c), $H_z = (v/c)E_y$, then confirms Eq. (19) in Chap. II.

Also, substituting in (70a, b, c) for E'_x , E'_y from Eqs. (68a, b) and likewise

$$x' = \gamma(x - vt), \quad y' = y$$

from the Lorentz transformation or Eqs. (52), we find

$$E_x = \frac{\gamma q}{r'^3}(x - vt), \quad E_y = \frac{\gamma qy}{r'^3}.$$

But vt is the x -coordinate of q , which is at the origin in S' . Hence, if r is the distance of P from q and if θ is the angle between the line qP and the x -axis (cf. Fig. 31a),

$$r^2 = (x - vt)^2 + y^2, \quad x - vt = r \cos \theta, \quad y = r \sin \theta.$$

Thus

$$\begin{aligned} r'^2 &= x'^2 + y'^2 = \gamma^2(x - vt)^2 + y^2 = r^2 + (\gamma^2 - 1)(x - vt)^2 \\ &= r^2[1 + (\gamma^2 - 1)\cos^2\theta], \end{aligned}$$

and the expressions just found for E_x and E_y can be written

$$E_x = \frac{\gamma \cos \theta}{[1 + (\gamma^2 - 1)\cos^2\theta]^{3/2}} \frac{q}{r^2}, \quad E_y = \frac{\gamma \sin \theta}{[1 + (\gamma^2 - 1)\cos^2\theta]^{3/2}} \frac{q}{r^2}. \quad (71a, b)$$

Thus $E_y/E_x = \tan \theta$, which shows that the electric field points radially outward from the instantaneous location of q . Furthermore,

$$E = (E_x^2 + E_y^2)^{1/2} = \frac{\gamma}{[1 + (\gamma^2 - 1) \cos^2 \theta]^{3/2}} \frac{q}{r^2}, \quad \mathbf{H} = \frac{v}{c} E \sin \theta$$

from [69c]. These last two equations were cited without proof in Chap. II.

In similar fashion, Eq. (23a) in Chap. II, for the radiation field emitted by any charge in linear accelerated motion, can be deduced from Eq. (22) in that chapter, which gives the field for a charge undergoing acceleration out of a state of rest.

Many other conclusions can be drawn by transforming fields from one frame to another; limitation of space allows only the mention of one other. The quantity $E^2 - IP$ is an "invariant," i.e., for any field, $E'^2 - I'^2 = E^2 - IP$, as the student can easily verify for himself. This invariance opens the possibility for plane waves in space to transform into plane waves when the frame of reference is changed; for, in such waves, $E = II$, $E^2 - IP = 0$. The direction of propagation and the frequency, however, are both altered by the transformation, in accordance with the phenomena of aberration and of the Doppler effect.

67. General Theory of Relativity.—In considering the bearing of the special theory of relativity upon physical laws, we have said nothing about one of the simplest of physical phenomena, viz., *gravitation*. After publishing the special theory, Einstein took up the problem of harmonizing the laws of gravitation with the requirements of that theory. Since we must suppose that no physical effect can be transmitted from one place to another with a velocity exceeding that of light, it may be assumed that gravitational effects are propagated with a finite velocity. What, then, is the law of this propagation?

At the same time, another idea was active in Einstein's mind. In the special theory, only *unaccelerated* frames of reference had been compared. Why this limitation? Could not the principle of relativity be generalized somehow so that *frames of all sorts* would stand on an equal footing?

In studying these questions, Einstein was impressed by the fact that gravitational acceleration is exactly the *same for all bodies*, however much they may differ in density or in other properties. In this respect, gravitational acceleration resembles the relative acceleration which appears when a frame of reference is itself subjected to acceleration. The latter effect is a matter of common experience. Every one knows how, when riding on an elevator, he seems momen-

tarily to become lighter whenever the elevator is accelerated downward and heavier when it is accelerated upward. This effect simulates closely an actual change in the force of gravity. By no mechanical experiment, indeed, can an apparent gravitational field thus produced be distinguished from a true field due to gravitational attraction.

Eventually¹ Einstein came to the conclusion that, in the neighborhood of any given point, there should be *no difference of any kind* between a gravitational field due to attracting matter and the "apparent" field due to acceleration of a frame of reference. This proposition he adopted as a *postulate* called the *principle of equivalence*. If the principle is accepted, it leads to the prediction of a number of physical effects hitherto unobserved.

Light, for example, had not commonly been supposed to be subject to gravitational action. But suppose the earth's gravitational field were abolished within a laboratory by allowing the whole laboratory to fall freely. Then, relative to the laboratory, there would be no gravitational attraction. A ball thrown out horizontally would travel in a straight line relative to the laboratory, not in a parabola. By the principle of equivalence, therefore, a ray of light projected horizontally would also appear to travel in a straight line; for conditions relative to the laboratory are the same as they would be out in space far from all attracting masses, and there is no doubt that in such locations rays of light are straight. Relative to the earth, however, the path of the ray of light would be slightly curved.

In view of the principle of equivalence, it can be said that we ourselves perceive a gravitational field on the earth only because we are using the wrong frame of reference. We ought to use a frame relative to which the earth is accelerated upward at the rate g ; using such a frame, we would find that the apparent gravitational field had completely disappeared. From this standpoint gravitational influence consists merely in determining what class of frames it is, relative to which there is no apparent field, and relative to which free bodies move in straight lines. Such frames are called, in a refined sense of the terms, *inertial* or *Galilean* frames of reference.

It does not follow, however, that the gravitational influence of one piece of matter on another is entirely illusory. For only a *uniform* gravitational field can be transformed away *in its entirety* by a proper choice of the frame of reference. In the neighborhood of a single point, any field can be transformed away; but, in general, the choice of frame that does this varies from point to point. For example, relative

¹ Cf. EINSTEIN, A., "Origins of the General Theory of Relativity," Jackson, Wylie and Co., Glasgow, 1933,

to a frame falling freely in New York there is no gravitational field in New York, but there is one of double strength in Australia.'

68. Einstein's Law of Gravitation.—There remains for consideration the problem as to the law according to which gravitating matter determines just which frames have the inertial property. The law must be such that its consequences agree with those derived from Newton's law of gravitation as a first approximation, since this law describes the motions of the solar system with high accuracy; and it must also be in harmony with the special theory of relativity. Einstein surmised that the law could probably be stated most simply in terms of a formulation that would permit the use, not only of any frame of reference in the ordinary sense, but of any sort of generalized coordinates. With the aid of the mathematician Grossmann, he found out how to write physical laws in a form that is valid *for any choice of space-time coordinates whatever*. The method involves the use of general tensor analysis, which is too complicated for any example of it to be given here. Suffice it to say that Einstein found that, among all possible guesses as to the correct law of gravitation, one stood out in contrast to all others as the simplest in mathematical form. Therefore he adopted this law as a tentative hypothesis¹ and then proceeded to look for predictions based on it which could be tested by experiment.

From the new law of gravitation thus obtained, Einstein deduced three novel effects that might be accessible to observation:

1. Rays of light passing close to a heavy body should be bent toward it. In the case of the sun, the deflection should be inversely proportional to the distance of closest approach of the ray to the sun's center and, for a ray just grazing the sun's surface, should amount to 1.75 seconds of arc. Stars seen near the sun, for example, during an eclipse, should appear to be displaced outward by this angular amount.

2. Physical processes in a region of low gravitational potential, when compared with similar processes at a point of high potential, should be found to take place more slowly. Consequently, atomic vibrations on the sun should appear to be slowed down, and spectral lines observed in the spectrum of sunlight should be shifted slightly toward the red, as compared with lines emitted or absorbed by the same elements on the earth.

3. The motion of the planets should be very slightly altered. In particular, the perihelion of the orbit of Mercury should be caused to precess about the sun at the rate of 43 seconds a century. The effect on the other planets would scarcely be detectable.

¹ EINSTEIN, *Ann. d. Physik*, vol. 49, p. 769 (1916).

It appears that all three of these effects actually occur. In the case of Mercury, calculation shows that perturbations by the other planets should cause an advance of the perihelion by more than 500 seconds a century, but the observed advance is about 43 seconds greater than can be accounted for in this manner. Thus Einstein's theory removed an annoying discrepancy in astronomical theory.

The new theory of space, time, and gravitation thus arrived at is known as the *general theory of relativity*. According to this theory, the spatial behavior of matter is not quite Euclidean. If a triangle of astronomical size near a heavy body like the sun were surveyed by means of rigid rods, with or without the help of light signals, the angles would not quite add up to 180° ; and so on. We have no space to pursue this fascinating subject further, however, nor to discuss its astronomical applications, which are connected with the question, not yet answered with certainty, as to the assumption that should be made concerning the outlying parts of space. Is space finite or infinite? Is it expanding, as the velocities of the nebulae are sometimes supposed to suggest?

For further study of such questions, the student must be referred to other books or to the literature.

CHAPTER V

THE ORIGIN OF THE QUANTUM THEORY

The quantum theory, first proposed by Planck in 1900, arose out of the inability of the classical physics to explain the experimentally observed distribution of energy in the spectrum of a black body. We have seen in Chap. I how, likewise, the older theory of radiation could not explain the experimentally observed facts of the photoelectric effect. In the present chapter, we shall first discuss certain observed phenomena of radiation which any theory must explain. We shall then show just how far the problem of black-body radiation can be solved by classical methods, and at precisely what point these classical methods failed and the introduction of the concept of quanta seemed to offer the only solution to the problem.

The phenomenon of black-body radiation is, however, of relatively minor importance in experimental physics. If the student does not wish to spend time upon it, he can without difficulty skip this chapter and proceed at once to the next.

69. Thermal Radiation.—It is a matter of common observation that bodies when heated emit radiant energy—or, more simply, radiation—the *quantity* and *quality* of which depend, for any given body, on the temperature of that body. Thus, the rate at which an incandescent lamp filament emits radiation increases rapidly with increase in the temperature of the filament; and the quality of the radiation, as observed visually, changes markedly as the temperature increases—the emitted light being "whiter" at higher temperatures. Radiation, the quantity and quality of which, emitted by any given body, depends *solely* on the temperature of that body, is called *thermal* radiation. It is a characteristic of thermal radiation that, when dispersed by a prism or other similar means, a *continuous* spectrum is formed. In order that thermal radiation may become visually observable, it is necessary that the temperature of the radiator should be 500 to 550°C. or above. Furthermore, thermal radiation is emitted, ordinarily, only by solids or liquids.

In the present chapter thermal radiation will be called simply "radiation," the adjective being omitted to save repetition. A complete account of the subject will not be attempted, the student being referred for this to the standard treatises on heat; only those features

will be discussed which bear on the fundamental problem that led to the quantum theory.

70. Some Fundamental Concepts and Definitions, (a) *Total Emissive Power.*—The rate at which a given body emits radiation depends upon the temperature of the body and on the nature of its surface. We may define the *total emissive power*, symbol E , of a body as the ^{if} total radiant energy emitted per unit time per unit area of surface of the radiating body?⁷ The total emissive power E increases rapidly with increasing temperature. Thus, the total emissive power of cast iron at a temperature of 1600°K.¹ is about 1.1×10^8 ergs per square centimeter per second; that of tungsten at 2450°K. is about 5×10^8 ergs cm.⁻² sec.⁻¹ (or 50 watts cm.⁻²).

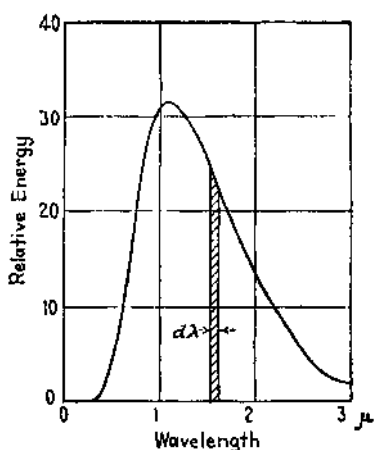


FIG. 32.—Distribution of energy in the spectrum of a tungsten lamp.

(b) *Monochromatic Emissive Power.*—If the radiation from a heated body, such as an incandescent lamp filament at a given temperature, be dispersed into a spectrum by a prism or other suitable device, it will be found, by means of a sensitive thermopile, that the energy in the spectrum is distributed among the various wave lengths in a regular manner, as is shown by the curve in Fig. 32, which shows the distribution of energy in the spectrum of tungsten at a temperature of 2450°K. We

may define the *monochromatic emissive power* e_λ at any given wave length λ by saying that the radiant energy emitted in the spectral range λ to $\lambda + d\lambda$, per unit area per unit time, is given by $e_\lambda d\lambda$. Thus, the ordinates of the curve in Fig. 32 are e_λ , and the area of the shaded strip is $e_\lambda d\lambda$. Obviously E and e_λ are related as follows:

$$E = \int_0^\infty e_\lambda d\lambda.$$

That is, the total emissive power is proportional to the area between the curve and the wave-length axis.

"Emissive power" is not to be confused with "emissivity," which we shall define later.

"K." refers to the absolute Kelvin scale of temperature, in centigrade degrees.

(c) *Intensity of Radiation from a Surface*.—Let dS (Fig. 33) be a small surface element of a radiating body. Describe about dS a hemisphere of radius p and let dB , located at P , be a small element of the surface of this hemisphere, the radius OP to this element making an angle θ with the radius ON which is normal to dS . The rate dQ/dt at which radiant energy is incident upon dB is easily seen to be proportional to (1) the area dB , (2) the area dS of the radiator, and (3) to $1/p^2$, because of the inverse-square law. It will depend also upon θ . The apparent area of dS as seen from P , or the area of the projection of dS upon a plane perpendicular to OP , is $dS \cos \theta$. It might be thought, therefore, that dQ/dt would be proportional to $\cos \theta$; and this well-known "cosine law" is, in fact, approximately obeyed by most surfaces. We can write, therefore,

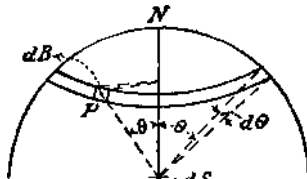


FIG. 33.

$$\frac{dQ}{dt} = i dS \cos \theta \frac{dB}{p^2} \quad (72)$$

where i is a quantity that is nearly or quite constant for a given radiator at a given temperature. Or, since dB/p^2 is the element of solid angle $d\omega$ which the area dB subtends at O , i may be defined by the equation

$$i = \frac{dQ/dt}{dS d\omega \cos \theta}.$$

The coefficient i is called the (total) *intensity of radiation* from dS in the direction OP . It represents the rate at which dS radiates energy in a given direction per unit solid angle and per unit of its own area as projected on a plane perpendicular to the given direction.¹

The definition just stated applies to the *total* radiation from the surface. As in the case of emissive power, we may designate by "monochromatic intensity of radiation," symbol i_λ , the intensity of radiation in a wave-length range $d\lambda$ at wave length λ .

When i is the same in all directions, there is a very simple relation between it and the total emissive power E . Referring to Fig. 33, we may choose as an element of area of the hemisphere a ring of width $p d\theta$ and of length $2\pi p \sin \theta$, the area of the ring being, $2\pi p^2 \sin \theta d\theta$

¹ The projection of a plane surface dS on a plane making an angle θ with dS has an area $dS \cos \theta$, since all lines on dS that are perpendicular to the intersection of the two planes are shortened by projection in the ratio $\cos \theta$, whereas lines **parallel** to the **line** of intersection are not shortened at all.

The flux of energy dQ/dt through this ring is given, as in Eq. (72), by

$$\frac{dQ}{dt} = i dS \cos \theta \frac{2\pi \rho \sin \theta \rho d\theta}{\rho^2} = 2\pi i dS \cos \theta \sin \theta d\theta.$$

If this equation is integrated from $\theta = 0$ to $\theta = \pi/2$, we shall have the total rate of flux of energy dQ/dt from dS . We have designated by E the total rate of flux of energy per unit area from a surface. Therefore,

$$E = \frac{dQ/dt}{dS} = 2\pi i \int_0^{\pi/2} \cos \theta \sin \theta d\theta = \pi i \sin^2 \theta \Big|_0^{\pi/2}.$$

Therefore

$$E = \pi i.$$

And similarly we have

$$e_\lambda = \pi i_\lambda.$$

(d) *Absorptivity*.—In general, radiation falling upon a surface is partly absorbed, partly reflected, and, unless the body be very thick or very opaque, partly transmitted. We shall define the *absorptivity* of a surface, symbol A , as the fraction of the radiant energy, incident on the surface, which is absorbed. Absorptivity is (1) a pure numeric, (2) for any actual body, less than unity, and (3) varies greatly with wave length of the incident radiation and, to a lesser extent, with the temperature of the absorber.

A surface whose absorptivity is unity for all wave lengths is called an "ideal" black surface. No such surface actually occurs in nature, but some bodies, such as black velvet or lampblack, reflect only a very small fraction of the incident radiation. It will be seen presently that in the theory of radiation a special interest attaches to the ideal black surface or body.

A very simple relation exists between the absorptivity of a surface and its total emissive power E . Suppose that two surfaces have total emissive powers E_1 and E_2 and absorptivities A_1 and A_2 , respectively. By considering the thermal equilibrium of such surfaces when present in an isothermal enclosure, as described in the next section, it can be shown that necessarily

$$\frac{E_2}{E_1} = \frac{A_2}{A_1}. \quad (74)$$

If, in particular, we make $A_1 = 1$, so that the first surface is black, E_x has obviously the maximum value that is possible at the given temperature; for A_1 in the last equation cannot exceed unity. Thus

no surface can emit more strongly than a black body. If E_0 is the total emissive power of a black body and E the total emissive power of any other body whose absorptivity is A , we find from the last equation that

$$E = AE_0. \quad (75)$$

These conclusions are known as "Kirchhoffs law" and have all been confirmed by experiment. The same relations can be shown to hold for each wave length separately.

(e) *Reflectivity*.—We may define the reflectivity of a surface, symbol R , as the fraction of the radiant energy incident upon the surface, which is reflected. R is a pure numeric.

(f) *Transmissivity*.—Likewise, we may define the transmissivity, symbol T , of a body as the fraction of the radiant energy, incident on the surface of the body, which is transmitted. We shall, in this chapter, confine our attention to cases where the body is so thick or so opaque as to transmit no energy, *i.e.*, to bodies for which $T = 0$. In these cases, since all the incident radiation is, then, either reflected or absorbed, we may write

$$A + R = 1.$$

(g) *The Density of Radiation*.—The radiant energy per unit volume in a stream of radiation is spoken of as the "energy density of radiation," symbol ψ . Thus, the solar constant¹ is about 1.94 calories per square centimeter per minute or 1.3×10^6 ergs per square centimeter per second. This is the amount of energy contained in a column of solar radiation 1 cm.² in cross section and 3×10^{10} cm. long, *i.e.*, in 3×10^{10} cm.³ The energy density of the sun's radiation in the neighborhood of the earth is, therefore,

$$\psi = \frac{1.3 \times 10^6}{3 \times 10^{10}} = 4.3 \times 10^{-5} \text{ erg per cubic centimeter.}$$

In the case just considered, the radiation is streaming in parallel directions. A different condition exists in the interior of a hollow, heated enclosure. Here radiation is streaming back and forth in all possible directions. The term "energy density" then refers to the total quantity of radiant energy in unit volume.

If a surface obeys the cosine law, so that the intensity i of the radiation emitted by it is the same in all directions, then a simple relation

¹ The solar constant is the amount of the sun's radiation received on unit area in unit time, the receiving area being perpendicular to the sun's rays and at a distance from the sun equal to the mean radius of the earth's orbit.

exists between the total emissive power E of the surface and the energy density in the emitted radiation at points near the surface. For, referring again to Fig. 33, let the radius p of the sphere be made very large. Then all rays drawn from points on dS to points on dB may be regarded as parallel. The radiant energy emitted in 1 second from dS , in the direction of the radius to dB , will be contained, therefore, in a cylinder whose length is the velocity of light c (vacuum being assumed). The cross-sectional area of this cylinder is $dS \cos \theta$ (this being the area of the projection of dS on a plane perpendicular to the radius); hence its volume is

$$dV = c dS \cos \theta.$$

The magnitude of the radiant energy is dQ/dt times 1 second or, by Eq. (72),

$$dE = i dS \cos \theta \frac{dB}{\rho^2}.$$

Dividing dE by dV , we have for the energy density due to this radiation anywhere inside the cylinder and hence, in particular, at points near the surface,

$$d\psi = \frac{i dB}{c \rho^2}.$$

Since i is assumed to be uniform and since $\int dB = 2\pi\rho^2$ (or half the area of the sphere), the total radiation density near the surface is thus

$$\psi = \int d\psi = \frac{i}{c \rho^2} \int dB = \frac{2\pi i}{c}.$$

Comparison of this equation with (73) above shows that

$$\psi = \frac{2E}{c}. \quad (70)$$

Furthermore, if we define as "monochromatic energy density," symbol ψ_λ , a quantity such that $\psi_\lambda d\lambda$ is the amount of energy present in unit volume in the form of radiation in a wave-length range $d\lambda$ at wave length λ , then clearly

$$\psi_\lambda = \frac{2e_\lambda}{c}. \quad (77)$$

Note.—There is considerable confusion in the terminology of the emission and absorption of radiation. Thus, for the quantity which we have herein denned as "total emissive" power," one finds used, variously, such terms as "radiating power," "emission," and "emissivity." And for the quantity which we have

called "absorptivity," one finds "absorbing power" and "absorptive power." The use of the word "power" in "total emissive power" is consistent with its use in mechanics. For "power" in the technical sense, in physics and engineering, means a "rate of doing work" or of delivering energy; and a radiating surface is emitting energy at a definite rate. But "power" used in connection with absorption, as in the term "absorbing power" is clearly inappropriate; for the quantity *defined* does not involve a *rate* of absorbing energy.

One must distinguish between "absorptivity," as herein defined, and "coefficient of absorption," which is an entirely different term, referring to the absorption of radiation in its passage *through* matter. Thus, if a beam of radiation of intensity I_0 is incident upon and passes through a slab of absorbing material, the thickness of which is d , the intensity I of the emerging beam is given by the equation $I = I_0 e^{-\kappa d}$, where "e" is the Napierian base of logarithms and κ the *coefficient of absorption* of the material.

It is essential, also, to distinguish between the terms "total emissive power" and "emissivity." If E is the total emissive power of a body at temperature T , and E_0 is the total emissive power of a black body at the same temperature, we have seen that $E = A E_0$, where A is the absorptivity of the substance. Now the emissivity of a surface is defined as "the ratio of its total emissive power to the total emissive power of a black body at the same temperature"; i. e., emissivity equals E/E_0 . Thus the emissivity of a substance is a pure numeric and is equal to its absorptivity. The same relations hold between monochromatic emissive power and the corresponding absorptivity.¹

71. The Isothermal Enclosure and Black-body Radiation.—A cavity whose walls and contents are all at a common temperature is called an *isothermal enclosure*. Such an enclosure is of special interest because the radiation field inside it possesses some remarkable properties.

It can be shown that *in an isothermal enclosure the stream of radiation in any given direction must be the same as in any other direction; it must be the same at every point inside the enclosure; and it must be the same in all enclosures at a given temperature, irrespective of the materials composing them. Furthermore, all of these statements hold for each spectral component of the radiation taken separately.*

The proof of these statements, which can be made quite rigorous, proceeds by showing that, if any one of them were not true, it would be possible to construct a device that would violate the second law of thermodynamics. For example, if the stream of radiation traveling west were greater than that traveling north, we could introduce two similar absorbers, one facing east and the other south. One of these absorbers would then become hotter than the other by absorbing radiant energy from the stronger stream. We could, therefore, operate a Carnot engine, using the two absorbers as source and sink, respec-

¹ So far as possible, the terminology defined in the "International Critical Tables," vol. I, pp. 34-42, has been adopted in this chapter.

tively, and so could convert heat continuously into work without leaving other changes in the system, in violation of the second law. Radiation of a particular wave length can be tested by using selective absorbers.

A simple relation exists between the stream of radiation in an isothermal enclosure and the radiation emitted by a black body. Suppose that such an enclosure contains a black surface. Then the radiation leaving this surface will consist entirely of radiation emitted by it, since such a surface reflects none of the radiation that falls upon it. Hence *the stream of radiation emitted by a black surface or body in any direction is the same as the stream of radiation traveling in any one direction in an isothermal enclosure at the same temperature.* Hereafter, we shall refer to both types of radiation indiscriminately as black-body radiation.

The *energy density* in the enclosure, however, will be twice that produced just in front of a black surface by radiation emitted by that surface alone. For the radiation emitted by the body is confined to a hemisphere of directions, whereas in the enclosure radiation is traveling in all directions. Hence, for the total energy density ψ , or the monochromatic energy density ψ_λ , in the enclosure, we can write, from Eq. (75) or (76),

$$\psi = 4 \frac{E}{c}, \quad \psi_\lambda = 4 \frac{e_\lambda}{c}, \quad (78a,6)$$

E being the total emissive power or e_λ the monochromatic emissive power of a black body at the same temperature.

To study black-body radiation, an electric furnace is employed, consisting of a long tube, preferably (but not necessarily) with blackened walls, heated by an electric current flowing in a wire wound around the tube. The temperature of the central part of the interior is measured with some sort of thermometer. A small hole is made through the wall, and the radiation issuing through it is observed. This radiation approximates very closely that which would be emitted by a black body at the temperature of the furnace.

Black-body radiation is a phenomenon of great interest from the theoretical standpoint, because its properties have a universal character, being independent of the properties of any particular material substance. Several questions press at once for an answer. How does the energy density in black-body radiation vary with the temperature? And what is the spectral distribution of the radiation? Furthermore, we wish to understand how this particular distribution is brought into existence by the atomic processes going on in matter.

Concerning the first two questions, it was found possible, during the last century, to obtain further information from thermodynamics *without making any assumption as to the atomic process*. The method consisted in considering the effect of expanding or contracting an isothermal enclosure and taking account of the work done on the walls by the radiation in consequence of *radiation pressure*. In order to follow the argument, we must ascertain the relation between the pressure and the energy density in uniformly distributed radiation.

72. Pressure Due to Isotropic Radiation.—Suppose a stream of radiation in a vacuum falls normally on the surface of a body. Then, if w is the mean energy density in the oncoming waves, they carry also w/c units of momentum per unit volume [Eq. (16) in Sec. 33]. Thus the waves bring up to each unit area of the surface, along with cw ergs of energy [Eq. (12) in Sec. 32], w units of momentum per second, the momentum as a vector being directed normally toward the surface. If the waves are absorbed by the surface, it receives this momentum and experiences, therefore, a pressure equal to w .¹

Suppose, next, that the radiation is incident at an angle θ . Then the energy that crosses a unit area drawn perpendicular to the rays (PQ in Fig. 34) is received by a larger area of magnitude $1/\cos \theta$ on the surface (PR in Fig. 34). Furthermore, the component of the momentum normal to the surface is less than in the case of normal incidence in proportion to $\cos \theta$. Thus the momentum in the direction of the normal that is delivered to unit area of the surface per second is decreased by the obliqueness of incidence in the ratio $\cos^2 \theta$, and the resulting pressure, if the radiation is entirely absorbed,

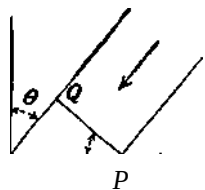


FIG. 34.

$$p = w \cos^2 \theta, \quad (79)$$

w being, as before, the energy density in the incident radiation. The same expression holds for the pressure caused by the *emission* of a beam at an angle θ , or for the additional pressure caused by the occurrence of a reflected beam. If an incident beam in which the energy density is w is specularly reflected from a surface at the angle of incidence θ , the total pressure on the surface is $2w \cos^2 \theta$.

Finally, let radiation be streaming toward a surface and also away from it with equal intensities in all directions, as in an isothermal enclosure. Such a distribution of radiation is equivalent to a large

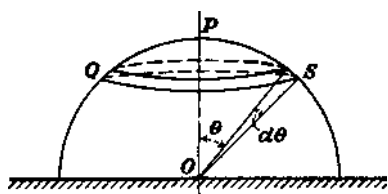
¹ Pressure and energy density are easily seen to be physical quantities of the same dimensions.

number of beams of plane waves, all of equal intensity, with their directions of propagation distributed equally in direction. Let there be N beams in all, and let the energy density due to any one of them be w . Then the total energy density just in front of the surface and the pressure p on it are

$$\psi = Nw, \quad p = \Sigma w \cos^2 \theta = w \Sigma \cos^2 \theta, \quad (80a,6)$$

by (79), $\Sigma \cos^2 \theta$ denoting the sum of the values of $\cos^2 \theta$ for all of the beams.

To find this latter sum, imagine lines drawn outward from a point O on the surface to represent the various directions of the beams, whether moving toward the surface or away from it, and then about O as center draw a hemispherical surface of unit radius with its base on the surface (*cf.* Fig. 35, where only two of the lines are shown). From



the hemisphere cut out a ring-shaped element of area, QS , by means of two cones of semiangle θ and $\theta + d\theta$, drawn from O as apex and with the normal OP as axis. The edge of this element is a circle of perimeter $2\pi \sin \theta$, and its width is $d\theta$, hence its area is $2\pi \sin \theta d\theta$;

whereas the area of the whole hemisphere is 2π . Now the lines of approach of the TV beams of radiation, if drawn through O , will cut the hemisphere in points equally distributed over its surface. Hence, if we let dN denote the number of these lines that pass through the ring-shaped element, dN will be to N in the ratio of the area of the ring to the area of the hemisphere, whence¹

$$\frac{dN}{N} = \frac{2\pi \sin \theta d\theta}{2\pi} = \sin \theta d\theta.$$

$$\sum_{\pi/2} \cos^2 \theta = \int \cos^2 \theta dN = N \int_0^{\pi/2} \cos^2 \theta \sin \theta d\theta = \frac{1}{3} N. \quad (81)$$

$$p = \frac{1}{3} \psi$$

$$p = \frac{1}{3} \psi.$$

have been reflected and are now leaving the mirror at an angle of reflection θ' , a distance λ' apart. It is obvious from the figure that $\lambda = BE \sin \theta$, $\lambda' = BE \sin \theta'$ whence

$$\frac{\lambda'}{\lambda} = \frac{\sin \theta'}{\sin \theta} \quad (88)$$

If the mirror is stationary, we have $\theta = \theta'$, hence $\lambda = \lambda'$. Otherwise, a second relation is necessary to fix λ' and θ' . To obtain this, consider *two successive positions of the same wave*. In Fig. 38(6), the part AB of a wave is just beginning to fall on the mirror whose instantaneous position is MM . The same portion of the wave at a later instant, after it has been reflected at the angle θ' , is shown by $A'B'$, the point A being now in contact at A' with the mirror, which is in the new position $M'M'$. While A traveled along the ray AA' , B traversed BB' , hence

$$AA' = BB'.$$

The angles BAA' and $BB'A'$, being angles between ray and wave, are right angles. It follows, therefore, by similar triangles, that angle

$$B'A'B = ABA'$$

or

$$6' - \epsilon = 6 + e, \quad (89)$$

where e is the angle QBA' . Furthermore, while A went from A to A' with the speed c of light, the mirror moved from MM to $M'M'$ at the speed u . Hence, if $A'Q$ is a perpendicular dropped from A' onto MM ,

$$\frac{A'Q}{AA'} = \frac{u}{c}$$

or

$$\frac{BA' \sin \epsilon}{\sin(\theta' + \theta)} = \frac{\sin \epsilon}{\sin(\theta' + \theta)} = \frac{u}{c}.$$

Substituting here $\epsilon = \frac{1}{2}(\theta' - \theta)$, from (89), we have

$$\frac{\sin \frac{1}{2}(\theta' - \theta)}{\sin \frac{1}{2}(\theta' + \theta)} = \frac{u}{c}. \quad (90)$$

From this one easily obtains

$$\tan \frac{1}{2}\theta' = \frac{c+u}{c-u} \tan \frac{1}{2}\theta \quad (91)$$

as the law of reflection from a moving mirror. If $u = 0$, it reduces to the familiar law $\theta' = \theta$.

The change of wave length implied by Eq. (88) represents a Doppler effect due to reflection from a moving mirror. It is this change which is needed for the further study of thermal radiation. We shall be dealing only with infinitesimal velocities, however, so that θ' and θ are nearly equal. For this case we can put $\theta = \theta'$ in the denominator in Eq. (90), and we can replace the sine by the angle in the numerator, obtaining

$$\frac{1}{2}(\theta' - \theta) = \frac{u}{c} \sin \theta;$$

and, similarly, subtracting 1 from both sides in (88), we have

$$\frac{\lambda' - \lambda}{\lambda} = \frac{\sin \theta' - \sin \theta}{\sin \theta} = \frac{2 \cos \frac{1}{2}(\theta' + \theta) \sin \frac{1}{2}(\theta' - \theta)}{\sin \theta} = \frac{\theta' - \theta}{\sin \theta} \cos \theta,$$

valid in the limit as $\theta' \rightarrow \theta$. From these two equations we obtain

$$\frac{\lambda' - \lambda}{\lambda} = 2 \frac{u}{c} \cos \theta;$$

whence, writing $\Delta\lambda = \lambda' - \lambda$, we have, for the increase in wave length upon reflection from a mirror receding with indefinitely small velocity v ,

$$\Delta\lambda = 2 \frac{u}{c} \lambda \cos \theta. \quad (92)$$

76. Effect of an Adiabatic Expansion upon Black-body Radiation.—

Returning now to the sequence of operations described in Sec. 73, let us consider the effect of the adiabatic process (*b*) upon the spectral distribution of the radiation. In this process, black-body radiation initially at temperature T_1 imprisoned in a cylinder with perfectly reflecting walls is slowly expanded from an initial energy density ψ_1 to a new energy density ψ_2 . Let the former restriction to a small expansion be dropped. The change in direction of the rays that is produced by the moving piston, according to Eq. (91) above, will tend to make the radiation no longer isotropic. We can obviate this inconvenient effect, however, by letting part of the walls of the cylinder reflect perfectly *but diffusely*. A surface of magnesium oxide does this very well. Then, if the expansion is made very slowly, because all rays (except a negligible few) strike the diffusing surface repeatedly, the radiation will be kept effectively isotropic; and the pressure on the piston, according to Eq. (82) above, will be at all times equal to $1/3\psi$.

Let the cylinder have a cross section A and a (variable) length Z . Then, if ψ is the energy density at any moment, when the piston moves outward a distance dl , work $p dV = 1/3\psi A dl$ is done on it by the force

due to radiation pressure. This work is done at the expense of the enclosed energy, the total amount of which is $IA\psi$. Hence

$$\frac{1}{3}\psi A dl = -d(lA\psi) = -A\psi dl - Ald\psi,$$

$$\frac{d\psi}{\psi} = -\frac{4dl}{3l},$$

and, after integration,

$$\log \psi = -\log l^{4/3} + \text{const.}, \quad \psi = Cl^{-4/3}, \quad (93)$$

C denoting a constant.

Now it can be shown by thermodynamical reasoning that an expansion of the type considered here cannot destroy the black-body property of the radiation. For, at a certain instant, suppose that the expansion has reduced the total energy density to ψ_2 . Let T_2 be the temperature of an enclosure in which the density has this same value ψ_2 . Suppose that in the cylinder there were more radiation per unit volume of wave lengths near some value λ' than at the same wave lengths in the enclosure, and less radiation near some other wave length X . It would then be possible to cause a little radiation to pass from the cylinder into a second enclosure at a temperature T_2' slightly above T_2 , by covering the opening in the base of the cylinder with a plate transmitting wave lengths near λ' but reflecting all others¹ and putting the cylinder into communication with the second enclosure through this opening. In a similar way, enough radiation near λ' could be passed *into* the cylinder from an enclosure at a slightly *lower* temperature T_2'' to restore the total energy to ψ_2 . Then the radiation could be compressed back to $1/3$, the changes in ψ and l and the amount of the work done being just the reverse of these quantities during the expansion. Finally, putting the cylinder again into communication with the enclosure at T_2 , we could allow the spectral distribution to be restored to that proper to a black body at T_2 but without any net transfer of energy between cylinder and enclosure, since the total energy density has already been restored to that corresponding to T_2 . Thus we should have performed a cyclic operation the *only* effect of which is to transfer heat energy from an enclosure at T_2'' to one at a higher temperature T_2' . But this is inconsistent with the second law of thermodynamics.

Thus black radiation must remain black during any slow adiabatic expansion. Its density, and therefore its temperature, however, decrease. On the other hand, we can determine from the Doppler effect how the spectral distribution of the radiation changes. In this

¹ Selective reflection and transmission are well-known optical phenomena.

way, we may arrive at a knowledge of the effect of changes of temperature upon the spectral distribution of black-body radiation.

Suppose, first, that the walls of the cylinder are perfectly reflecting. Then any ray preserves its angle of inclination to the axis of the cylinder in spite of successive reflections (c/. Fig. 39) and has, therefore, a constant component of velocity $c \cos \theta$ perpendicular to the piston. Thus the ray strikes the piston $ccos\theta/2Z$ times a second, l being the length of the enclosed space; and each time, according to (92), the wave length is increased $2u \cos \theta / c$, u being the speed of the piston. The wave length thus increases at an average rate of

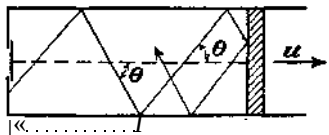


FIG. 39.

$$\frac{d\lambda}{dt} = \frac{c \cos \theta}{2l} \frac{2u \lambda \cos \theta}{c} = \frac{u \lambda}{l} \cos^2 \theta \text{ cm. per sec.}$$

This expression for $d\lambda/dt$ varies with θ . If, however, as assumed above, *diffuse* reflection occurs over part of the walls, the waves take turns moving in different directions, and hence they all experience the same *average* change in length. To simplify the calculation of this average change, imagine that each wave moves so as to be perpendicular to one of a very large number N of lines equally spaced in direction. Then, since the radiation remains isotropic, during a second of time any wave must spend $1/N$ second moving in each of the N directions. While moving in one of these directions, such that the wave makes an angle θ with the axis, its wave length increases by the amount

$$\frac{1}{N} \frac{d\lambda}{dt} = \frac{u \lambda}{l N} \cos^2 \theta.$$

The total change in λ in a second is thus

$$\sum \frac{u \lambda}{l N} \cos^2 \theta = \frac{u \lambda}{l N} \sum \cos^2 \theta,$$

Σ indicating the sum for all N directions. But $2 \cos^2 \theta$ was evaluated above and found to equal $4\pi/3$ [Eq. (81) in Sec. 72]. Hence, all waves of length λ increase during the expansion at the average rate

$$\frac{d\lambda}{dt} = \frac{u \lambda}{3l}.$$

Since $u = dl/dt$, this equation leads to the equations, for a given wave,

$$\frac{d\lambda}{\lambda} = \frac{1}{3} \frac{dl}{l}, \quad \log \lambda = \log l^{1/3} + \text{const.}, \\ \lambda \propto l^{1/3}$$

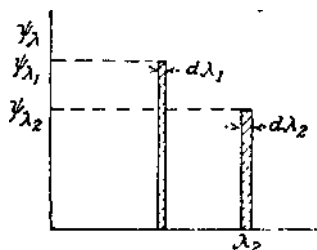
But, by (93), $\lambda \propto \psi^{-1/4}$; and, also, by the Stefan-Boltzmann law, $\psi \propto T^4$. Hence $\lambda \propto \psi^{-1/4} \propto T^{-1}$. Thus we reach the conclusion that the wave length of any given spectral component of black-body radiation changes during a slow adiabatic expansion in such a way that

$$\lambda \propto \frac{1}{T}.$$

The Wien Displacement Law.—The result just obtained leads to an important conclusion concerning *the manner in which the spectral distribution of black-body radiation depends upon the temperature*. Let $\psi_\lambda d\lambda$ denote the part of the energy in unit volume that is due to waves of length between λ and $\lambda + d\lambda$, the total energy density being thus $\int_0^\infty \psi_\lambda d\lambda$. Let us fix our attention on a particular spectral range $d\lambda$ at λ_1 containing energy $\psi_{\lambda_1} d\lambda_1$ in the black-body radiation at temperature T_1 . Then an adiabatic expansion which lowers the temperature to some new value T_2 changes the wave lengths of this part of the radiation in the ratio T_1/T_2 , λ_1 becoming λ_2 and $\lambda_1 + d\lambda_1$, $\lambda_2 + d\lambda_2$, where

$$\frac{\lambda_2}{\lambda_1} = \frac{T_1}{T_2}, \quad \frac{\lambda_2 + d\lambda_2}{\lambda_1 + d\lambda_1} = \frac{T_1}{T_2}, \quad \therefore \frac{d\lambda_2}{d\lambda_1} = \frac{T_1}{T_2}. \quad (95a, 6, c)$$

(Cf. Fig. 40.) The energy originally in $d\lambda_1$ is also decreased, and in the same ratio as is the total energy; for, in the argument that led up to Eq. (93), we might have started with only the radiation in $d\lambda_1$ present in the cylinder, and then Eq. (93) would have been obtained for this part of the radiation alone. Hence, ψ_{λ_2} being the new monochromatic density of energy,



$$\frac{\psi_{\lambda_2} d\lambda_2}{\psi_{\lambda_1} d\lambda_1} = \frac{\psi_2}{\psi_1} = \frac{T_2^4}{T_1^4},$$

by the Stefan-Boltzmann law. Substituting here the value of $d\lambda_2/d\lambda_1$ from (95c), we obtain

$$\frac{\psi_{\lambda_2}}{\psi_{\lambda_1}} = \frac{T_2^5}{T_1^5}. \quad (96)$$

Since ψ_λ is proportional to the monochromatic emissive power of a black body at the same temperature, we can also conclude that

$$\frac{e_{\lambda_2}}{e_{\lambda_1}} = \frac{T_2^5}{T_1^5}. \quad (97)$$

Wave lengths λ_2, λ_1 , such that $\lambda_2/\lambda_1 = T_1/T_2$ may be called *corresponding wave lengths* for black radiation at these temperatures. Equations (96) and (97) then state that the *monochromatic energy density* ψ_λ in an isothermal enclosure and the *monochromatic emissive power* of a black body when taken at corresponding wave lengths are both directly proportional to the fifth power of the absolute temperature. This important conclusion is known as the *Wien displacement law*.

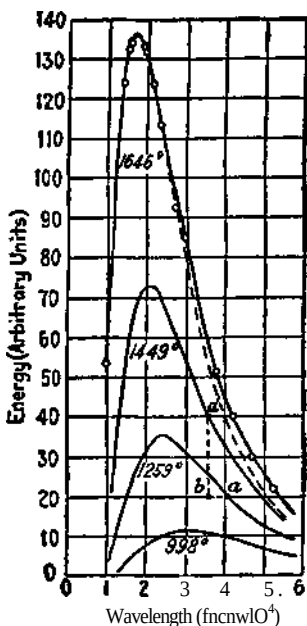


FIG. 41.—Distribution of energy in the spectrum of a black body at various temperatures.

It follows that, if λ_n is the wave length at which e_λ and ψ_λ have their *maximum* value,

$$\lambda_n T = A = \text{const.}$$

for all temperatures—a special case of the displacement law that is often cited. The value of A is discussed in Sec. 85.

77. The Formula for Black-body Radiation.—The relationship represented by Fig. 42 can also be expressed mathematically. It is obvious that e_λ/T^5 is some function of the product λT we may write

$$\frac{e_\lambda}{T^5} = f(\lambda T).$$

¹ D. phys. Ges. "Verhandlungen, vol. 1, pp. 23 and 215 (1899); vol. 2, p. 163 (1900).

Or, since then

$$e_{\lambda} = T^5 f(\lambda T) = (\lambda T)^5 \frac{f(\lambda T)}{\lambda^5},$$

we can write

$$e_{\lambda} = \frac{1}{\lambda^5} F(\lambda T), \quad (98)$$

where $F(\lambda T) = (\lambda T)^5 f(\lambda T)$.

Thus, by reasoning based on thermodynamics, the problem of black-body radiation is reduced to the determination of the single unknown

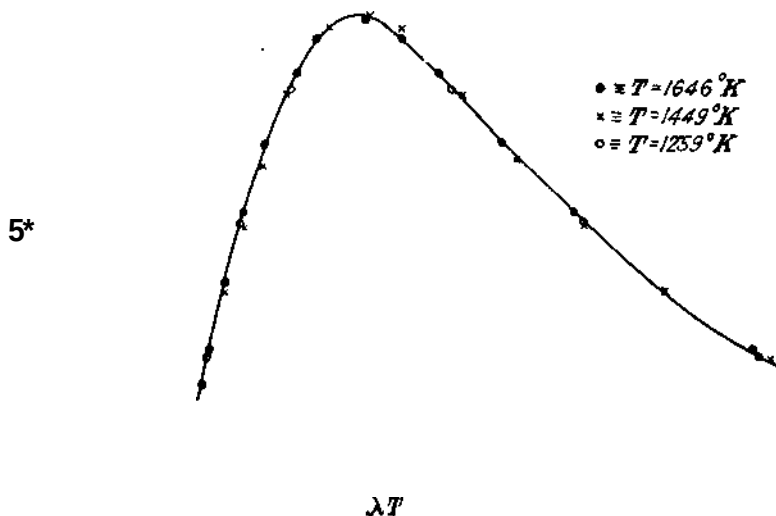


FIG. 42.—Experimental verification of the black-body displacement law.

function $F(\lambda T)$. In order to determine this function, we must resort to other methods. All attempts to find the function $F(\lambda T)$ on the basis of classical theory, however, failed. Equation (98) represents the limit of success of classical theory in dealing with the problem of black-body radiation.

Two of the formulas that were proposed on the classical basis, although incorrect, deserve some consideration. Wien proposed a formula that was derived from special assumptions concerning the process of emission and absorption, viz.

$$\psi_{\lambda} = \frac{4}{c} e_{\lambda} = c_1 \lambda^{-5} e^{-\frac{c_2}{\lambda T}} \quad (99)$$

where c_1 and C_2 were certain constants which remained to be found. Rayleigh made a suggestion based on very general reasoning,¹ and on the basis of this suggestion Jeans arrived at formulas for ψ and E that were free from all unknown constants. The Rayleigh-Jeans formula was found to fit the experimental curve at very long wave lengths; whereas, by assigning proper values to C_1 and C_2 , Wien's formula could be made to fit at wave lengths shorter than the point of maximum emission. Otherwise, both formulas failed (cf. Fig. 48 in Sec. 85). The method of approach employed by Rayleigh and Jeans is full of interest, and it also furnishes an excellent background for the consideration of Planck's introduction of quanta, by means of which he succeeded in finding the (apparently) *correct* black-body formula; hence the reasoning of Rayleigh and Jeans will now be described. It is based upon the classical law of the equipartition of energy; therefore this will be discussed first.

78. Degrees of Freedom.—A gas molecule within an enclosure is *free to move* in a variety of ways. It may, in general, have a motion of translation, a motion of rotation, and, *if polyatomic*, the atoms may vibrate with respect to each other. A billiard ball upon the table does not have so great freedom of motion as the gas molecule, for, although the ball has as much freedom of rotation as the molecule, its motion of translation is confined to a plane. A flat disk sliding upon the ice is not so free as is the billiard ball, for the disk is capable of rotation about only one axis, *viz.*, the vertical axis. A block sliding in a closely fitting groove is still more restricted in its freedom of motion.

Now, it is obvious that the relative freedom of motion of these several bodies can be expressed by giving the number of independent quantities which need to be known to express completely the position and orientation of the bodies. Thus, in the case of the block sliding in the groove, its position and orientation are completely specified by one quantity—its distance from some origin; for the disk on the ice, three quantities are required—two to express the position of some point, say the center of gravity of the disk with respect to some coordinate system on the ice, and one more to specify the orientation of the disk with respect to some fixed line; and so on. *These separate and independent quantities which need to be known to specify completely the position and configuration of a body are called its "degrees of freedom."*

This concept of degrees of freedom is readily extended to a *system* of bodies. If we have *three* disks sliding on the ice, *nine* independent quantities are necessary to express the "position and configuration

¹ For an outline of the arguments of Wien and of Rayleigh see Preston's "Theory of Heat."

of the system" Degrees of freedom are additive—the total number of degrees of freedom of a system of bodies is the sum of the number of degrees of freedom possessed by the several bodies which make up the system.

To each degree of freedom of a system there corresponds an independent term in the expression representing its kinetic energy as a function of its coordinates. The number of these terms is equal to the number of the degrees of freedom of the system.

For example, a monatomic gas molecule behaves as if it had a motion of translation only. Accordingly, its kinetic energy is completely expressed with respect to an x-, y-, z-coordinate system by the terms

$$\frac{1}{2}mv_x^2 + \frac{1}{2}mv_y^2 + \frac{1}{2}mv_z^2,$$

where v_x , v_y , and v_z refer to the component velocities in the three coordinate directions and m is the mass of the molecule. If there are N monatomic molecules in a given mass of gas, $3N$ such independent terms are necessary to express the kinetic energy of the gas at any given instant. On the contrary, a rigid diatomic gas molecule would have 5 degrees of freedom, 3 for translation (of the center of gravity) and 2 for rotation of the "dumbbell-shaped" molecule about each of a pair of axes at right angles to each other and to the line joining the atoms. If the molecule is not rigid, but the atoms can vibrate with respect to each other, then there is an additional degree of freedom corresponding to this vibration, and the total number is 6. A triatomic molecule will have at least 6 degrees of freedom (unless all atoms lie on a line), 3 for translation and 3 for rotation about 3 mutually perpendicular axes; if vibration of the atoms within the molecule is also possible, there will be 9 degrees of freedom in all.

79. Relation between Energy per Degree of Freedom and the Temperature.—A simple means for calculating the energy of a system, according to classical theory, is furnished by the principle of the *equipartition of energy*. To illustrate the meaning of this principle, consider a monatomic gas in an enclosure at a uniform temperature T . The molecules of such a gas are in motion in all possible directions with a wide range of velocities. Their degrees of freedom may be divided into three groups, corresponding to the three coordinate axes, viz., in x group, a y group, and a z group. The classical theorem of the *equipartition of energy* now says that *each degree of freedom in one group has, on the average, exactly the same amount of kinetic energy as has, on the average, a degree of freedom in any other group*. For example, let us take a census, or "snapshot," of 1,000 of the gas molecules. Each

molecule has a component of velocity in each of the coordinate directions x , y , z . Let v_x , v_y , and v_z be these components, and $\frac{1}{2}mv_x^2$, $\frac{1}{2}mv_y^2$, $\frac{1}{2}mv_z^2$ the corresponding kinetic energies. The *average* kinetic energy per molecule $\bar{E}_x$ in the x direction for the 1,000 molecules is obtained by adding the terms $\frac{1}{2}mv_x^2$ for all the molecules and dividing by 1,000, and similarly for the y - and the z -components. The theorem of the equipartition of energy then states that

$$\bar{E}_x = \bar{E}_y = \bar{E}_z.$$

Let us designate by $\bar{E}_1$ the common value of $\bar{E}_x$, $\bar{E}_y$, and $\bar{E}_z$. Then, if in the gas under consideration there are N molecules each with 3 degrees of freedom, the total kinetic energy of the system, due to the random motion of its molecules, is $3NE_1$. The same principle holds for more complicated molecules. The mean kinetic energy of a rigid diatomic molecule, for example, according to classical theory, would be $5E_1$; of a rigid triatomic one, $6E_1$.

Even in the case of vibratory motion, such as a vibration of the 2 atoms composing a diatomic molecule along the line joining their centers, the principle asserts that the average *kinetic* energy associated with such a degree of freedom will likewise be $\bar{E}_1$. If the motion is simple harmonic, however, it is easily shown that the average amount of potential energy is the same as the average kinetic energy. Hence the total amount of energy associated with such a vibratory motion, according to classical theory, is $2\bar{E}_1$. The *total* average energy of a *vibrating* diatomic molecule will then be $5\bar{E}_1 + 2\bar{E}_1 = 7\bar{E}_1$.

Now the quantity E_1 is a very simple universal function of the temperature. This may be seen as follows:

According to the kinetic theory of gases, the pressure p which a gas exerts on an enclosure is given by¹

$$p = \frac{1}{3} \rho v_g^2,$$

where p is the density of the gas and v_g is the "root mean square" velocity of its molecules. Since $p = nm$, where n is the number of molecules per cubic centimeter in the gas and m is the mass of each molecule, we have

$$p = \frac{1}{3} nm v_g^2.$$

Assume that we are dealing with a monatomic gas at temperature T . Then the average kinetic energy E_K of each molecule of the gas is given by

$$E_K = \frac{1}{2} m v_g^2,$$

¹ See any textbook on general physics.

and we may write, instead of the last equation,

$$p = \frac{2}{3} n \bar{E}_K.$$

But

$$pV_M = RT,$$

where VM is the volume of a gram-molecule of the gas at temperature r , and R is a universal constant known as the "gas constant," the numerical value of which is 8.315×10^7 ergs per mole per degree, or 1.986 calories per mole per degree.

Eliminating p between the last two equations, we have

$$\frac{2}{3} n \bar{E}_K V_M = RT.$$

The product of n , the number of molecules per cubic centimeter, by VM , the volume of the gram-molecule, is equal to N_0 , the number of molecules in a gram-molecule:

$$nV_M = N_0.$$

N_0 , known as *Avogadro's number*, is constant for all substances. Its numerical value, as given in Sec. 42, is 6.02×10^{23} molecules per gram-molecule. Eliminating VM between the last two equations and solving for E_K , we have

$$\bar{E}_K = \frac{3}{2} \frac{R}{N_0} T = \frac{1}{2} kT \quad (100)$$

where

$$k = \frac{R}{N_0}. \quad (101)$$

The new universal constant k is known as "Boltzmann's constant," or as the "molecular gas constant"; its value is

$$k = \frac{8.315 \times 10^7}{6.02 \times 10^{23}} = 1.381 \times 10^{-16} \text{ erg per molecule per degree.}$$

Equation (100) gives the average kinetic energy of the monatomic molecule which we are considering. Since such a molecule has 3 degrees of freedom, its *average kinetic energy per degree of freedom* is

$$\bar{E}_1 = \frac{1}{2} kT; \quad (102)$$

and the average total energy associated with a simple-harmonic degree of freedom is

$$2\bar{E}_1 = kT; \quad (103)$$

The principle of the equipartition of energy has been shown in a fairly convincing way to follow as a consequence of Newton's laws of motion as applied to complex systems. It has been strikingly confirmed by observations upon the Brownian movements of small particles suspended in a fluid, which we have not space here to describe;¹ these particles move about in thermal agitation like gigantic molecules. It is abundantly clear today, however, that the principle has only limited validity in relation to the internal motions of molecules of ordinary size.

80. Degrees of Freedom in an Enclosure.—It was suggested by Lord Rayleigh² and by Jeans³ that the law of the equipartition of energy might be applied to the radiation problem by computing the number of modes of free vibrations in the ether in an enclosure and by assuming that with each mode of vibration, or degree of freedom, there is associated, on the average, energy of magnitude kT per degree of freedom.

Any vibrating system is capable, in general, of a great many modes of vibration. Thus, a violin string or an organ pipe may

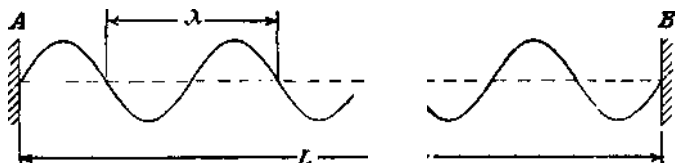


FIG 43.

vibrate not only in its fundamental mode but also in a great many overtones. We may think of these several overtones as corresponding to the various degrees of freedom of the vibrating system.

An organ pipe is a one-dimensional system, so far as its standing sound waves are concerned, since the vibrations take place longitudinally and in one direction only—viz., parallel to the axis of the pipe. Standing waves may be set up, also, in a "three-dimensional organ pipe," i.e., in a hollow enclosure such as a large concrete-lined room. In this case, too, we may have a great many possible overtones, or modes of vibration. We shall compute the number of such modes of vibration within a given frequency range for sound waves, and shall then extend the results to the analogous case of standing electromagnetic waves within an enclosure.

¹ See KENNAKD, E. H., "Kinetic Theory of Gases," p. 281, 1938; or MILLIKAN, ~R. A., "The Electron."

² *PhU. Mag.*, vol. 49, p. 539 (1900).

³ *Phil Mag.*, vol. 10, p. 91 (1905).

It will be instructive to apply the method of computation first to a one-dimensional case. Let a string of great length L (Fig. 43) be stretched between two supports A and B . The condition that standing waves may be set up requires (1) that A and B (i.e., the ends of the string) shall be nodes of motion and (2) that there shall be an integral number of equal loops between A and B . Since each loop is equivalent to one-half of a wave length, we may say that standing waves may occur in the string only for those wave lengths λ defined by

$$\frac{L}{\lambda/2} = i = \frac{2L}{\lambda}, \quad (104)$$

where i is some positive integer. The standing waves may be regarded as due to the superposition of two trains of waves of equal amplitude and wave length running in opposite directions along the string.

For a string of given length, i is equal, also, to the number of possible modes of vibration of the string for wave lengths equal to or greater than the value of λ as given by (104). We wish to know how many of these waves have lengths within a range $\Delta\lambda$. Now if we increase i by some integer, Δi , the value of λ given by (104) decreases to $\lambda - \Delta\lambda$, where

$$\frac{2L}{\lambda - \Delta\lambda} = i + \Delta i.$$

Subtracting Eq. (104) from this equation, we obtain

$$\Delta i = 2L \left(\frac{1}{\lambda - \Delta\lambda} - \frac{1}{\lambda} \right) = \frac{2L\Delta\lambda}{\lambda(\lambda - \Delta\lambda)}.$$

Since L is very great, however, $\Delta\lambda$, as given by this last equation must be very small; hence, in the denominator of the last fraction, $\lambda - \Delta\lambda$ can be neglected in comparison with λ , and we can write

$$\Delta i = 2L \frac{\Delta\lambda}{\lambda^2}.$$

Here Δi represents the number of modes of vibration whose wave lengths lie in the range $\Delta\lambda$.

Now, associated with each of these modes of vibration, there are 2 degrees of freedom, since the vibration is transverse and any point on the string is free to move in a *plane* at right angles to the string. The total number of degrees of freedom per unit length of the string in the wave-length range from λ to $\lambda - \Delta\lambda$ is, therefore,

$$\Delta n = 2 \frac{\Delta i}{L} = 4 \frac{\Delta\lambda}{\lambda^2}. \quad (105)$$

The situation is more complex but involves identically similar principles when we consider the possible systems of standing waves in a hollow enclosure. Let us discuss first the system of sound waves in a rectangular flat box, a sort of two-dimensional organ pipe. Let $A B C D$ (Fig. 44) represent the box, and consider a system of plane-parallel sound waves moving in directions parallel to the top and bottom of the box and being reflected, without absorption, from

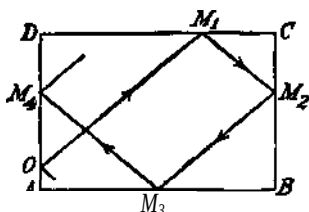


FIG. 44.—Reflection of waves inside a box.

its sides, according to the ordinary law of reflection. A set of waves moving initially in the direction OM_1 will, after reflection at the face CD , move in the direction M_1M_2 ; after reflection at M_2 , the direction M_2M_3 will be parallel but opposite to OM_1 , etc. For this group of waves, only four directions of motion are possible, $+OM_1$ and $\pm M_1M_2$.

Under suitable conditions the four trains of waves thus formed will combine to produce a set of *standing waves*. To describe these, divide the sides AB and DC of the box into i_1 equal segments and the sides AD and BC into i_2 segments, as in Fig. 45; and join the points of division by two sets of parallel lines, lines at each point making equal angles with the sides (e.g., $\angle r_3q_2B = \angle s_3q_2C$). Then a group of waves, such as those just described, can be supposed to move so that they are always parallel to one or the other of these two sets of lines. At a given instant, let alternate parallel lines represent, respectively, crests and troughs of these waves. At the instant of observation, suppose a crest moving toward C , for example, coincides with the line p_3r_3 ; as time goes on, this crest becomes extended at the r_3 end by reflection of a crest now at r_2q_2 and in motion toward B , and of another crest now at p_3s_1 and moving toward D . A similar crest, s_3q_2 , moving toward C is itself undergoing progressive reflection with the effect of extending crests p_3s_3 and r_3q_2 ; and so on. Upon these waves, then, let there be superposed a set of waves of equal amplitude moving in the opposite directions, crests moving toward A coinciding momentarily with p_3r_3 and s_3q_2 ; and so on. We obtain in this way four sets of waves mutually producing each other by reflection at the walls, as in Fig. 44.

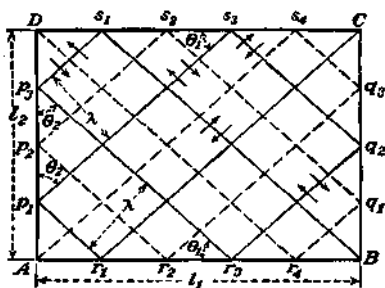


FIG. 45.—Standing waves in a two-dimensional box.

The set of waves just described will combine to form standing waves. For after half a period has elapsed, the crests, initially on the solid lines, will be meeting on the dotted lines, whereas troughs will now meet on the solid lines; the result is a disturbance of which the phase is momentarily just opposite to the previous one. After another half period, the original distribution recurs. Plane surfaces parallel to the sides of the box through the points of subdivision are, like the sides themselves, surfaces of no disturbance. It can be shown that, by proceeding in the manner described, using in turn all pairs of values for i_1 and i_2 , we obtain all possible modes of vibration of the medium inside the box.

For our purpose, the connection between the numbers i_1 and i_2 and the length of the waves is important. In Fig. 45, the wave length is the distance between adjacent parallel solid lines. Let the waves make angles θ_1 and θ_2 with the sides of the box, of which the lengths are h and Z_2 respectively. Then, since $r_1 r_2 = r_2 r_3 = l_1 / i_1$,

$$(r_1 r_3) \sin \theta_1 = \frac{2l_1}{i_1} \sin \theta_1 = \lambda,$$

$$(p_1 p_2) \sin \theta_2 = \frac{2l_2}{i_2} \sin \theta_2 = \lambda.$$

Solving these equations for $\sin \theta_1$ and $\sin \theta_2$ in terms of l_1, l_2 and substituting in the equation $\sin^2 \theta_1 + \sin^2 \theta_2 = \sin^2 \theta_1 + \cos^2 \theta_1 = 1$, we obtain

$$\frac{i_1^2}{l_1^2} + \frac{i_2^2}{l_2^2} = \frac{4}{\lambda^2}.$$

This method of constructing standing waves is readily extended to a *three-dimensional* rectangular box, with edges Z_1, l_2, l_3 long. The edges are divided into i_1, i_2 , and i_3 segments, respectively; and eight sets of waves are drawn. It is then found that instead of the last equation the condition for standing waves is

$$\frac{i_1^2}{l_1^2} + \frac{i_2^2}{l_2^2} + \frac{i_3^2}{l_3^2} = \frac{4}{\lambda^2}. \quad (106)$$

From this equation we can find the number of possible modes of vibration within the box for which the wave length is greater than any given value λ_m . The number of such modes will be equal to the number of possible combinations of positive integers i_1, i_2, i_3 , which make the left-hand member of (106) less than $4/\lambda_m^2$. To find this number, imagine each possible set of the i 's represented on a three-dimensional plot by a point the coordinates of which are $x = i_1/l_1, y = i_2/l_2, z = i_3/l_3$ (Fig. 46). These points will lie at the corners of

rectangular parallelepipeds or cells the edges of which are $1/l_1$, $1/l_2$, $1/l_3$ long (two of them being shown in the figure). There are just as many cells as there are points; each cell has eight points at its corners, but, on the other hand, each point serves as a corner to eight cells. Hence, since the volume of a cell is $1/l_1 l_2 l_3$, there are $l_1 l_2 l_3$ cells or points in each unit volume of the plot. The points we wish to count, according to Eq. (106), lie inside the sphere defined by the equation

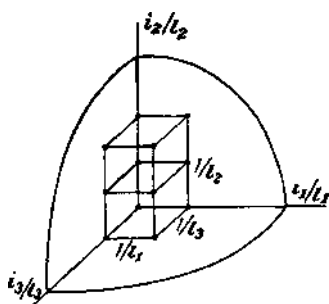


FIG. 46.

$$x^2 + y^2 + z^2 = \frac{4}{\lambda_m^2},$$

of which the radius is $2/\lambda_m$. The number of these points will be equal to the volume of 1 octant of this sphere multiplied by the number of points in unit volume, or to¹

$$\left[\frac{1}{8} \times \frac{4}{3} \pi \left(\frac{2}{\lambda_m} \right)^3 \right] \times l_1 l_2 l_3 = \frac{4\pi}{3} \frac{l_1 l_2 l_3}{\lambda_m^3}.$$

This is the total number of the possible modes of vibration for which $\lambda > \lambda_m$. Dividing it by the volume of the box, which is $h h h$, and dropping the subscript m , we have, therefore, for the number of modes of vibration, or of degrees of freedom, per unit volume of the box, corresponding to wave lengths greater than λ ,

$$n = \frac{4\pi}{3} \frac{1}{\lambda^3}.$$

Differentiation of this equation gives as the number of degrees of freedom per unit volume in a wave-length range λ to $\lambda + d\lambda$

$$dn = 4\pi \frac{d\lambda}{\lambda^4}, \quad (107)$$

both dn and $d\lambda$ being taken positive for simplicity. This is analogous to Eq. (105) for the string.

In the discussion up to this point we have been dealing with the longitudinal vibrations of sound waves. An identical analysis will hold for electromagnetic waves inside a rectangular enclosure with reflecting walls, with one exception. In the latter case, since these waves are transverse, the number of degrees of freedom is twice as great as for longitudinal waves, corresponding to the two possible

¹ This is not quite accurate, of course, because of errors at the surface, but, when n is large, the *relative* error is negligible.

independent planes of polarization. For *electromagnetic* waves in an enclosure, therefore, we may write

$$dn = 8\pi \frac{d\lambda}{\lambda^4}, \quad (108)$$

where dn is the number of degrees of freedom of the system of waves, per unit volume of the enclosure, in the wave-length range λ to $\lambda + d\lambda$.

81. The Rayleigh-Jeans Formula.—We are now in position to apply to the radiation problem the law of the equipartition of energy. Consider the radiation within an enclosure the walls of which are at a temperature T , equilibrium having been reached between the radiation and the walls. Whatever may be the nature of the radiating and absorbing mechanism of the walls, each degree of freedom of the *electromagnetic field* within the walls should have associated with it an *average* energy kT , and this quantity, multiplied by the number of degrees of freedom per unit volume for the wave-length range λ to $\lambda + d\lambda$, should give the amount of radiant energy per unit volume in this range, which was denoted previously (Sec. 70) by $\psi_\lambda d\lambda$. Thus, using (108), we have

$$\psi_\lambda d\lambda = 8\pi kT \lambda^{-4} d\lambda \quad (109)$$

as the equation for the spectral energy distribution in black-body radiation as deduced from the theorem of the equipartition of energy. This is the famous Rayleigh-Jeans formula.

It is obvious, at once, however, that Eq. (109) cannot actually represent the distribution of energy *throughout the spectrum* of a black body. For at a given temperature T , the value of ψ_λ , according to this equation, should vary inversely as the fourth power of the wave length and should *increase rapidly to infinity as the wave length approaches zero*, whereas the actual curves shown in Fig. 41 rise to a maximum with decreasing wave length *and then decrease to zero*. Only at long wave lengths, in fact, is the Rayleigh-Jeans expression found to be in accord with experiment.

A little consideration shows, in fact, that the law of the equipartition of energy and the computation of the number of degrees of freedom in the electromagnetic field, resulting in Eq. (108), cannot simultaneously be correct. For the field should be capable of containing waves of wave length varying from infinity to zero. Hence, by integration of Eq. (108) from $\lambda = \infty$ to $\lambda = 0$, we find that the number of degrees of freedom in the field should be infinite. And, if the law of the equipartition of energy is correct, it follows that the energy density of black-body radiation, when in equilibrium with the

walls of an enclosure at a given temperature, should be vastly greater, indeed infinitely greater, than the energy density in the walls themselves. In reality quite the opposite is the case. Thus the radiation density in an enclosure at 1000°K, is only of the order of 8×10^{-3} erg per cm.³, whereas the energy density of the thermal agitation of the molecules of the walls, say of iron, is of the order of 10^{10} ergs per cm.³

82.. Planck's Investigation of Black-body Radiation.—The correct black-body formula was discovered in 1901¹ by Max Planck. By introducing a radical innovation quite at variance with previous concepts, he discovered a function of λT which gave a formula in complete agreement with experiment. *This was the birth of the quantum theory.* The remainder of this chapter will be devoted to a discussion of the ideas leading up to Planck's quantum hypothesis and of his new radiation law.

Planck arrived at his formula as the result of a long persistent effort to make the theory fit the facts. The experiments of Hertz on electromagnetic waves had seemed to give final confirmation to the electromagnetic theory of light, and this convinced Planck that the key to the black-body spectrum would be found in the laws governing the absorption and emission of radiation by electrical oscillators. We may imagine that the walls of an isothermal enclosure contain electrical oscillators of all frequencies, essentially similar to the Hertzian oscillator, and that the emission and absorption of radiation by the walls are caused by these oscillators

Investigation by means of electromagnetic theory, however, led Planck, before 1900, to the conclusion that an electrical oscillator, in the long run, would affect only radiation of the same frequency as that of the oscillator itself. Thus, in the state of equilibrium, there would be a definite ratio between the density of radiation of any frequency ν and the average energy of the oscillators of that frequency. The problem of the black-body spectrum was thus reduced to the problem of the average energy of an oscillator at a given temperature. Now we have seen (Sees. 78 and 79) that according to classical theory a harmonic oscillator in thermal equilibrium with its surroundings would have an average energy equal to kT . If this conclusion from the theory is combined with the value obtained by Planck for the ratio between oscillator energy and radiation density in the enclosure, the result is again the Rayleigh-Jeans radiation law, which was derived above in a different manner. Planck himself, however, did not accept the principle of equipartition for the oscillators. On the basis of a different assumption he was led, at first, to Wien's formula for the

¹ *Ann. d. Physik*, vol. 4, p. 553 (1901).

radiation density as stated in Eq. (99); and he believed for a time that this formula must necessarily be correct.

In 1900, however, new measurements of the black-body spectrum by Lummer and Pringsheim, and by Rubens and Kurlbaum, showed definitely that Wien's formula could not be correct. So Planck set to work to find an improved formula. First he discovered an empirical modification of Wien's formula that fitted the observations. Then he sought to modify the statistical theory so that it would lead to this new formula. He succeeded in doing this only after making a new assumption that broke drastically with classical principles.

83. Distribution and Average Energy of Harmonic Oscillators in Thermal Equilibrium.—The theory of black-body radiation as developed by Planck requires a knowledge of the average energy of an oscillator at a given temperature. This is a problem in statistical mechanics.¹ It is solved by finding first the distribution of energy among a large number of oscillators in thermal equilibrium.

Consider a linear oscillator consisting of a mass m that vibrates in simple harmonic motion along a line. At any instant the vibrating mass will have a certain displacement x from its position of equilibrium and also a certain velocity. For statistical purposes, however, it is more convenient to employ, instead of the velocity dx/dt , the momentum p , whose magnitude is

$$p = m \frac{dx}{dt}.$$

The potential energy of the mass can be written $\frac{1}{2} \beta x^2$, where the constant β is analogous to a spring constant; whereas the kinetic energy of the mass is $\frac{1}{2} m(dx/dt)^2 = p^2/2m$. Thus its total energy is

Elementary theory gives, also, for its frequency of vibration

$$\nu = \frac{1}{2\pi} \sqrt{\frac{\beta}{m}}. \quad (111)$$

The statistical law for a large set of N such oscillators in thermal equilibrium, as found from classical theory, is easily stated. At any given moment, the number dN of the oscillators that have their values

¹ Cf. KENNARD, E. H., "Kinetic Theory of Gases," Chap. IX, 1938; TOLMAN, R. C., "Principles of Statistical Mechanics," 1938; MAYER, J. E. and M. G., "Statistical Mechanics," 1940.

of x and p lying in any given ranges of value dx and dp , respectively, is given by the formula

$$dN = NCe^{-\epsilon/kT} dx dp, \quad (112)$$

C being a constant. This law is obtained by an argument similar to that which leads in classical theory to the law of the equipartition of energy, or to Maxwell's law for the distribution of the molecular velocities in a gas. Equation (112) is very similar, in fact, to the mathematical formula that expresses Maxwell's law for molecular motions.

The distribution law will better serve our present purpose, however, if it is thrown into a slightly different form. Let us construct a plot

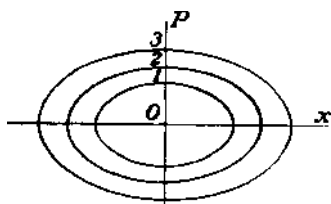


FIG. 47.— The momentum-coordinate plane for a harmonic oscillator

on which x and p are taken as Cartesian coordinates (Fig. 47). On this plot $dx dp$ is an element of area. Now, as the oscillator moves, free from the action of disturbing forces, x and p change with the time in such a way that the energy ϵ remains constant; hence the point whose coordinates are x and p , representing the instantaneous state of the oscillator, moves on a curve given by Eq. (110) with a fixed value of ϵ . This curve is obviously an ellipse.

It is natural, therefore, to take as an element of area, instead of $dx dp$, the elliptical ring between two ellipses corresponding respectively, to energies ϵ and $\epsilon + d\epsilon$. The area of such a ring can be expressed in terms of ϵ . For the entire area inside one of the ellipses is

$$S = \pi p_m x_m,$$

where p_m , x_m are its semiaxes, representing the maximum values of p and x , respectively, during a vibration. Putting in (110) first $p = p_m$ and $x = 0$, then $p = 0$ and $x = x_m$, we find

$$p_m = \sqrt{2m\epsilon}, \quad x_m = \sqrt{\frac{2\epsilon}{\beta}}.$$

Hence

$$S = 2\pi\epsilon \sqrt{\frac{m}{\beta}} = \frac{\epsilon}{\nu} \quad (113)$$

by (111). The area of a ring corresponding to an increment of energy ch is thus

$$dS = \frac{d\epsilon}{\nu}.$$

Substituting this value of dS for $dx dp$ in (112) and writing C_1 for C/ν , we have

$$dN = NC_1 e^{-\epsilon/kT} d\epsilon \quad (114)$$

as the number of oscillators of which the energy lies between given values ϵ and $\epsilon + d\epsilon$, when a large number N of them are in the thermal equilibrium at absolute temperature T .

The *total energy* E of all N oscillators can now be found by summing their individual energies, and E divided by N gives then the average energy per oscillator. The summation will be made with use of a somewhat unusual notation in order to prepare for the introduction of Planck's new principle.

Let ellipses be drawn, as in Fig. 47, in such a way as to divide the entire xp plane into elliptical rings of equal area; and let these rings be numbered 0, 1, 2, 3, $\cdot$ $\cdot$ $\cdot$ from the center outward, as in Fig. 47. Thus "ring" number 0 is actually an elliptical area. Then, if we denote the area of a ring by h , the total area inside of ring number r is

$$S = r h;$$

and by Eq. (113) the energy of an oscillator represented by a point on the inner boundary of ring number r is

$$\epsilon = S\nu = r h \nu.$$

Thus the range of energy represented by points lying on a given ring is

$$\Delta\epsilon = \nu \Delta S = h \nu,$$

since r increases by unity from one ring boundary to the next.

The number of oscillators represented by points on ring number r can be found by integrating dN as given by Eq. (114) over the range $\Delta\epsilon$ of ϵ within the ring. Calling this number N_r , we thus find

$$N_r = NC_1 \int_{\Delta\epsilon} e^{-\epsilon/kT} d\epsilon.$$

If h and $A\epsilon$ are small, however, we can treat $e^{-\epsilon/kT}$ as constant in this integral and equal to its value at a point on the inner boundary of the ring; then

$$\int_{\Delta\epsilon} e^{-\epsilon/kT} d\epsilon = e^{-\epsilon/kT} \int_{\Delta\epsilon} d\epsilon = e^{-\epsilon/kT} \Delta\epsilon.$$

Thus, replacing $NC_1 \Delta\epsilon$ by N_0 , we have for the number of oscillators on ring number r

$$N_r = N_0 e^{-\frac{r h \nu}{kT}}. \quad (115)$$

Obviously the constant N_0 represents the number of oscillators in ring number 0. Its connection with the total number of oscillators N is easily found by noting that

$$N = N_0 + N_1 + N_2 + \cdots = N_0(1 + e^{-\frac{h\nu}{kT}} + e^{-\frac{2h\nu}{kT}} + \cdots) = \frac{N_0}{1 - e^{-h\nu/kT}}.$$

(For $1/(1 - x) = 1 + x + x^2 + \cdots$).

The *total energy* of the oscillators can then be found by multiplying the number in each ring by the energy of an oscillator when in that ring, for which, since h and $\Delta\epsilon$ have been assumed to be small, we can use the value $h\nu$ belonging to the inner boundary, and then summing over all rings. Thus the total energy is

$$\begin{aligned} E &= (N_0 \times 0) + (N_0 e^{-\frac{h\nu}{kT}} \times h\nu) + (N_0 e^{-\frac{2h\nu}{kT}} \times 2h\nu) + \cdots \\ &= N_0 h\nu e^{-h\nu/kT} (1 + 2e^{-h\nu/kT} + 3e^{-2h\nu/kT} + \cdots), \\ &= \frac{N_0 h\nu e^{-h\nu/kT}}{(1 - e^{-h\nu/kT})^2}, \end{aligned}$$

since the last series is of the form

$$1 + 2x + 3x^2 + \cdots = \frac{1}{(1 - x)^2}.$$

Dividing this expression for E by that found above for AT , we find finally for the *average energy* of an oscillator

$$\bar{\epsilon} = \frac{E}{N} = \frac{h\nu e^{-h\nu/kT}}{1 - e^{-h\nu/kT}} = \frac{h\nu}{e^{h\nu/kT} - 1}. \quad (116)$$

This expression represents the average value at a given instant of the energies of the individual oscillators. The same expression represents also the energy of a particular oscillator averaged over any length of time that is not too short. For, the oscillators, being similar, will all have the same average energies in the long run, and this average energy must be a fraction $1/iV$ of the average energy of the entire set, which is simply E .

The *classical value* for I can now be obtained from this expression by letting $h \rightarrow 0$. This converts Ae into an infinitesimal, de , so that our sums become integrals, as in the preceding calculation for the mean energy of a molecule. We use the series

$$e^x = 1 + x + \frac{1}{2}x^2 + \cdots,$$

so that to the first order in h

$$e^{h\nu/kT} - 1 = \frac{h\nu}{kT}$$

Then it is clear that according to (116), as $h \rightarrow 0$,

$$\bar{\epsilon} \rightarrow kT.$$

This is the same value for the average energy of a harmonic oscillator that was deduced from the equipartition of energy in Sees. 78 and 79, and it tartus, as we have seen, to the incorrect Rayleigh-Jeans formula.

84. Planck's Quantum Hypothesis.—The new assumption introduced by Planck was equivalent to *keeping the quantity ϵ finite* in the preceding formulas, instead of letting it go to zero.

In the first formulation of the new theory¹ Planck assumed that the oscillators associated with a given ring all have the energy proper to the inner boundary of that ring. According to this assumption, the energy of an oscillator *cannot vary continuously but must take on one of the discrete set of values*: $0, hv, 2hv, \dots, rhv, \dots$. Actually, Planck's original assumption was somewhat more general. He assumed that the energy of the oscillator must always be an integral multiple of a certain quantity, ϵ_0 . He then showed, however, that ϵ_0 must be proportional to ν if the radiation law is to harmonize with the Wien displacement law. Thus he assumed that $\epsilon_0 = h\nu$, where h is a constant of proportionality. The connection between the constant h and areas on the xp plane for the oscillator, as described above, was not recognized by Planck until later.

It must be emphasized that the assumption of a *discrete set* of possible energy values, or *energy levels*, for an oscillator was completely at variance with classical ideas. According to this assumption, if the energies of a large number of oscillators were measured, some might be found to have zero energy; some, energy hv ergs each; others, $2hv$; and so on. But not a single oscillator would be found which had energy, say $1.75 hv$ or $3.94 hv$. When the energy of a given oscillator changes, therefore, it must change suddenly and discontinuously. According to the older conceptions, on the other hand, the interchange of energy between two "systems," as, for example, between one gas molecule and another or between radiation and oscillators, should be a perfectly *continuous* process. Thus, if one could follow the "history" of a particular gas molecule or oscillator and observe its energy at short intervals of time over a very long period, one would find, of course, that its energy varied over enormously wide limits. But, if the numerical values of the energy observed at the various instants were to be arranged in a sequence as to magnitude, this sequence would become more and more *continuous* as the number of observations becomes greater. Or, if,

¹ PLANCK, *Ann. d. Physik*, vol. 4, p. 553 (1901).

instead of observing a single molecule over a long period, one were to observe, at a given instant, the energy of a very large number of gas molecules or of oscillators chosen at random, one ought to find that a similar sequence of values of energy would approach more and more nearly to continuity as the number of observations becomes larger. This continuity of these energy values is not only in accord with but is also imperatively demanded by classical physics. For example, the electric and magnetic vectors in a light wave may have any values whatsoever, *from zero up*; and, accordingly, the wave may have any intensity, *from zero up*. Furthermore, the emission and absorption of this energy by the walls of an enclosure should, likewise, be a perfectly continuous process.

The problem of the absorption and emission of radiation, in fact, presented serious difficulties for the new theory. If the energy of an oscillator can vary only discontinuously, the *absorption and emission of radiation* must likewise be *discontinuous processes*. As long as the oscillator remains in one of its "quantum states," as we now call them, with its energy equal to one of the allowed discrete values, it cannot be emitting or absorbing radiation according to the laws of classical physics, for then the conservation of energy would be violated. This is entirely contrary to classical electromagnetic theory; for we have seen (Sec. 35) that classical theory absolutely requires an isolated, accelerated electrical charge to radiate energy.

According to Planck's new theory, *emission* of radiation could occur only when the oscillator "jumps" from one energy level to another; if it jumps down to the next lower energy level, the energy $h\nu$ that it loses is emitted in the form of a short pulse of radiation. *Absorption* was also assumed at first to be discontinuous. An oscillator, Planck assumed, can absorb a quantum $h\nu$ of radiant energy and jump instantaneously (or nearly so) up to its next higher energy level. This assumption met with difficulties, however. For the quantum of radiant energy emitted by an oscillator, according to the classical wave theory, would spread out over an ever expanding wave front, and it is hard to see how another oscillator could ever gather this energy together again so as to absorb it all and thereby acquire the energy for an upward quantum jump. Absorption ought, therefore, on Planck's theory, to be impossible.

For this reason, Planck later modified his theory so as to allow the oscillators to absorb in a continuous manner, only the process of emission being discontinuous. The energy of an oscillator could then take on all values, as in classical theory; but, every time that the energy passed one of the critical values, $rh\nu$, there was assumed to

exist a certain chance that the oscillator would jump down to a lower energy level, emitting its excess energy as a quantum of radiation. This came to be known as the "second form" of Planck's quantum theory. Calculation showed that, on the theory as thus modified, the average energy of an oscillator, when in thermal equilibrium with its surroundings, would be

$$\bar{\epsilon} = \frac{h\nu}{e^{h\nu/kT} - 1} + \frac{1}{2}h\nu, \quad (\text{U7})$$

in place of the value given in Eq. (116) above.

Having read thus far, the student may perhaps have reached a state of fairly complete confusion as to *what were the essential* assumptions of Planck's quantum theory! This confusion can be no worse than that which existed in the minds of most physicists in the year, say, 1911. The situation was made still more puzzling by the success of Einstein's theory of the photoelectric effect, described above (Sec. 46); for Einstein assumed not only that radiation came in quantized spurts but that each spurt was closely concentrated in space, contrary to the wave theory. Confusion usually reigns while important physical advances are being made; it is only afterward that a clear-cut logical path can be laid down leading straight to the goal.

One of the principal aims of the rest of this book will be to show how the theory has gradually become clarified. To assist the student at this point, it may be stated that the following two new ideas introduced by Planck have been retained permanently and form a part of modern "wave mechanics" (Chap. VII):

1. *An oscillator, or any other physical system, has (in general) a discrete set of possible energy values or levels; energies intermediate between these allowed values never occur. [In wave mechanics, however, the energy of level number r for a harmonic oscillator is $(T + x^{1/2})h\nu$, and their average energy at temperature T is given by (117), not (116).]*

2. *The emission and absorption of radiation are associated with transitions or jumps between two of these levels, the energy thereby lost or gained by the oscillator being emitted or absorbed, respectively, as a quantum of radiant energy, of magnitude $h\nu$, ν being the frequency of the radiation.*

It should be emphasized that Planck's revolutionary assumptions were not based upon an extension of the ordinary lines of reasoning of classical physics. Quite the contrary: they represented an *empirical modification* of classical ideas made in order to bring the theoretical deductions into harmony with experiment.

Had the magnitude of the quantum of energy turned out to be not $h\nu$ but something *independent of the frequency*, the new theory might well have taken the form of a simple atomicity of energy, similar to the atomicity of electricity represented by the electronic charge. Such is not the case, however. Rather, it is the new universal constant h that represents the essentially new element introduced into physics by the quantum theory. We shall find h playing an important part in a wide variety of atomic phenomena.

85. Planck's Radiation Law.—Planck derived his new radiation formula by considering the interaction between the radiation inside an isothermal enclosure and electrical oscillators which he imagined to exist in the walls of the enclosure. A modern version of this deduction, in which gaseous atoms take the place of oscillators, first given by Einstein, will be sketched later [Sec. 116(6)]. An alternative and equally satisfactory procedure is to combine Planck's expression for the mean energy of an oscillator with the analysis of the electromagnetic field by the method of Rayleigh and Jeans, in which the various modes of oscillation of the field inside an enclosure are treated as if they were oscillators.

In Sec. 80, Eq. (108), we found that there would be

$$8\pi \frac{d\lambda}{\lambda^4}$$

such modes of oscillation or degrees of freedom per unit volume in the wave length range λ to $\lambda + d\lambda$. If we multiply this number, not by kT , the mean energy of an oscillator according to the law of equipartition, but by $\bar{\epsilon}$ as given by Eq. (116), we obtain

$$8\pi \frac{d\lambda}{\lambda^4} \frac{h\nu}{e^{h\nu/kT} - 1}.$$

This is the value of $\lambda d\lambda$, the energy density belonging to the range $d\lambda$. Let us substitute in it $\nu = c/\lambda$, c being the speed of light in vacuum. Thus we obtain, as Planck's new radiation law,

$$\psi_\lambda = \frac{8\pi ch}{\lambda^5} \frac{1}{e^{ch/\lambda kT} - 1} \quad (118)$$

The emissivity of a black body, e_λ , is then equal to ψ_λ as given by this formula multiplied by $c/4$ [cf. Eq. (786) in Sec. 71]. [Strictly speaking, we should have used for $\bar{\epsilon}$ the value given by Eq. (117), which agrees with the value obtained from wave mechanics. The effect of this change would be to add in ψ_λ a term independent of temperature. Since only *changes* in ψ are perceptible, however, this term would

be without physical effect. Its true significance is still unknown, and it is never included in the expression for ψ_λ or e_λ .]

It is easily shown that Planck's formula reduces to a form of Wien's formula near one end of the spectrum, and to the Rayleigh-Jeans formula near the other end. Thus, if the value of the product λT is *small* enough, *i.e.*, for sufficiently small wave lengths or at sufficiently low temperatures, the exponential term in the denominator in (118) is much larger than unity, so that in comparison with it the term -1 can be dropped. Then (118) becomes

$$\psi_\lambda = 8\pi ch\lambda^{-5}e^{-ch/\lambda kT},$$

which agrees with Eq. (99) or Wien's formula if in that equation we choose $c_1 = 8\pi ch$, $c_2 = ch/k$. For *large* values of λT , on the other hand, *i.e.*, for sufficiently long waves or at sufficiently high temperatures, we may expand the denominator in (118) by means of the series,

$$e^x = 1 + x + \frac{x^2}{2} + \cdots,$$

obtaining

$$\frac{ch}{e^{\lambda kT}} - 1 = \frac{ch}{\lambda kT} + \frac{c^2 h^2}{2\lambda^2 k^2 T^2} + \cdots$$

If λT is large enough, only the first term of this series need be kept; then we have, from (118),

$$\psi_\lambda = \frac{8\pi kT}{\lambda^4},$$

approximately. This agrees with the Rayleigh-Jeans formula as stated in Eq. (109) in Sec. 81.

In Fig. 48 is shown a comparison between the several spectral-energy distribution formulas and the experimental data. The circles show observations by Coblentz¹ on the energy distribution in the spectrum of a black body at 1600°K. The full line shows the distribution predicted by Planck's formula. The lower dotted line, which coincides with the full line from short wave lengths up to about $\lambda = 2.2/x$, corresponds to Wien's formula as given in Eq. (99). The upper dotted line is from the Rayleigh-Jeans formula. The superiority of Planck's formula is at once evident. Thus, whatever one may think of the theoretical assumptions and reasoning by means of which Planck arrived at his formula, there seems to be no doubt that it correctly represents the observations on black-body radiation.

¹ Bureau of Standards, Bull, vol. 13, p. 476 (1916).

The reason that the curve for Planck's formula drops below the Rayleigh-Jeans formula is to be found, of course, in the failure of the classical principle of the equipartition of energy. This is easily seen from Eq. (117). At short wave lengths, the -1 can be dropped in the denominator in that equation and we can write, as the approximate average energy of an oscillator,

$$\bar{\epsilon} = h\nu e^{-h\nu/kT} + \frac{1}{2}h\nu.$$

Thus, the part of $\bar{\epsilon}$ that varies with temperature decreases rapidly toward zero as $\nu \rightarrow \infty$. The high-frequency modes of oscillation of the

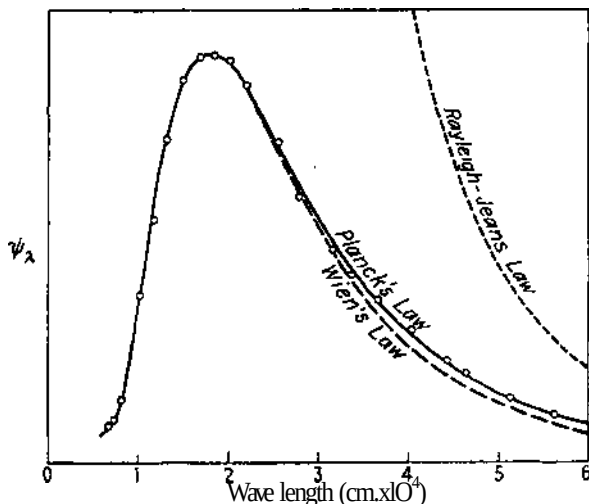


FIG. 48.—Companion of the three radiation laws with experiment at 1600°K. The ordinate represents ψ_λ or e_λ , on an arbitrary scale.

electromagnetic field in the enclosure, which should all have mean energy kT according to classical theory, remain, therefore, almost entirely in their lowest quantum states and so contribute nothing to the observable density of radiant energy.

It may be noted, also, that Planck's formula has the form that was found in Sec. 77 to be required by thermodynamics. For, using Eq. (786) in Sec. 71 and Planck's formula for e_λ , the emissive power of a black body can be written

$$e_\lambda = \frac{c}{4} \psi_\lambda = \frac{F(\lambda T)}{\lambda^5}, \quad F(\lambda T) = \frac{2\pi c^2 h}{e^{\frac{ch}{\lambda T}} - 1},$$

so that e_λ has the form given in Eq. (98).

Finally, as to numerical values, Planck's formula may be used to calculate the values of h and k , as was done by Planck himself in his

original paper. Observations of thermal quantities are difficult to make with accuracy, however, and more reliable values of h and k can be obtained in other ways. A good estimate based on recent measures is¹

$$h = 6.61 \times 10^{-27} \text{ erg sec.} \quad (119)$$

The value of k was given in Sec. 79. If these values are inserted, Planck's formula or Eq. (118) becomes, when λ is measured in centimeters and T in degrees centigrade,

$$\psi_\lambda = \frac{4.981 \times 10^{-15}}{\lambda^5} \frac{1}{e^{\frac{1.435}{\lambda T}} - 1} \frac{\text{ergs}}{\text{cm.}^2}.$$

Other quantities can also be calculated which are more easily compared with experiment. Thus, the constant a or a , defined in Sec. 73, can be derived by integration. The total radiant energy in unit volume of an isothermal enclosure is, using (118),

$$\psi = \int_0^\infty \psi_\lambda d\lambda = 8\pi ch \int_0^\infty \frac{d\lambda}{\lambda^5} \frac{1}{e^{\frac{ch}{\lambda kT}} - 1}.$$

Let us change the variable here to

$$x = \frac{ch}{\lambda kT}; \quad \therefore \lambda = \frac{ch}{xkT}, \quad d\lambda = -\frac{ch dx}{x^2 kT}.$$

Then

$$\psi = \frac{8\pi k^4 T^4}{c^3 h^3} \int_0^\infty \frac{x^3 dx}{e^x - 1}.$$

This integral can be

$$\psi = a T^4 \frac{\text{ergs}}{\text{cm.}^2}, \quad a = \frac{8\pi^5}{15} \frac{k^4}{c^3 h^3}. \quad (120a,6)$$

Hence

¹ DUNNINGTON, *Rev. Mod. Phys.*, vol. 1, p. 95 (1931).

² Since $(e^x - 1)^{-1} = \sum_{n=0}^{\infty} e^{-nx} = 1 - e^{-x} + e^{-2x} - e^{-3x} + \dots$, and

$$\int_0^\infty x^3 e^{-nx} dx = \frac{6}{n^4},$$

we have

$$\int_0^\infty \frac{x^3 dx}{e^x - 1} = 6 \left(1 + \frac{1}{2^4} + \frac{1}{3^4} + \dots \right).$$

The value of the series written here in parentheses, as given on p. 140 of the "Smithsonian Mathematical Formulae and Tables of Elliptic Functions" (Smithsonian Institution, 1922) is $\pi^4/90$. Cf. also Bierens de Hahn, "Nouvelle Table d'Integrates Definitives," No. 4, p. 124, where $B_{4,0} = \frac{1}{90}$.

$$\frac{d\psi_\lambda}{d\lambda} = -\frac{40\pi ch}{\lambda^6(e^{ch/\lambda kT} - 1)} + \frac{8\pi ch}{\lambda^5} \frac{ch}{\lambda^2 kT} \frac{e^{ch/\lambda kT}}{(e^{ch/\lambda kT} - 1)^2} = 0;$$

$$\left(1 - \frac{ch}{5\lambda kT}\right) e^{ch/\lambda kT} = 1.$$

$$\left(1 - \frac{x}{5}\right) e^x = 1, \quad x = \frac{ch}{\lambda kT}.$$

One root is obviously $x = 0$, but this one does not interest us. Any other positive root must be < 5 , for, if $x > 5$, the factor in parentheses is negative. If $x = 5$, the left-hand member is zero, whereas, if $x = 4$, it is much bigger than 1. Another root must lie, therefore, just below 5. To facilitate finding it, let us write $5 - x = y$, so that $x = 5 - y$ and the equation becomes

$$y e^{5-y} = 5$$

or

$$\log_e y - y = \log_e 5 - 5 = -3.3906.$$

To make the solution still easier, set $y = 2/10$; then

$$-\log_e z + \frac{z}{10} = 3.3906 - \log_e 10 = 1.0880.$$

It is now easy to see, from a table of natural logarithms, that to three figures $z = 0.349$; hence, $y = 0.0349$, $x = 4.965$. Thus, λ_m is given by

$$\frac{ch}{\lambda_m kT} = 4.965,$$

whence

$$\lambda_m T = \frac{ch}{4.965k}. \quad (121)$$

Inserting $h = 6.61 \times 10^{-27}$ and $k = 1.3805 \times 10^{-16}$, it is found from (120b) and (121) that

$$a = 7.619 \times 10^{-15} \text{ erg cm.}^{-3} \text{ deg.}^{-4}; \quad \lambda_m T = 0.2891 \text{ cm. deg.}$$

when λ_m is in centimeters and T in degrees centigrade; also, in Eq. (86) or $E = \sigma T^4$

$$\sigma = \frac{c}{4} a = 5.710 \times 10^{-5} \text{ erg cm.}^{-2} \text{ sec.}^{-1} \text{ deg.}^{-4}.$$

The pressure on the walls of an isothermal enclosure at absolute temperature T is thus, from Eq. (82),

$$p = \frac{1}{3} \psi = \frac{1}{3} a T^4 = 2.54 \times 10^{-15} T^4 \text{ dyne cm.}^{-2}$$

These values of a and of $\lambda_m T$ differ slightly from values based on observations of thermal radiation, viz.,¹

$$a = 7.652 \times 10^{-15}, \quad \lambda_m T = 0.2884.$$

The discrepancy is increased if in calculating a and $\lambda_m T$ from theory the value of h is employed that is derived from the Rydberg constant (Sec. 95), viz., $h = 6.624 \times 10^{-27}$ j with this value of h ,

$$a = 7.571 \times 10^{-15}, \quad \lambda_m T = 0.2897.$$

It is believed, however, that the thermal observations may easily be in error even to this extent.²

In his original paper (1901) Planck used as experimental values $a = 7.061 \times 10^{-15}$, $\lambda_m T = 0.294$, and calculated from these values and his formulas that

$$h = 6.55 \times 10^{-27}, \quad k = 1.346 \times 10^{-16}.$$

¹ Cf. BIRGE, *Rev. Modern Phys.*, vol. 1, p. 1 (1929). Values of a are by Kussman (1924) and Coblenz (1920), of $\lambda_m T$ by Coblenz (1922) and Ladenburg.

² *Ibid.*, vol. 13, p. 233 (1944).

CHAPTER VI

THE NUCLEAR ATOM AND THE ORIGIN OF SPECTRAL LINES

We have seen in the last chapter that attempts to explain the experimentally observed laws of the distribution of energy in the continuous spectrum emitted by a black body were unsuccessful until Planck introduced the revolutionary concept of radiation quanta. Planck's hypothesis became both *possible and necessary*, because the very careful experiments of Lummer and Pringsheim had proved that Wien's law of temperature radiation was untenable.

A somewhat similar sequence of events is to be found in the development of our present extensive knowledge of the *line* spectra of atoms and molecules. Corresponding to the empirical laws of temperature radiation, there was accumulated a vast array of very accurate measurements of the wave lengths of lines in the spectra of various substances, considerable impetus being given to this work because of the rigorous demands of spectroscopy as a method of chemical analysis. From these data, certain relations were discovered empirically between the frequencies of various lines in the spectra of certain elements. These relations pointed to some fundamental mechanism, common to all atoms, as the origin of characteristic line spectra. The gradual accumulation of evidence bearing on the problem of atomic structure, on the one hand, and the increasing importance of these spectral relations, on the other, culminated, about 1913, in the proposal, by Bohr, of the theory of atomic structure and the origin of spectra which bears his name. This theory constituted an extension in a new field of the quantum theory which had been introduced by Planck to explain the law of the distribution of energy in temperature radiation.

In the present chapter, we shall begin by considering very briefly the development of the empirical laws of spectral series. Next we shall take up the lines of evidence that were accumulated bearing on the structure of the atom, culminating in the famous nuclear theory of atomic structure proposed by Rutherford. Then we shall consider the way in which these two lines of development were combined by Bohr in his famous quantum theory of the atom. Bohr's theory has been superseded by modern quantum mechanics (or wave mechanics), but certain features of his theory retain a permanent interest.

86. Spectroscopic Units.—In stating the wave lengths of spectral lines, which are always very small, various submultiples of the meter are commonly employed as units of length in different parts of the spectrum, *viz.*,

The micron, symbol μ = 10^{-4} cm. (or 10^{-6} meter)

The millimicron, symbol $m\mu$ = 10^{-7} cm.

The angstrom,¹ symbol $\text{\AA}$ = 10^{-8} cm.

(or Ångström, symbol $\text{\AA}$.)

The X-unit, symbol X.U. = 10^{-11} cm. (approximately;

see Sec. 175).

The last named is used in the X-ray region of the spectrum.

Wave lengths of lines are of fundamental importance in the technique of spectroscopy and in spectroscopic analysis. In physical theory, on the contrary, *frequency* ν is more fundamental than wave length. We do not, however, measure frequencies *directly*; laboratory measurements yield *wave lengths*, and frequencies are *computed* from these and from the velocity of light c by the relation, $c = \nu\lambda$.

Frequencies may be expressed in vibrations per second, but this involves very large numbers. Furthermore, the calculated frequency is affected, not only by errors in the measured wave lengths, but also by any error there may be in the assumed velocity of light. For these reasons, it is customary in spectroscopy to use, instead of the frequency itself, the *wave number*, or number of waves per centimeter in vacuum, which we shall denote by $\bar{\nu}$. The unit for $\bar{\nu}$ is written cm.^{-1} ; it may be read "waves per centimeter." Thus

$$\bar{\nu} = \frac{1}{\lambda} \text{cm.}^{-1}, \quad \nu = c\bar{\nu},$$

¹ Strictly speaking, the angstrom is not defined from the meter as a *primary* standard of length. Michelson and Benoist in 1895 and, later (1907), Fabry, Perot, and Benoist measured the wave length of the red cadmium line in terms of the standard meter; this line is very narrow, so that its wave length can be measured with great accuracy. The two measurements were almost exactly in agreement, the wave length according to the latter measurement being

6,438.4696 angstroms.

The International Union for Solar Research, in 1907, adopted this value of the wave length of the red cadmium line as the *primary* standard of wave lengths on the basis of which all other wave lengths were to be expressed. It is specified that the medium is to be dry air at 15°C. (hydrogen scale) and a pressure equal to 760 mm. Hg at a place where the acceleration due to gravity is 980.67 cm. sec.⁻². Formally, this amounts to a new definition of the angstrom in terms of the wave length of the cadmium line such that this wave length is *exactly* 6,438.4696 angstroms.

λ being the wave length in vacuum expressed in centimeters. Or, if λ_μ is the wave length in microns, or λ_A the wave length in angstroms,

$$\bar{\nu} = 10^4 \times \frac{1}{\lambda_\mu} \text{ cm.}^{-1} = 10^8 \times \frac{1}{\lambda_A} \text{ cm.}^{-1}.$$

For example, the wave number of the red cadmium line

($X = 6,438.4696 \text{ \AA} = 0.6438\text{M} = 6.438 \times 10^{-5} \text{ cm.}$, approximately)

is $\nu = 15,531.64 \text{ cm.}^{-1}$; its frequency is 2.99796×10^{10} times ν or $4.65632 \times 10^{14} \text{ sec.}^{-1}$. Very roughly, the visible region (4,000 to 7,000 $\text{\AA}$) extends from 14,000 to 25,000 cm.^{-1} ; from 14,000 down to 1,000 is the near infrared region (out to 10 μ); from 1,000 to 100, the far infrared. The ultraviolet region extends from 25,000 to 100,000 cm.^{-1} (1,000 $\text{\AA}$). The principal X-ray region is from 10^7 to 10^8 cm.^{-1} .

The wave length of a line *in air* is slightly different from the wave length *in vacuum*. Since ordinary spectroscopic work is done in air, wave lengths above (about) 2,000 angstroms are commonly given *in air*.¹ Below 2,000 $\text{\AA}$, however, they are usually given in vacuum, since vacuum spectrographs are commonly employed for work in that region. The two wave lengths are related by the equation

$$\lambda_{\text{vac}} = \mu \lambda_{\text{air}},$$

where μ is the refractive index of air; thus the relation actually used ordinarily in obtaining wave numbers from optical wave lengths in air, or vice versa, is

$$\bar{\nu} = \frac{1}{\mu \lambda_{\text{air}}}.$$

The difference between λ_{vac} and λ_{air} is somewhat less than 1 part in 3,000. Tables have been published to facilitate the conversion from λ_{air} to ν or vice versa.²

Measurements of wave lengths in the visible and near-visible region of the spectrum are now possible with an accuracy of at least 1 part in 1 million.

87. Early Search for Regularities in Spectra.—One of the first features of spectra to be noticed was that the spectrum of a given element depended a great deal upon its mode of excitation. When the spectrum was produced by a *spark*, many lines were observed which were absent, or at least much weaker, when a steady arc

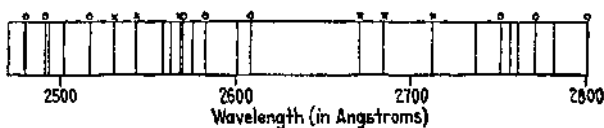
¹ Cf. the recent extensive table, "Massachusetts Institute of Technology Wave-length Tables," John Wiley & Sons, Inc., New York, 1939.

² Cf. KAYSER, H., "Tabelle der Schwingungszahlen," S. Hirzel, Leipzig, 1925.

or a flame was employed. These lines are said to form the *spark spectrum* (formerly, the "enhanced" spectrum) of the element, those lines present in the arc forming its *arc spectrum*. It was later established that arc spectra are emitted by neutral atoms, spark spectra by ionized atoms.

As soon as dependable wave-length measurements became available, numerous investigators, reasoning from the analogy of overtones in acoustics, sought for harmonic relations in the lines found in the spectrum of a given element. This search proved fruitless; but certain relations of a different type were discovered.

Liveing and Dewar,¹ in the early eighties, emphasized the physical similarities occurring in the spectra of such elements as the alkali metals. They called attention to the successive *pairs* of lines in the arc spectrum of sodium and pointed out that these pairs were alternately "sharp" and "diffuse" and that they were more closely crowded together toward the short-wave-length end of the spectrum,



P'IG. 49.—Some lines in the spectrum of zinc in a limited wave-length range in the ultraviolet. The lines marked with a cross (X) belong to one series of triplets, lines marked (O) belong to another series.

suggesting some kind of series relation, which, however, they were unable to discover. A little later, Hartley² discovered an important numerical relationship between the components of doublets or triplets in the spectrum of a given element. If frequencies, instead of wave lengths, are used, Hartley found that *the difference in frequency between the components of a multiplet (i.e., doublet or triplet) in a particular spectrum is the same for all similar multiplets of lines in that spectrum.*

Hartley's law made it possible to isolate from the large number of lines in any given spectrum those groups of lines which were undoubtedly related. But the task was not an easy one. Figure 49 shows the lines in the zinc spectrum from about 2,500 to 2,800 Å. Two triplets having constant frequency differences are marked by crosses (X) in Fig. 49. This same spectral region, however, contains another overlapping series of triplets, which are designated by circles (O). The sorting out of these related lines required a great deal of diligent and patient study.

The beginning of our knowledge of spectral-series formulas dates

¹ *Roy. Soc. Proc.*, vol. 29, p. 398 (1879); vol. 30, p. 93 (1880).

² *Chem. Soc. J.*, vol. 43, p. 390 (1883).

from the discovery by Balmer,¹ in 1885, that the wave lengths of the nine then known lines in the spectrum of hydrogen could be expressed by the very simple formula

$$\lambda = b \frac{n^2}{n^2 - 4}, \quad (122)$$

where b is a constant the numerical value of which, to give λ in angstroms, is 3,645.6 and n is a variable integer which takes on the successive values 3, 4, 5, $\dots$ for, respectively, the first (beginning at the red), second, third, $\dots$ line in the spectrum. Balmer compared the predictions of this formula with the best values then available for the wave lengths of the hydrogen lines. The four lines in the visible region had been measured by Angstrom and others; five ultraviolet lines in the spectrum of white stars had been measured by Huggins. Table I, taken from Balmer's paper, shows the comparison

TABLE I.—WAVE LENGTHS OF THE FIRST NINE HYDROGEN LINES
COMPUTED BY BALMER FROM HIS FORMULA $\lambda = 3,645.6 \frac{n^2}{n^2 - 4}$ ($a = 2$)

Line.		X (computed), angstroms	X (observed), angstroms
H _a	3	6,562 08	6,562 10 (Angstrom)
H _B	4	4,860 80	4,860.74 (Angstrom)
H _r	5	4,340.0	4,340 10 (Angstrom)
H ₈	6	4,101.3	4,101.2 (Angstrom)
H _E	7	3,969.7	3,968 1 (Huggins)
H _f	8	3,888 6	3,887 5 (Huggins)
H _n	9	3,835.0	3,834 0 (Huggins)
H _o	10	3,797 5	3,795 0 (Huggins)
H _i	11	3,770.2	3,767.5 (Huggins)

of the formula with the measured wave lengths. The agreement is seen to be excellent in the visible spectrum. The discrepancy between Balmer's computed values and the measurements of Huggins increases, however, to nearly 1 part in 1,000 for H_i . Balmer questioned whether this discrepancy indicated that the formula was only an approximation or whether the data were in error. Recent measurements have considerably revised Huggins' data but have also revealed the need for a slight correction to Balmer's formula [*cf.* Sec. 136(b)].

Balmer correctly predicted that in this series of lines in hydrogen no lines of longer wave length than **H_a** would be found and that the

¹ *Ann. d. Physik*, vol. 25, p. 80 (1885).

series should "converge" at $\lambda = 3,645.6$ Å, since the fraction $\frac{n^2 - 4}{n^2}$ approaches unity as n becomes large. In Fig. 50 is reproduced a

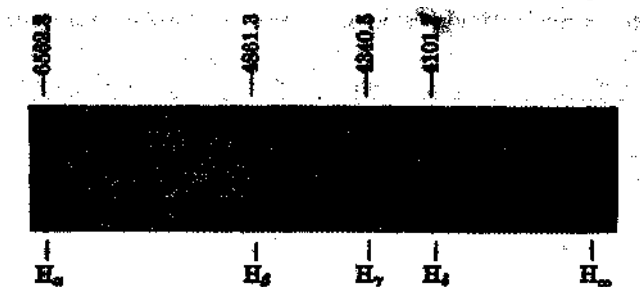


FIG. 50.—The Balmer series of atomic hydrogen, in emission. H_∞ marks the theoretical position of the series limit. [Photograph by G. Herzberg, *Ann. d. Physik*, vol. 84, p. 565 (1927); reprinted by courtesy of Prentice-Hall, Inc., New York, from G. Herzberg, "Atomic Spectra and Atomic Structure" (1927).]

photograph of the first few lines of the Balmer Series; in Fig. 51 is shown another photograph of the same series starting from the seventh line, more strongly exposed so as to bring out more lines.

The impetus which Balmer's discovery gave to work in spectral series is another illustration of the highly convincing nature of relations which are expressible in *quantitative* form. Soon after the



FIG. 51.—Higher members of the Balmer series, in emission, starting from the seventh line and showing the continuum. {After G. Herzberg, see credit under Fig. 50.)

publication of Balmer's work, intensive investigations in spectral series were initiated by Kayser and Runge and by Rydberg. We shall discuss next some of the results that were obtained by Rydberg.

88. Spectral Series and Their Interrelations.—In announcing his discovery, Balmer raised the question as to whether or not his formula might be a special case of a more general formula applicable to other series of lines in other elements. Rydberg set out to find such a formula.¹ Using the comparatively large mass of wave-length data then available and starting from the above-mentioned work of Liveing and Dewar, Rydberg isolated other series of doublets and triplets of constant frequency difference, according to Hartley's law of con-

¹ A brief account of Rydberg's work is given by him in *Phil. Mag.*, vol. 29, p. 331 (1890). For a fuller account see E. C. C. Baly's "Spectroscopy."

stant wave-number separation. In all cases, these series showed a tendency to converge to some limit in the ultraviolet. He found that he could distinguish two types of such series: a type in which the lines were comparatively sharp, which he called, therefore, *sharp*

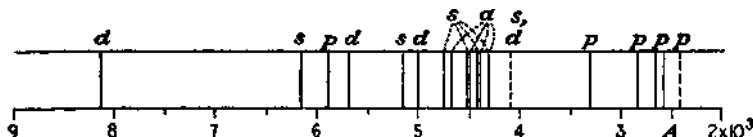


FIG. 52.—Principal lines in the spectrum of the neutral sodium atom: *p* = principal, *s* — sharp, *d* = diffuse series. Dotted lines indicate series limits. The first line of the sharp series, in the infrared at 11,393 Å, is not shown.

series; and a type which, because the lines were comparatively broad, he called *diffuse* series. Both types of series occurred in the arc spectrum of the same element. In many arc spectra, he found also a

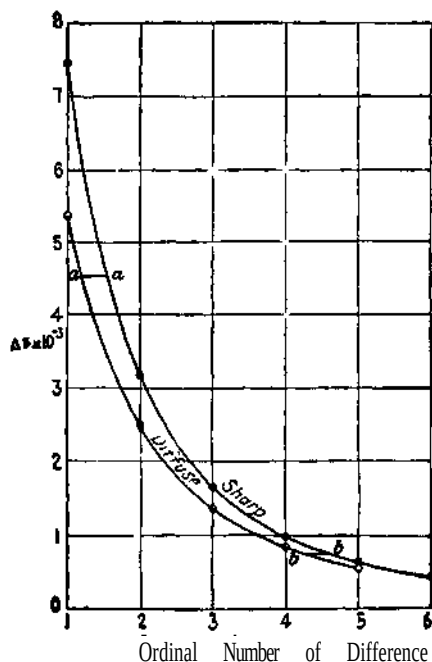


FIG. 53.—Wave-number differences between successive lines in the sharp and the diffuse series of sodium.

He observed that a similarity existed not only among the several series of a given element but also between series of different elements. He then showed that this similarity was *quantitative*. For example, in Fig. 53, the *differences* in wave numbers between successive lines in the sharp series of sodium, and similar differences

the doublet or triplet spacing decreased as the frequency or ordinal number of the line increased, as if tending to vanish at the convergence limit of the series; these he called *principal* series, because they commonly contained the brightest and most persistent lines in the spectrum. The chief lines of the principal, sharp, and diffuse series of sodium are plotted in Fig. 52 and marked *p*, *s*, *d*, respectively.

In searching for a formula, Rydberg plotted the wave numbers of the lines, or their differences, in various ways. By a "line" in this connection is meant a complex of actual lines forming a doublet or triplet,

which is seen as a single line in a spectroscope of moderate resolving power.

for the diffuse series of the same element, are plotted as ordinates, represented by the circles, against the *ordinal number* of the difference as abscissas. The resolving power is supposed to be limited so that each doublet appears as a single line. It is found that exactly parallel smooth curves can be drawn through the observed points; the two curves would coincide if the one for the diffuse series were displaced toward the right by a certain distance δ , equal to the length of the line aa or bb . Thus, we can say that the values of the successive differences for the sharp series are the same function of $(m + \delta)$ as the corresponding differences for the diffuse series are of m , where m is the ordinal number of the difference. Let us write ν_s , ν_d for the wave numbers of sharp and diffuse lines, respectively, and $\Delta \tilde{\nu}_s$, $\Delta \tilde{\nu}_d$ for the successive differences. Then, if

$$\Delta \tilde{\nu}_d = f(m), \quad \Delta \tilde{\nu}_s = f(m + \delta). \quad (123a,b)$$

A similar relationship must exist, then, for ν_d and $\tilde{\nu}_d$ themselves. Let us write for the m th line of each series

$$\tilde{\nu}_{dm} = F_d(m), \quad \tilde{\nu}_{sm} = F_s(m). \quad (124a,b)$$

Then, from (124a and 123a),

$$\tilde{\nu}_{d,m+1} - \tilde{\nu}_{dm} = F_d(m + 1) - F_d(m) = f(m), \quad (125)$$

provided the ordinal numbers of the differences are properly chosen. Similarly, for the sharp series,

$$\tilde{\nu}_{s,m+1} - \tilde{\nu}_{sm} = F_s(m + 1) - F_s(m) = f(m + \delta). \quad (126)$$

But, if we regard m as continuously variable in (125) and replace m by $m + \delta$, we have

$$f(m + \delta) = F_d(m + 1 + \delta) - F_d(m + \delta).$$

Hence, from (126),

$$F_s(m + 1) - F_s(m) = F_d(m + 1 + \delta) - F_d(m + \delta),$$

or

$$F_s(m + 1) - F_d(m + 1 + \delta) = F_s(m) - F_d(m + \delta).$$

Replacing m by $m + 1$ in this equation, we find also that

$$F_s(m + 2) - F_d(m + 2 + \delta) = F_s(m + 1) - F_d(m + 1 + \delta)$$

so, by the same equation,

$$F_s(m + 2) - F_d(m + 2 + \delta) = F_s(m) - F_d(m + \delta),$$

and, continuing,

$$F_s(m+n) - F_d(m+n+\delta) = F_s(m) - F_d(m+\delta)$$

for any integral n . Hence, replacing $m+n$ by m simply, we can write, for any m ,

$$F_s(m) - F_d(m+\delta) = C,$$

where C is independent of m , or

$$F_s(m) = C + F_d(m+\delta). \quad (127)$$

Thus, if we know $F_d(m)$ and δ , we can at once write down $F_s(m)$.

The form of F_d could only be discovered by a judicious guess. Rydberg found that many observed series could be represented closely by an equation of the form

$$\bar{\nu}_m = \bar{\nu}_\infty - \frac{R}{(m+\mu)^2} \quad (128)$$

u and v_w being constants which vary from one series to another. Obviously, by properly choosing the ordinal number m of the lines, u can always be made less than 1 in such a formula. The constant v_m represents the high-frequency limit to which the lines in the series ultimately converge. The Balmer formula is a special case of this more general "Rydberg formula," as it is called; for Eq. (122) can be written

$$\bar{\nu} = \frac{1}{\lambda} = \frac{1}{b} - \frac{4}{bm^2} = \bar{\nu}_\infty - \frac{R}{m^2},$$

with $v_\infty = 1/b$, $R = 4/b$, $m = 3, 4, 5, \dots$. This is of the type of (128) with $u = 0$.

The constant R in Eq. (128) was found to be the same for a large group of series for each substance, and very nearly the same for all substances. The slight variation of R from one atom to another is now known to be connected with differences in the atomic weight, the effect of which can be calculated theoretically (cf. Sec. 94). It is thus possible to calculate what the value of R would be for an atom of infinite weight. This latter value of R is denoted by R_8 ; the value for a particular kind of atom may then be indicated by a subscript. According to Birge¹ the probable value of R , obtained in this manner from spectroscopic observations as represented by Balmer's formula with a slight correction [Sec. 136(6)], is:

For hydrogen.....	R_H	= 109,677.58 cm. ⁻¹
For helium.....	R_{He}	= 109,722.26 cm. ⁻¹
For infinite atomic weight.....	R_∞	= 109,737.30 cm. ⁻¹

¹ BIRGE, *Rev. Modern Physics*, vol. 13, p. 233 (1941).

The value of R for the particular atom under discussion is known as the *Rydberg constant* for that atom.

89. Further Relationships between Series. Spectral Terms.—In addition to the qualitative similarities between spectral series already described, Rydberg discovered quantitative relationships between them. For one thing, he noted that the *sharp and the diffuse series appeared to have a common convergence limit*. Next, he observed that this common limit was equal to a *term* in the formula for the *principal series*, viz., to the term $R/(m + \mu)^2$ with m set equal to 1; and that a similar equality held between the *limit* of the *principal series* itself and the variable *term*, for $m = 1$, in the formula for the *sharp series*.

Because of these remarkable relations, the formulas for the three series in question could be written in the following form:

$$\tilde{\nu}_p = \frac{R}{(1 + S)^2} - \frac{R}{(m + P)^2}, \quad m = 1, 2, \dots \quad (129a)$$

$$\tilde{\nu}_s = \frac{R}{(1 + P)^2} - \frac{R}{(m + S)^2}, \quad m = 2, 3, \dots \quad (129b)$$

$$\tilde{\nu}_d = \frac{R}{(1 + P)^2} - \frac{R}{(m + D)^2}, \quad m = 2, 3, \dots \quad (129c)$$

Here we have written P, S, D for the values of u that occur in the variable terms of the formulas for the principal, sharp, and diffuse series, respectively; and we have indicated that, at least for the alkali metals, m starts from 1 in the principal series but from 2 in the others.

Two additional relations involving wave numbers of lines are at once apparent from the formulas. If we set $m = 1$ in the formula for ν_2 , we obtain the same number as the value of ν_p for $m = 1$, but with reversed sign! Furthermore, the difference between the limit of the principal series and the common limit of the other two is just the wave number of the first line of the principal series (the "Rydberg-Schuster" law, specifically enunciated by Rydberg in 1896 and independently in the same year by Schuster).

The structure of these formulas suggested to Rydberg the possibility that the first term on the right might also vary, in the same manner as does the second, thus giving rise to additional series of lines; for example, we might expect to find a series represented by the formula

$$\tilde{\nu} = \frac{R}{(2 + S)^2} - \frac{R}{(m + P)^2}, \quad m = 3, 4, \dots$$

Lines or series of this sort were actually discovered later by Ritz. Such lines are called *intercombination lines* or *series*, and the possi-

bility of their occurrence is known as the *Ritz combination principle*. Many examples of them are now known.

The most significant features about atomic spectra in general thus seem to be the following:

1. *The wave number of each line is conveniently represented as the difference between two numbers. These numbers have come to be called terms.*

2. *The terms group themselves naturally into ordered sequences, the terms of each sequence converging toward zero.*

3. *The terms can be combined in various ways to give the wave numbers of spectral lines.*

4. *A series of lines, all having similar character, results from the combination of all terms of one sequence in succession with a fixed term of another sequence. Series formed in this manner have wave numbers which, when arranged in order of increasing magnitude, converge to an upper limit.*

In the further development of spectroscopy, the spectral terms have tended increasingly to become the primary object of study. In the analysis of a new spectrum, one of the first steps is to represent the wave numbers of all of the observed lines as differences between terms, using as few terms as possible. Rydberg's approximate formulas for the spectral frequencies result from the use of an approximate formula for the terms of the type

$$\frac{R}{(m + \mu)^2},$$

c being a constant.

The simple picture presented here, however, requires considerable extension in order to be adequate even regarding the spectra emitted by single atoms. For one thing, in writing Rydberg formulas for spectral "lines," we have ignored the fine structure of the lines, by means of which series were first picked out; *singlet* series, in which each line is single, are known in many elements, but more commonly the lines form *doublets*, *triplets*, or groups of even more components. In such cases, a separate Rydberg formula must be written for each component line; in the spectra of the alkali metals, for example, doublets occur, so that six formulas instead of three are required for a complete representation of the chief series. Furthermore, the Rydberg formula itself is only a first approximation. The whole subject can be much better understood when the discussion can be centered about the atomic mechanism by which spectral lines are emitted. Accordingly, further discussion of details will be postponed to a later chapter.

The remarkable properties possessed by spectral series pointed to the existence within atoms of a comparatively simple and universal mechanism by which spectra are emitted. In terms of classical ideas, however, it was very difficult to imagine a mechanism which could emit spectra having the observed features. It was natural to assume that the higher members of a series were of the nature of overtones. Among acoustic vibrations, many cases are met with in which the frequencies of the overtones are not integral multiples of the fundamental frequency; examples are the vibrations of bells and of the common tuning fork. But no cases are known in which the frequencies of the overtones converge to an upper limit. Furthermore, the Ritz combination principle is without analogy in the classical theory of vibrating systems.

The key to the origin of spectral lines was not discovered until certain other lines of evidence had led to the adoption of radically new conceptions concerning the structure of atoms.

90 Early Views on Atomic Structure.—Speculations as to the structure of the atom date from the early years of the nineteenth century. In 1815, Prout proposed a hypothesis asserting that *all elements are made up of the atoms of hydrogen as a primordial substance*. The hypothesis was based on the fact that the atomic weights of a large number of the elements are very nearly simple multiples of that of hydrogen, and Prout had no further data to support his views. Accordingly, when more accurate determinations showed that, in general, atomic weights were *not* exact integral multiples of the atomic weight of hydrogen and that there were such notable exceptions as chlorine, with an atomic weight of 35.5, Prout's hypothesis was abandoned—to be revived again decades later in a new form and, of course, on the basis of newly discovered *experimental* evidence (see Chap. XI).

With the discovery of the electron by Sir J. J. Thomson, in 1897, theories of atomic structure began to assume a more definite form, since it became obvious then that the atom must be made up of numerically equal quantities of negative and of positive charges. Two questions of importance then arose: (1) How many electrons are there in atoms? and (2) How are the electrons and the positive charges in the atom arranged? Two independent lines of evidence gave an answer to the first of these two questions.

1. On the basis of classical theory, Sir J. J. Thomson showed that, when a beam of X-rays passes through matter, it should be scattered, the scattering coefficient σ being given by

$$\sigma = \frac{8\pi}{3} \frac{e^4}{m^2 c^4} N,$$

where e and m are, respectively, the charge and mass of the electron, c is the velocity of light, and N is the number of electrons per unit volume [Sec. 172(6)]. Knowing N from the experimentally determined value of σ and knowing the number of atoms per unit volume of the scattering material, the number of electrons *per atom* could be computed. From early measurements of σ , Thomson¹ computed that the number of electrons per atom was of the order of the atomic weight; but, from later measurements,² Barkla concluded that the number of electrons per atom is, for the lighter atoms at least (except hydrogen), more nearly *one-half the atomic weight* [see Sec. 172(6)]. The carbon atom, for example, with atomic weight 12, was found in this way to have 6 electrons. It was later shown that the scattering coefficient for hydrogen is such as to indicate that it has only 1 electron per atom.

2. A stream of electrons traveling with very high velocity is able to pass through thin sheets of matter but, in so doing, is diffused, the electrons being deflected, more or less, from their initial direction as a result of the electrostatic forces acting on the moving electrons when they pass near the electrons in the atoms of the thin sheet. Comparing the computed value of the diffusion with that observed experimentally, Thomson concluded that the number of electrons per atom should be of the order of the atomic weight.

These conclusions were corroborated in a general way by calculations in regard to the dispersion of light by monatomic gases. The number of electrons in an atom was thus not determined definitely, but its order of magnitude was indicated. Since the normal atom is electrically neutral, the quantity of *positive* electricity per atom was thus roughly determined as well. Now, Thomson had shown that the mass of the electron is of the order of one two-thousandth of the mass of the hydrogen atom. It was obvious, therefore, that the mass of the atom is vastly greater than the total mass of the comparatively few electrons which it contains; and it was logical to assume that practically the entire mass of the atom is associated with its positive charge.

The problem as to the *arrangement* of the electrons and of the positive charge in the atom then came to the fore. On the basis of classical ideas, it seemed obvious that this arrangement must be such as to meet two conditions. (1) The ensemble of positive charge and negative electrons which make up the atom must be stable; the electrons, for example, must be held by (electrostatic?) forces in fixed

¹" Electron Theory of Matter," p. 145.

²*Phil. Mag.*, vol. 21, p. 648 (1911).

positions of equilibrium about which they may vibrate, when disturbed, with the definite frequencies required to explain the characteristic line spectra of the elements. (2) Except when so disturbed, the electrons must be at rest, since otherwise they would emit radiation as required by the electromagnetic theory. The much greater mass of the positive charge made it reasonable to assume that it is the electrons, rather than the positive charges, which vibrate in the process of emitting radiation.

A possibility considered by J. J. Thomson was that the positive electricity might be distributed continuously throughout a certain small region, perhaps with uniform density throughout a sphere. The electrons might then be embedded in the positive electricity, occupying normally certain positions of equilibrium, and executing harmonic vibrations about these positions when slightly disturbed. Frequencies in the visible spectrum might thereby be emitted if the sphere of positive electricity were of the order of 10^{-8} cm. in radius. Thomson was unable to show, however, that these frequencies might be such as to form a series converging to an upper limit, and eventually his theory came into conflict with the experiments of Rutherford and his collaborators on the scattering of α particles, which are now to be described.

91. The Scattering of Alpha Particles by Atoms.—The α -rays from radioactive materials have been shown to be positively charged particles which have (1) a mass almost exactly equal to that of the helium atom and (2) a positive charge numerically equal to twice the charge on the electron. Since helium is known to be produced by radioactive substances which emit α -rays, the latter are identified with helium atoms which have lost 2 electrons (*cf.* Sec. 198 in Chap. XI). The initial velocity of the α particles, although depending somewhat on the radioactive material from which they originate, is of the order of 2×10^9 cm. per second. These particles can be studied by means of the flashes of light or scintillations which they produce on striking a zinc sulfide screen, the impact of a *single* particle producing a visible flash. These flashes are readily observed under a low-power microscope.

If a stream of α particles, limited by means of suitable diaphragms to a narrow cylindrical pencil, be allowed to strike a zinc sulfide screen placed at right angles to the path of the particles, the scintillations will occur over a well-defined circular area equal to the cross section of the pencil. If, now, a thin film of matter, such as gold or silver foil, is interposed in the path of the rays, it is found that they pass quite readily through the foil, but that the area over which the scintillations occur becomes larger and loses its definite boundary, indicating that some of the particles have been deflected from their

original direction. This spreading out of the stream of particles on passing through thin layers of matter, solid or gaseous, is called "scattering."

Qualitatively, it is easy to explain the origin of the forces which cause the deflection of the α particle. The particle itself has a twofold positive charge. The atoms of the scattering material contain charges, both positive and negative. In its passage through the scattering material, the particle experiences electrostatic forces the magnitude and direction of which depend on how near the particle happens to approach to the centers of the atoms past which *or through which* it moves.

If we assume the Thomson model of the atom, with its sphere of positive electrification, inside which are electrons, the path of an α particle in passing *through* such an atom might be as indicated in Fig. 54, the net result of the passage being to deflect the path of the particle

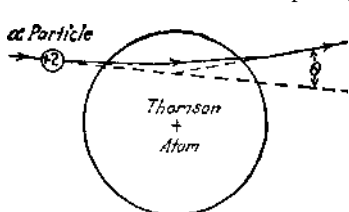


FIG. 54.—Deflection of an α particle by a Thomson atom.

through a small angle θ . The major part of this deflection arises from the electrostatic repulsion on the α particle due to the charge on the positive sphere, which, for the heavier atoms at least, has a mass many times that of the α particle. The electrons

within the positive sphere, being capable of motion about their respec-

tive positions of equilibrium and possessing a mass which is very small compared with the mass of the α particle, will produce no appreciable deflection of the latter. Rather, the electrons themselves would be pulled from their positions of equilibrium and set vibrating or even expelled from the atom by the passage of the α particle. The total deflection of any given particle in passing through or past a number of such atoms in a thin layer of scattering material will be governed by the laws of probability. Thomson showed¹ that the mean deflection ϕ_m of a particle in passing through a thin plate of thickness t should be

$$\phi_m = \theta \sqrt{N\pi a^2 t},$$

where θ is the average deflection due to a single atom, N is the number of atoms *per unit volume*, and a is the radius of the positive sphere.

This process of scattering of the α particles as a result of a large number of small deflections produced by the action of a large number of atoms of the scattering material on a single α particle is called

¹ *Cambridge Phil. Soc. Proc.*, vol. 15, p. 465 (1910).

multiple, or *compound*, scattering. It is readily seen that the structure of the atom assumed by Thomson would not result in a large deflection due to any *single* encounter. According to Rutherford, the number of a particles N_ϕ which, as a result of *multiple* scattering, should be scattered by such an atom through an angle ϕ or greater is given by

$$N_\phi = N_0 e^{-(\phi/\phi_m)^2},$$

where N_0 is the number of particles for $\phi = 0$, and ϕ_m is the average deflection after passing through the scattering material.

Now Geiger had shown¹ experimentally that the most probable angle of deflection of a pencil of a particles in passing through gold foil 1/2,000 mm. thick is of the order of 1° . It is evident, therefore, from the last equation, that the probability for scattering through large angles becomes vanishingly small; for 30° , for example, it would be of the order of 10^{-13} . Geiger showed that the *observed* scattering obeyed this probability law for *small* angles of scattering but that *the number of particles scattered through large angles was much greater than the theory of multiple scattering predicted*. Indeed, Geiger and Marsden showed² that 1 in 8,000 a particles was turned through an angle of *more than* 90° by a thin film of platinum, *i.e.*, was, *in effect*, diffusely reflected. This so-called "reflection" however, was shown to be not a surface phenomenon but rather a volume effect, since the number of particles turned through more than 90° increased, up to a certain point, with increasing thickness of the scattering foil. It was also found that the proportion of particles diffusely reflected increased approximately as the $3/2$ power of the atomic weight of the foil.

It was impossible to explain this excess scattering of a particles at large angles on the basis of multiple scattering by a Thomson atom and the laws of probability. There must be something wrong, then, with Thomson's picture of the atom.

92' Rutherford's Nuclear Atom.—(a) *Rutherford's Hypothesis.*—Accordingly, Rutherford, in a classic article³ which may be regarded as the starting point of our modern ideas on atomic structure, proposed a new type of atom model capable of giving to an a particle a *large* deflection as a result of a *single* encounter. He assumed *that the positive charge of the atom, instead of being distributed continuously throughout a region of atomic dimensions, is concentrated in a very small*

¹ Roy. Soc, Proc, vol. 83, p. 492 (1910).

² Ibid., vol. 82, p. 495 (1909).

³ Phil. Mag., vol. 21, p. 669 (1911). Every student of modern physics should read and thoroughly digest this article.

region less than 10^{-12} cm. in diameter. This concentrated charge, later called the "nucleus," was assumed to be surrounded by the electrons in some sort of configuration.

The difference in the effect which the two atoms, Thomson's and Rutherford's, have on an α particle passing in their neighborhood is shown in Fig. 55. In (a) is represented diagrammatically the path of an α particle through Thomson's atom, the initial path of the

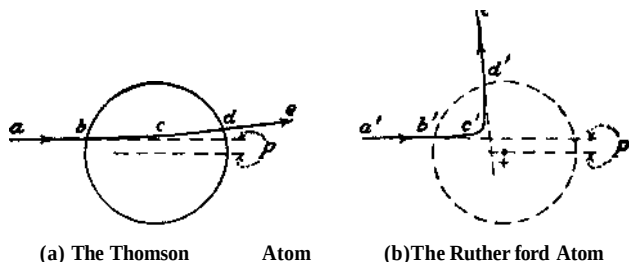


FIG. 55.—Comparison of deflections of an α particle by (a) the Thomson atom and (b) the Rutherford nuclear atom, for similar conditions of incidence.

particle being so directed that, if it were not deflected, it would pass through the atom at a distance p from its center. The force which the particle experiences as it penetrates more deeply into the Thomson atom becomes less and less. At point c , the force, although at right angles to the path, is a minimum. The deflection produced, equal to the angle between ab and de , is small. In Rutherford's model [Fig. 55(b)] we have a different state of affairs. Over the path $a'b'$,

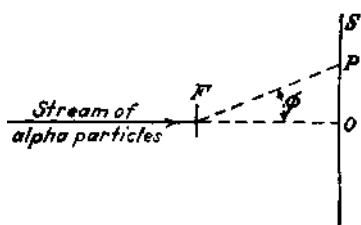


FIG. 56.

the forces which the particle experiences are the same as for the corresponding path in (a). After passing point b' , however, the forces continue to increase according to the inverse square of the distance between the particle and the nucleus, instead of decreasing as was the case in (a). The difference in the forces

experienced by the α particle in the two cases becomes very great as the particle approaches the nucleus. It is readily seen that the net result of all this is that the Rutherford model gives to the particle a much greater deflection than the Thomson model.

Rutherford calculated mathematically the distribution of the particles to be expected, as the result of *single scattering* processes by atoms of the type assumed. The scattering is usually measured by allowing the particles, after passing through the foil F (Fig. 56), to

fall on the zinc sulfide screen *S* placed normal to the initial path of the particles. By means of a low-power microscope, the number of particles striking the screen in the neighborhood of *P* is observed for various angles of scattering ϕ . Rutherford showed that the number of particles per unit area striking the screen should be proportional (1) to $1/\sin^4 \frac{\phi}{2}$; (2) to the thickness *t* of the scattering material, for small values of *t*; (3) to $(Ze)^2$, where *e* is the charge on the electron and *Z* is an integer depending on the kind of atom; and (4) inversely proportional to the square of the initial kinetic energy of the α particles.

(b) *Experimental Confirmation*.—These predictions were completely verified by the experiments of Geiger and Marsden.¹ Their data are

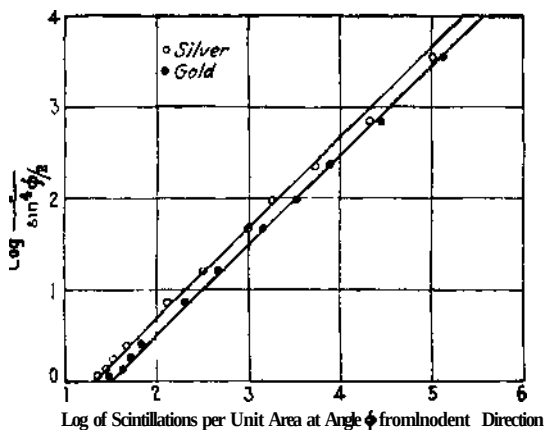


FIG. 57.—The law of the scattering of α particles.

shown graphically in Fig. 57, in which the logarithm of the number of scintillations on the screen per minute is plotted as abscissa against the logarithm of $1/\sin^4(\phi/2)$ ordinate. If these two quantities are proportional to each other, the points for each substance should lie on a straight line inclined at 45° with the axes. The two lines in the figure are drawn at exactly 45° , and the observed points are seen to agree well with the predictions. This is the more remarkable since the numbers of scintillations varied in the experiment over a very wide range of values, the left-hand points on the plots representing 22 particles per minute for silver and 33 for gold, whereas the right-hand points represent 105,400 and 132,000, respectively.

The prediction of Rutherford that the scattering should, for small thicknesses, be proportional to thickness was confirmed by the observa-

¹ GEIGER and MARSDEN, *Phil. Mag.*, vol. 25, p. 605 (1913),

tions. It will be recalled that, on the theory of multiple scattering by the Thomson atom, the *square root* of the thickness is involved. Geiger and Marsden also showed, in confirmation of the fourth of Rutherford's predictions, that "the amount of scattering by a given foil is approximately proportional to the inverse fourth power of the velocity (inverse square of the energy) of the incident α particles," over a range of velocities such that the number of scattered particles varied as 1:10.

(c) *Atomic Number*.—Geiger and Marsden further concluded, from a study of the variation of scattering with atomic weight and of the fraction of the total number of incident particles scattered through a given angle, (1) that the scattering is approximately proportional to the atomic weight of the scatterer over the range of elements from carbon to gold, and (2) that "*the number of elementary charges composing the center of the atoms is equal to half the atomic weight*." This second conclusion was in agreement with Barkla's experiments on the scattering of X-rays previously mentioned (Sec. 90), which determined the number of *electrons* associated with an atom. According to these results, carbon, nitrogen, and oxygen should have, respectively, 6, 7, and 8 electrons, around a nucleus containing, in each case, an equal amount of positive charge. Now, these elements are, respectively, the sixth, seventh, and eighth elements in the periodic table. The hypothesis was natural, therefore, that the number of electrons in the atom, *or the number of units of positive charge on its nucleus*, is numerically equal to the ordinal number which the atom occupies in the series of the elements, counting hydrogen as the first. This assumption gives to hydrogen 1 electron, and an equal charge on its nucleus, in agreement with the data on the scattering of X-rays by hydrogen. Helium would then have 2 electrons and a twofold positive charge on the nucleus; and we see, therefore, that the α particles, which have been shown to be helium atoms with a twofold positive charge, are helium nuclei. Lithium, the third element, should have 3 electrons and a threefold positive charge on the nucleus. Neon, the tenth element, should have 10 electrons; and so on. Thus originated the concept of *atomic number*, the importance of which was soon to be emphasized by the pioneer work of Moseley in the X-ray spectra of the elements. The *atomic number*, symbol Z , of an element we may think of variously as (1) the ordinal number of the element in the series of the elements starting with $Z = 1$ for hydrogen, or (2) the positive charge carried by the nucleus of the atom, in terms of the electronic charge e as a unit, or (3) the number of electrons surrounding the nucleus in the normal, neutral atom.

These experiments of Geiger and Marsden so completely confirmed the conclusions which Rutherford had reached by postulating the nuclear type of atom that, in spite of certain weighty objections, the Rutherford atom model was at once universally adopted.

(d) *Some Difficulties.*—The objections to the nuclear type of atom are, in large part, nonced with questions of stability. It is obvious that equilibrium cannot be secured, if the electrons are at rest, by the operation of electrostatic forces alone. For, consider [Fig. 58(a)] a nucleus with a dotible positive charge $+2e$ and with 2 electrons symmetrically placed at a distance r from the nucleus. Assuming the inverse-square law, the attraction of the nucleus for each electron is $2e^2/r^2$, while the electrons repel each other with a force equal to only one-eighth of this, viz., $e^2/(2r)^2$. The electrons will, therefore, "fall into" the nucleus.

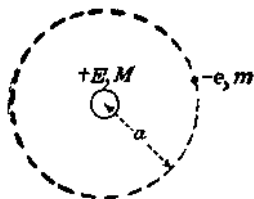


FIG. 58a.

FIG. 58b.

Something is gained by giving to the electron an orbital motion, such as that of the earth round the sun, but not enough. Let a nucleus [Fig. 58(6)] have a charge $+E$ and a mass M , and let a single electron, charge $-e$, mass m very small compared with M , revolve around it in a circular orbit of radius a at such velocity v that

$$\frac{Ee}{r^2} = \frac{mv^2}{r}.$$

Newton's mechanical laws of motion are then satisfied. But, according to the fundamental laws of the electromagnetic theory, such a revolving electron, since it is subject to a constant acceleration toward the center of its orbit, should *radiate energy*. This energy can come only from the energy of the system. The system will, therefore, "run down"; the electron will approach the nucleus along a spiral path and, as can easily be shown, will *give out radiation of constantly increasing frequency*. This, however, does not agree with the observed emission of spectral lines of fixed frequency.

It was at this point that Bohr introduced his epoch-making theory of the structure of the atom and of the origin of spectra. His theory constituted an extension of Planck's theory of quanta to Rutherford's

nuclear atom, in an attempt, extraordinarily successful, both to remove the difficulties of the nuclear model and to explain the origin of the characteristic spectra of the elements.

93. The Bohr Theory of Atomic Hydrogen.—As we have seen in the last chapter (Sec. 84), the essential features of Planck's original quantum theory were two:

1. An oscillator can exist only in one of a number of discrete quantum states, and to each of these states there corresponds a definite allowed value of its energy.

2. No radiation is emitted while the oscillator remains in one of its quantum states, but it is capable of jumping from one quantum state into another one of lower energy, the energy lost in doing so being emitted in the form of a pulse or quantum of radiation.

Successful applications of the first of these assumptions had already been made in other fields, notably by Einstein and especially by Debye in the specific heat of solids, which will be discussed in a later chapter (Sec. 163, Chap. IX). Nicholson had also attempted to apply the theory to spectra, and with some success, but he was unable to make it yield a series of lines converging to a limit. Bohr discovered how to apply similar ideas to a hydrogen atom of the Rutherford type and succeeded in arriving at a theoretical formula for its spectrum that agrees with observation.¹

Bohr assumed that an electron in the field of a nucleus was not capable of moving along every one of the paths that were possible according to classical theory but was restricted to move along one of a discrete set of allowed paths. While so moving, he assumed that it did not radiate, contrary to the conclusion from classical theory (Sec. 35), so that its energy remained constant; but he assumed that the electron could jump from one allowed path to another one of lower energy and that, when it did so, radiation was emitted containing an amount of energy equal to the difference in the energies corresponding to the two paths.

As to the *frequency* of the emitted radiation, he considered several alternative hypotheses but finally adopted the same assumption that Planck had made for his oscillators. That is, if W_1 and W_2 are the energies of the atom when moving in its initial and final paths, respectively, the *frequency* of the radiation emitted is determined by the condition that

$$\nu = \frac{W_1 - W_2}{h}, \quad (130)$$

¹ BOHR, N., *Phil Mag.*, vol. 26, p. 1 (1913).

where h is Planck's constant (Sees. 84 and 85). This assumption had the additional advantage of agreeing with that made by Einstein in arriving at his highly successful photoelectric equation [Sec. 46, Eq. (48)]. For the latter reason, Eq. (130) came later to be known as the *Einstein frequency condition*.

Concerning the formulation of the condition that determines the allowed paths, Bohr was also in some doubt. While the electron remains in one of its "stationary states," as he called them, he supposed it to revolve in a circular or elliptical orbit about the nucleus, just as the earth revolves about the sun, in accordance with the classical laws of mechanics. But what fixes the size and shape of this orbit? Bohr showed that it would suffice to assume a certain relation between the frequency of the emitted radiation and the frequency of revolution of the electron in its orbit. In the end, however, he preferred to postulate that the orbit is a circle, with the nucleus at its center, and of

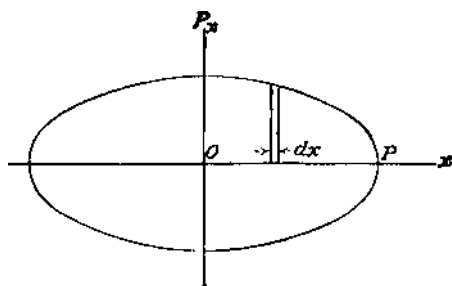


FIG. 51).

such size that the angular momentum of the electron about the nucleus is equal to an integral multiple of the natural unit $\hbar/2\pi$.

It was recognized later that the postulate in this form is in harmony with the final interpretation that Planck had found for h in the case of the harmonic oscillator. We saw in Sec. 84 that the state of the oscillator might be represented by a point on a plane, with the coordinate x and the momentum p of the oscillator taken as Cartesian coordinates, and that, as long as the energy of the oscillator remains constant, this point would move along an ellipse (Fig. 59). The condition for a stationary state or quantum state of the oscillator could be said to be that the area enclosed by this ellipse should be an integral multiple of h . For the enclosed area, we can obviously write $\oint p_x dx$, $\oint$ meaning integration over a complete cycle. While x ranges from its minimum to its maximum value, p_x is positive (above Ox in Fig. 59); whereas during the other half-cycle, as x decreases again so that $dx < 0$, p_x is negative, and $p_x dx$ is again positive. Thus a positive value is

obtained for the area. Planck's condition for a quantum state of the oscillator can, therefore, be stated as follows:

$$\oint p dx = nh, \quad (131)$$

where n stands for any positive integer.

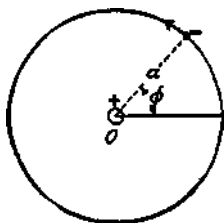


FIG. 60.

This rule can at once be modified so as to apply to the revolving electron. If we take, as its coordinate, its angular displacement from a fixed line, ϕ (Fig. 60), the (generalized) momentum p_ϕ corresponding to ϕ is the ordinary angular momentum. Equation (131) thus takes the form, for the revolving electron,

$$\oint p_\phi d\phi = nh.$$

But here p_ϕ is constant during a revolution, so that $\oint p_\phi d\phi = p_\phi \oint d\phi$ and $\oint d\phi = \int_0^{2\pi} d\phi = 2\pi$. Thus, from the last equation, we have, as the condition for a quantum state, Bohr's quantum condition:

$$p_\phi = n \frac{h}{2\pi}, \quad (132)$$

n being any positive integer.

Bohr's theory represents a remarkable combination of principles taken over from classical theory with postulates radically at variance with that theory. He solved the old problem of stability merely by *postulating* that the cause of instability, the emission of radiation and the accompanying radiation reaction, did not exist so long as the electron remained in one of its allowed stationary orbits. The electron could thus remain in its stationary state of lowest energy indefinitely, without spiraling down into the nucleus, as classical theory would require. But the problem of stability was solved only at the expense of throwing away the only picture we had of the mechanism by which the atom could radiate energy. For Bohr's postulates provide no picture of the sequence of events *during transitions between orbits*.

A hybrid theory of this sort was widely felt to be unsatisfactory; but we shall see in the next section that it was astonishingly successful (in the case of atomic hydrogen). In the next chapter, on wave mechanics (Chap. VII), we shall describe the theory that ultimately replaced the Bohr theory. This theory leads, at least approximately, to the same set of energy values for the stationary states or quantum states of the hydrogen atom as does the theory of Bohr, but it suggests quite a different picture of the behavior of the electron while in a quantum state. In the new theory, it is only about half true that the

electron is in motion in the atom, even when it is in one of its higher quantum states; at least, it cannot be said to follow a definite orbit. Because of the abstractness of the new theory, the original simple Bohr picture is commonly felt to retain something more than mere historical interest.

94. Quantum States of One Electron in an Atom.—The allowed values of the energy as given by Bohr's theory for an atom containing a single electron are easily deduced from his basic assumptions. For greater generality, let the nuclear charge be Ze , Z being the atomic number and, therefore, an integer, and e being the numerical electronic charge, in electrostatic units. Our calculations will then hold not only for the neutral hydrogen atom but also for a singly ionized atom of helium, for doubly ionized lithium, and so on.

It is assumed that the electron revolves in a circle with angular momentum

$$p_{\phi} = ma^2\omega = n\frac{\hbar}{2\pi}, \quad (133)$$

where n is a positive integer, called the "quantum number" of the state in question, m is the electronic mass, a the radius of the orbit, and ω the angular velocity of revolution. To satisfy the laws of mechanics, we must have the necessary centripetal force on the electron supplied by the force of attraction due to the nucleus, whence

$$ma\omega^2 = \frac{Ze^2}{a^2}. \quad (134)$$

Eliminating ω from these two equations, we find for the radius of the orbit

$$a = \frac{n^2\hbar^2}{4\pi^2me^2Z}. \quad (135)$$

The energy of the electron will be partly kinetic and partly potential. If we call the energy zero when the electron is at rest at infinity, its potential energy in the presence of the nucleus, according to the usual electrostatic formula (charge times potential), is

$$U = -\frac{Ze^2}{a}.$$

Its kinetic energy is

$$K = \frac{1}{2}mv^2 = \frac{1}{2}ma^2\omega^2 = \frac{1}{2}\frac{Ze^2}{a},$$

by (134). Thus $K = -\frac{1}{2}U$ a relation peculiar to motion under an inverse-square force. The total energy is $W = K + U$ or

$$W_n = -\frac{1}{2} \frac{Ze^2}{a} = -\frac{2\pi^2 me^4 Z^2}{h^2 n^2} \quad (136)$$

by (135). This is the energy of the atom when the electron is in its n th stationary or quantum state. Here and in Eq. (135), the quantum number n may take on any integral value: $n = 1, 2, 3, \dots$.

From Eq. (136), we see that, the larger the value of n , the *smaller numerically* but the larger in *algebraic value* is the energy of the system. The lowest value of W_n is that corresponding to the *first orbit*. This state is called the *normal* (quantum) state, or *ground state*, of the atom, since it should be the *most stable state and the one ordinarily occupied by the electron*.

It is interesting to note that according to Eq. (135) the radii of the successive allowed orbits are proportional to n^2 , to $1^2, 2^2, 3^2, 4^2, \dots$ or to $1, 4, 9, 16, \dots$. If in (135) we substitute $Z = 1$ for hydrogen and $h = 6.61 \times 10^{-27}$, $e = 4.803 \times 10^{-10}$, $m = 0.9107 \times 10^{-28}$ gram, and $n = 1$, we find for the radius of the smallest Bohr circle for hydrogen, or the radius of its orbit in the normal atom,

$$a = a_0 = \frac{h^2}{4\pi^2 me^2} = 5.27 \times 10^{-9} \text{ cm.} \quad (137)$$

The diameter of the orbit is thus close to 10^{-8} cm., which agrees very well with estimates of the atomic diameter obtained from kinetic theory. This is a first indication that the new theory may be able to explain, among other things, the apparent sizes of molecules. In atoms containing only 1 electron but a larger nuclear charge, the orbits are all smaller in the ratio $1/Z$.

It must be remarked, however, that our calculations contain a slight error. What we have actually developed is a *fixed-nucleus* theory. In reality, if the electron revolves in a circle about the nucleus, as we have supposed it to do, its path in space will be a circle about the *center of mass of the combined system*, which is slightly displaced from the nucleus (Fig. 61). At the same time the nucleus itself revolves about the center of mass in a smaller circle. If a is the distance of the electron from the nucleus, and a' , a'' the respective distances of electron and nucleus from the center of mass, then

$$a' = \frac{M}{m+M} a, \quad a'' = \frac{m}{m+M} a,$$

where M is the mass of the nucleus.



FIG. 61.

The total angular momentum about the center of mass is then the sum of that due to the electron and that due to the nuclear motion or

$$ma'^2\omega + Ma''^2\omega = \left[m \frac{M^2}{(m+M)^2} + M \frac{m^2}{(m+M)^2} \right] a^2\omega = \frac{mM}{m+M} a^2\omega.$$

Let us write

$$\checkmark m' = \frac{mM}{m+M}; \quad (138)$$

m' so defined is called the *reduced mass* of the electron. Then we can write for the total angular momentum $m'a^2\omega$, and, if we assume that Bohr's postulate has reference to the total angular momentum of the atom, we shall have, in place of (133),

$$m'a^2\omega = n \frac{h}{2\pi}. \quad (133')$$

Furthermore, in the left-hand member of the dynamical equation, (134), a , representing the radius of the electronic orbit in space, is now to be replaced by a' . The same result is obtained, however, if, retaining a , we replace m by m' as defined by Eq. (138). In the *right-hand* member of (134), however, a represents the distance between the electron and the nucleus and must be left unchanged. Thus in place of (134) we have

$$m'a\omega^2 = \frac{Ze^2}{a^2}. \quad (134')$$

The expression for the potential energy U is unaltered. The kinetic energy of the electron, however, will now be $\frac{1}{2}ma'W$, to which is to be added that of the nucleus, or $\frac{1}{2}Ma''^2\omega^2$; the total kinetic energy of the atom is thus

$$\begin{aligned} \left(\frac{1}{2}ma'^2 + \frac{1}{2}Ma''^2 \right) \omega^2 &= \frac{1}{2} \left[m \frac{M^2}{(m+M)^2} + M \frac{m^2}{(m+M)^2} \right] a^2\omega^2 \\ &= \frac{1}{2} \frac{mM}{m+M} a^2\omega^2 = \frac{1}{2} m' a^2\omega^2 = \frac{1}{2} \frac{Ze^2}{a}, \end{aligned}$$

by (134'). The total energy is thus

$$W_n = -\frac{1}{2} \frac{Ze^2}{a}$$

as before.

It is easily seen now that the effect of these changes is simply to replace m by m' as given by Eq. (138) in all of the equations previously

$$a = \frac{n^2 \hbar^2}{4\pi^2 m' e^2 Z}, \quad W_n = -\frac{2\pi^2 m' e^4 Z^2}{\hbar^2 n^2}.$$

$$\nu = \frac{W_1 - W_2}{h}.$$

$$\nu = \frac{2\pi^2 m e^4 Z^2}{\hbar^2} \left(\frac{1}{n_2^2} - \frac{1}{n_1^2} \right),$$

or, in terms of wave numbers,

$$\tilde{\nu} = RZ^2 \left(\frac{1}{n_2^2} - \frac{1}{n_1^2} \right), \quad R = \frac{2\pi^2 me^4}{ch^3}.$$

These expressions are positive, since necessarily $n_1 > n_2$.

For hydrogen ($Z = 1$), the latter formula becomes

$$\tilde{\nu} = R \left(\frac{1}{n_2^2} - \frac{1}{n_1^2} \right). \quad (141)$$

If we set $n_2 = 2$ and let n_1 take on any integral value from 3 up, this last formula is exactly of the type of Balmer's formula! It permitted at once a crucial quantitative test of the new theory. When Bohr inserted in it the best values of his day, $e = 4.7 \times 10^{-10}$, $e/m = 5.31 \times 10^{17}$ with e in electrostatic units, and $h = 6.5 \times 10^{-27}$ (as found from Planck's formula for black-body radiation), he found for the coefficient in (140a)

$$\frac{2\pi^2 me^4}{h^3} = 3.1 \times 10^{15}.$$

This, then, should be the value of the Rydberg constant R , when expressed in terms of frequency rather than wave number. As calculated from the frequencies of the lines of the Balmer series, its value was

$$3.290 \times 10^{15}.$$

Bohr considered that "the agreement between the theoretical and observed values is inside the uncertainty due to experimental errors in the constants entering in the expression for the theoretical value."

If we employ the best modern values, say $e = 4.803 \times 10^{-10}$, $c/m = 1.759 \times 2.9980 \times 10^{17}$ with e in electrostatic units, and $h = 6.61 \times 10^{-27}$, as obtained from independent measurements, we obtain, in reciprocal centimeters,

$$\frac{2\pi^2 me^4}{ch^3} = 1.105 \times 10^5 \text{ cm.}^{-1},$$

as against the best spectroscopic value¹ (for an atom of infinite mass):

$$R_\infty = 109,737.30 \text{ cm.}^{-1}$$

The difference is about 0.7 percent. This has been felt to be too large a difference to be due to errors in the experimental values of e , e/m , c , and h . It should be noted, however, that e and h enter into the formula

¹ BIBGE, *Rev. Mode Phys.*, vol. 13, p. 233 (1941).

in rather high powers, so that any errors in these constants become magnified. Furthermore, as the refinement of experimental methods has increased, the observed values of physical constants have shown in the past a general tendency to become changed by larger amounts than would have been expected in view of the precision of the methods employed, presumably owing to unsuspected systematic errors. The formula for R_∞ seems, on the other hand, to rest on a sound theoretical basis; the same formula is obtained on the more modern wave-mechanical theory as on Bohr's theory. Accordingly, Birge has recommended¹ that the value of h should be chosen so as to bring the theoretical formula for „ into agreement with the observed value, resulting in the value

$$h = 6.624 \times 10^{-27}.$$

In the present edition of this book, however, the value $h = 6.61 \times 10^{-27}$ will be retained in order to avoid the necessity of recalculating various other numbers.

The maximum possible value of ν in any given spectral series, for a 1-electron atom, is obtained from Eq. (1406) if we set $n_1 = \infty$. This gives the theoretical series limit,

$$\bar{\nu}_\infty = \frac{RZ^2}{n_2^2}.$$

For the corresponding atomic transition, we have $W_1 = 0$, the electron starting from a state of rest at infinity.

The range of possible transitions is not thereby exhausted, however. It can also be imagined that an electron moving along a hyperbolic orbit past a hydrogen nucleus might drop into one of the closed orbits that are allowed on the Bohr theory, emitting all of its excess energy in a single quantum. In this case, W_1 would have some *positive* value W and the frequency of the emitted photon would be

$$\bar{\nu} = \frac{W - W_2}{ch} = \bar{\nu}_\infty + \frac{W}{ch}.$$

Since W can vary continuously from 0 upward, such jumps would give rise to a continuous spectrum extending from $\bar{\nu}'$ toward shorter wave lengths. Such a continuum is clearly visible in Fig. 51 (Sec. 87).²

In spite of the radical nature of the assumptions underlying Bohr's

¹ *Ibid.*

² The continuum actually overlaps the series slightly in this figure. This may be due to some interference of the atoms with each other; in its higher quantum states, according to Eq. (135), the hydrogen atom must be effectively very large, e.g., 400 a_0 or 2×10^{-6} cm. for $n = 20$.

theory, its quantitative successes in explaining many facts quickly secured general acceptance of the theory. In his first paper in 1913, Bohr points out several other respects in which the predictions of his theory were in agreement with experiment. According to the theory, in the formulas for ν or ω the quantum number n_2 , as well as n_1 , might have any value. If we make $n_2 = 1$ and let n_1 vary, we obtain another series of lines, for which the final state is the normal state of the atom; $n_2 = 3$, with $n_1 > n_2$, gives us a third series, and so on. In Bohr's time a series of infrared lines corresponding to $n_2 = 3$, called the "Paschen series," was actually known. The series for $n_2 = 1$, however, lies in the ultraviolet and was not discovered until later; it is called, after its discoverer, the "Lyman series." Lines of other series corresponding to $n_2 = 4$ or 5 have also been observed.

[It is important to note clearly, however, which features of the theory are directly effective in determining the frequencies of the spectral lines. These frequencies are quite distinct from the frequency of orbital revolution of the electron, a feature of the new theory that was in most striking contrast with classical ideas. The frequency of orbital revolution can be found by solving (133) and (134) in the last section for ω ; it is

$$\nu_{\text{orb}} = \frac{\omega}{2\pi} = \frac{4\pi^2 m e^4 Z^2}{h^3} \frac{1}{n^3}.$$

For comparison, Eq. (140a) may be written thus:

$$\nu = \frac{4\pi^2 m e^4 Z^2}{h^3} \frac{n_1 + n_2}{2n_1^2 n_2^2} (n_1 - n_2).$$

It can easily be seen that, since $n_1 > n_2$,

$$\frac{1}{n_1^3} < \frac{n_1 + n_2}{2n_1^2 n_2^2} < \frac{1}{n_2^3}.$$

Hence, if $n_1 - n_2 = 1$, the frequency ν of the emitted radiation is intermediate between the frequencies of orbital revolution in the initial and final states. Only for very large n , for which the orbital frequencies in successive orbits become indistinguishable, do the emitted and orbital frequencies tend to coincide. Making in succession $n_1 - n_2 = 1, 2, 3, \dots$, we have an approximation to various harmonic overtones of a fundamental frequency. Thus, quantum jumps in which $\Delta n > 1$ correspond to the overtones in the case of classical vibrations. This is an example of a principle later elaborated by Bohr and known as the "correspondence principle."

In the present section we have discussed principally the spectrum of hydrogen, but similar results follow from Eq. (140a) or Eqs. (140c)

for any atom containing a single electron. As a further example, the spectrum of ionized helium will be discussed presently (Sec. 97).

96. The Spectrum of Atomic Hydrogen. Energy Levels and Spectral Series.—The energies of the lowest eight states of the hydrogen atom are shown in Table II. They are calculated from Eq. (136) with the observed value of the Rydberg constant R_H inserted in place of the coefficient $2\pi^2me^4Z^2/h^2$. It is often convenient to express such energies in wave-number units; the difference of two energies then gives at once the wave number of the corresponding spectral line. If W_e is an energy in ergs, its value W_z in wave-number units or cm.^{-1} is

$$W_z = \frac{W_e}{ch} \quad (143)$$

TABLE II.—ENERGY VALUES OF SOME STATES OF THE HYDROGEN ATOM

State	Energy		
	Wave-number units	Ergs	Electron-volts
$n = 1$	-109,678	-217.3×10^{-13}	-13.58
2	-27,420	-54.3	-3.394
3	-12,186	-24.2	-1.508
4	-6,855	-13.58	-0.849
5	-4,387	-8.69	-0.543
6	-3,047	-6.04	-0.377
7	-2,238	-4.44	-0.277
8	-1,714	-3.40	-0.212

For other purposes energies expressed in *electron-volts* are more convenient. An *electron-volt* (abbreviation, *ev*) is the work done on an electron, or its gain in energy, when it passes through a potential rise of 1 volt. Hence, a volt being $1/300$ electrostatic unit,

$$1 \text{ electron-volt} = \frac{e}{300} = 1.601 \times 10^{-12} \text{ erg}, \quad (144)$$

if $e = 4.803 \times 10^{-10}$ e.s.u. In Table II, the energies are given in all three units, ergs, cm.^{-1} , and electron-volts.

In spectroscopic work, however, it is more usual not to employ the energy itself, which is negative, but its numerical value. These numerical values of the atomic energies are called *term values* or *terms*. The wave number of a spectral line is then obtained by subtracting the term value for the *initial* state from that for the *final* state.

In Table III are shown the first four lines of each of the four known series in the spectrum of atomic hydrogen, and the associated term values. Column 4 gives the quantum number for each term. The first term value given in column 3 is in each case the convergence wave number or limit for that series; the wave number ν of each line is obtained by subtracting from the convergence limit the term opposite

TABLE III.—SOME TERM VALUER AND LINES FOR THE HYDROGEN ATOM

X, angstroms	ν , cm. ⁻¹	Terms, cm. ⁻¹	Quantum number, n
Lyman series		109,678	1
1,216 0	82,258	27,420	2
1,025 8	97,491	12,186	3
972.5	102,823	6,855	4
949 5	105,291	4,387	5
Balmer series		27,420	2
6,562 8	15,233	12,186	3
4,861 3	20,565	6,855	4
4,340 5	23,032	4,387	5
4,101 7	24,373	3,047	6
Paschen series		12,186	3
18,756	5,331	6,855	4
12,821	7,799	4,387	5
10,939	9,139	3,047	6
10,052	9,948	2,238	7
Brackett series		6,855	4
4 05M	2,468	4,387	5
2.63	3,808	3,047	6
2 16	4,617	2,238	7
1 94	5,141	1,714	8

the wave number of the line. The wave length X is calculated as $\nu^{-1} \times 10^8$. Many repetitions of term values necessarily occur in such a table.

The relations between the energies and the series of lines are much better seen from an *energy-level diagram*. In Fig. 62(a) in Sec. 98 is shown such a diagram, necessarily incomplete, **for** atomic hydrogen, Each horizontal line represents an **energy level**, **higher**

energies being plotted above. The line at $n = 1$ represents the normal state of the atom; the line at $n = \infty$ represents the electron at rest at infinity, the atom being just ionized; and above this is the continuum of positive energies for the free electron. Wave numbers are shown on a scale reading downward. Energies measured from the normal state as zero are also shown, expressed in electron-volts. A few of the transitions which give rise to lines are shown in an obvious manner by arrows. The diagram brings out clearly the fact that each series ends on a particular energy level.

97. Ionized Helium.—A helium atom which has lost both electrons is a bare nucleus and cannot radiate energy. One which has lost only a single electron, however, resembles a hydrogen atom, except that $Z = 2$ and the nucleus is nearly four times as heavy. The spectrum emitted by such atoms is known as the spark, sprriry. of helium, because it is emitted much more strongly when the helium is excited by a spark than when it is excited by an arc. The *arc spectrum* of helium, emitted by the neutral atom, will be considered later (Sec. 140).

Putting $Z = 2$ in Eq. (1406) we should have for the frequencies of the spectral lines emitted by ionized helium

$$\bar{\nu} = 4R \left(\frac{1}{n_2^2} - \frac{1}{n_1^2} \right). \quad (145)$$

Thus *ionized helium should emit the same spectrum as hydrogen except that all frequencies are four times as great, or all wave lengths a quarter as great.*

This conclusion from the theory agrees with observation except for a slight numerical discrepancy which becomes significant when measurements of precision are considered. As explained in Sec. 94, m in Eq. (136), or in formula (140c) for the Rydberg constant, must, strictly speaking, stand for the reduced mass m' of the electron as given by Eq. (138). Hence, the Rydberg constant for a 1-electron atom the nucleus of which has a mass M is, from (138) and (140c),

$$R = \frac{M}{m + M} R_{\infty}, \quad R_{\infty} = \frac{2\pi^2 m e^4}{c h^3}. \quad (146a,b)$$

In these latter formulas, m stands for the ordinary electronic mass, 0.911×10^{-27} gram.

One use that can be made of these relations is to deduce the ratio of the mass of the electron to the mass of a hydrogen or helium atom from values of R calculated from spectral data. If subscripts H, He

indicate values of R referring to hydrogen and helium, respectively, we have, from (146a),

$$\frac{R_{\text{He}}}{R_{\text{H}}} = \frac{M_{\text{He}}}{m + M_{\text{He}}} \frac{m + M_{\text{H}}}{M_{\text{H}}},$$

whence

$$\left(1 - \frac{M_{\text{H}} R_{\text{He}}}{M_{\text{He}} R_{\text{H}}}\right) \frac{m}{M_{\text{H}}} = \frac{R_{\text{He}} - R_{\text{H}}}{R_{\text{H}}}.$$

If we insert here, from Sec. 88, $R_{\text{H}} = 109,677.58$, $R_{\text{He}} = 109,722.26$, and $M_{\text{K}}/M_{\text{He}} = 1.0081/4.002$, we find

$$\frac{m}{M_{\text{H}}} = \frac{1}{1,837}.$$

That a value of m/M_{H} so found really possesses some precision, although it depends upon the difference of two numbers, R_{H} and R_{He} , which are almost equal, is due to the extraordinary precision of the spectroscopic values of R . We can then also proceed further and obtain a value of e/m for the electron. This is one of several spectroscopic methods for finding e/m . For

$$\frac{e}{m} = \frac{e}{M_{\text{H}}} \frac{M_{\text{H}}}{m},$$

and e/M_{H} equals the charge necessary to precipitate a gram-atom in electrolysis divided by the atomic weight of hydrogen, or 9,648.8 electromagnetic units divided by 1.0081. Thus,

$$\frac{e}{m} = \frac{9,648.8}{1.0081} \times \frac{1}{1,837} = 1.758 \times 10^7 \text{ e.m.u. gram}^{-1}.$$

This value of e/m agrees very closely with the results of recent measurements of the same quantity for *free* electrons, already mentioned in Sec. 42. Birge cites the average of four such measurements, made either by measuring the velocity imparted to electrons by a known potential drop or by the magnetic-deflection method, as 1.75955. The agreement of the values of e/m obtained by these two different methods constitutes a valuable check on the correctness of the assumptions underlying atomic theory. Furthermore, five other spectroscopic methods of evaluating e/m , listed by Birge, give values in good agreement with these two. The weighted average of all 10 values is

$$\frac{e}{m} = 1.7591 \times 10^7.$$

¹ BIRGE, *Phys. Rev.*, vol. 54, p. 972 (1938).

From the value just found for m/Mn we can calculate R_8 from the observed value of R_n by means of the formula, a special case of Eq. (146a),

$$R_{\infty} = \left(1 + \frac{m}{M_H}\right) R_H.$$

It is in this or an equivalent way that "experimental*" values of R are obtained.

In a similar way, *doubly ionized lithium* is found to emit the hydrogen spectrum with all frequencies multiplied (almost exactly) by 9; *trebly ionized beryllium* emits them increased in the ratio 16; and so on. The first line of the Lyman series for quadruply ionized boron ($Z = 5$) has been found by Edlen at 48.585 Å, with a frequency 25.04 times that of the first Lyman line of hydrogen.

98. Energy Levels and Series Relationships for Sodium.—Energy-level diagrams analogous to that for hydrogen can be constructed for all atoms. Such a diagram makes very evident a simple explanation in terms of the energy levels for those striking properties of spectral series and of their interrelations which were described above (Sees. 88 and 89). As an illustration we may return here to a further brief discussion of the spectrum of neutral sodium.

A partial energy-level diagram for sodium is shown in Fig. 62(6); it contains levels only for those quantum states that occur as the valence electron moves out toward infinity (for a more complete treatment of the sodium spectrum see Sees. 131, 132, 134). The levels are labeled on the diagram in modern spectroscopic notation, but this is immaterial for our present purpose. Most of the levels are really double, two levels lying very close together, but only one such pair is shown as such on the diagram (and on a much exaggerated scale), the others being represented by a single line each. (For the explanation of the doubling and of a similar but more minute fine structure of the hydrogen levels, see Sec. 136 below.) The levels labeled *S*, however, including the normal state, are strictly single.

A brief inspection of the level diagrams in Fig. 62 reveals the following explanation for the properties of the spectral series. That each series must converge to a finite limit arises from the simple fact that the largest energy difference possible between levels is limited to the energy required to ionize the atom, starting from the initial state for the series in question. The *principal* series of lines for sodium, including as its first member the ordinary *D* lines, arises from transitions ending on the normal state, which, as just stated, is single. The doublet character of the lines of this series is due, therefore, to the

doublet nature of the upper level; and, since the spacing of these levels decreases rapidly with increasing energy, the spectral doublets close up as their ordinal number increases. The *sharp* and the *diffuse* series both end on the doublet level next above the normal state; hence, they have a common convergence limit. The lines of these series, as

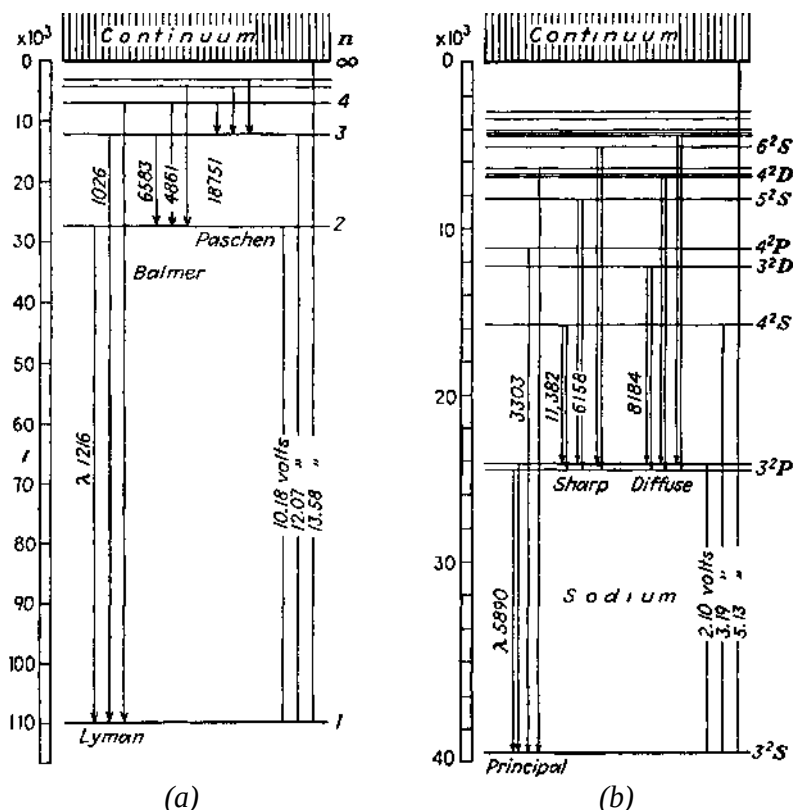


FIG. 62.—The lower energy levels of (a) hydrogen and (b) sodium atoms. Values at the left are in cm^{-1} ; a few differences are shown at the right in electron-volts. The transitions giving rise to the first three lines of the chief series are shown by arrows.

observed, usually appear to be doublets of constant frequency difference, this difference being that between the sublevels of the final state; hence, Hartley's law commonly appears to be obeyed by both series. (A further slight splitting of the *diffuse* lines will be discussed later in Sec. 134.) Finally, the basis for the Rydberg-Schruster law is simple and obvious: the difference between the common convergence limit of the sharp and the diffuse series and the limit of the principal series is just the difference between the lowest two energy levels; and this is **also** equal to the wave number of the first line of the principal series.

The convergence limit of the principal series, corresponding to a jump of the electron from infinity into the normal state, obviously corresponds to an energy that is (1) numerically equal to that of the normal state, when this is measured downward from zero at infinity, and (2) equal to the minimum energy that must be given to the atom to remove an electron from it, *i.e.*, to the *ionization energy* of the neutral atom. The ionization energy is usually expressed in electron-volts and called the *ionization potential*, representing the potential through which an electron must drop in order to acquire enough energy to ionize the atom by impact. It is 5.13 volts for sodium, 13.58 volts for a hydrogen atom.

The discussion of spectra in terms of energy levels is thus much simpler than the direct discussion in terms of the lines themselves. Whatever may happen in the future to our picture of the behavior of the electrons in an atom, energy levels are undoubtedly here to stay.

99. Excitation and Ionization of Atoms by Electrons.—In the last few sections, we have ascribed the emission of radiation to transitions made by atoms from one stationary state to another of lower energy, without considering how the atoms get into the higher states initially. Since the energy of the atom decreases with each emission of a quantum of radiation, the final result of emission processes can only be that the atom arrives in its normal state, or state of lowest energy, and then remains there without radiating farther. Thus, in order to be able to emit radiation, the atom must first be transferred by some means from its normal state into a state of higher energy.

When an atom is in such a state of higher energy, it is said to be *excited*, and the process of transferring or "raising*" it from its normal state into an excited state is referred to as *excitation*. The excited atom may be thrown into any one of its infinite series of discrete quantum states. It may also happen, however, that the process of excitation is so intense that the electron is completely removed from the nucleus. The atom is then left with a net positive charge and, if free, will tend to move in an electric field, like an ion in an electrolytic solution. Such an atom which has lost 1 electron (or more) is said to be in an *ionized* state, and the process of raising the atom from the normal state to the ionized state is called *ionization*. In the process either of excitation or of ionization, additional energy is given to the atom.

A full discussion of the subject of the excitation and ionization of atoms, and of the far-reaching conclusions drawn from experiments in this field, is beyond the scope of this book. We may, however, make brief mention of some of the ways by which experiment shows that

atoms may become excited or ionized **and** of the interpretation of some of these experiments.

The atoms of a gas or vapor may become excited or ionized by bombarding them with electrons. An atom becomes ionized when the energy given to it in a collision is sufficient to expel an electron from its normal orbit in the atom to infinity.

As one illustration of the mode of procedure followed in studying phenomena of this kind may be mentioned an experiment by Foote, Meggers, and Mohler.¹ Their apparatus is shown diagrammatically in Fig. 63. A filament *F* of tungsten or lime-coated platinum is heated by a battery *B*₁ to such temperature that it emits electrons.

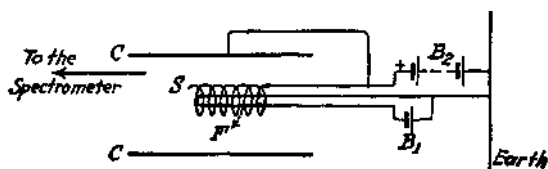


FIG. 63.—The apparatus of Foote, Meggers, and Mohler for studying excitation potentials.

Around the filament is a spiral grid *S*, which, by means of a battery *B*₂ or potentiometric source of *pd*, is maintained at any desired *positive* potential with respect to the filament. Around *S* and electrically connected thereto is a metal cylinder *CC*; *S* and *CC* are thus at the same potential. Inside *CC* is the gas or vapor under study, maintained at a suitable pressure. Electrons, emitted by the filament, are accelerated toward the grid *S*, which is so close to *F* that with proper regulation of the gas pressure comparatively few atoms of the gas are struck by the electrons in their passage from filament to grid. After leaving the grid, the electrons move in the force-free space between *S* and *CC*, in which space, because of the greater distance *S* to *CC*, the electrons collide with the gas molecules causing the excitation of the latter, provided the energy of the electrons is sufficiently great.

With sodium vapor inside the cylinder, no luminosity is observed in the vapor until the potential difference between the grid and the filament reaches 2.09 volts. For voltages slightly above this value, the spectrograph shows that the sodium vapor emits the well-known *D* lines, *and those only*. The mean wave length of these lines is 5,893 angstroms, corresponding to a frequency ν of $0.509 \times 10^{16} \text{ sec.}^{-1}$; the energy quantum $h\nu$ is

$$6.61 \times 10^{-27} \times 0.509 \times 10^{15} = 3.36 \times 10^{-12} \text{ erg.}$$

¹ FOOTE, MEGGERS, and MOHLER, *Astrophys. J.*, vol. 55, p. 145 (1922); cf. also FOOTE and MOHLER, "The Origin of Spectra."

According to the quantum picture of the origin of spectra, there must have been a *change of state* of the sodium atom in *emitting* this line corresponding to an *energy drop* of 3.36×10^{-12} erg—whatever may have been the absolute values of the energies in the initial and the final state. Now, the electrons, which, after emission from the filament, have dropped through 2.09 volts, have acquired a kinetic energy given by

$$eV = 4.80 \times 10^{-10} \times \frac{2.09}{3 \times 10^2} = 3.34 \times 10^{-12} \text{ erg.}$$

We see, therefore, that the kinetic energy possessed by the electrons as they pass through the space between the grid and the cylinder is equal to the quantum energy of the radiation emitted by the sodium vapor. The presumption is very strong, therefore, that the electrons have, by collision, transferred to the sodium atoms sufficient energy to raise the latter from their *normal* state to an excited state

$$3.36 \times 10^{-12} \text{ erg}$$

above the normal, and that the atoms in returning from that excited state to the normal state emit the doublet *D* line.

In discussing such observations, it is convenient to express all spectral frequencies in equivalent electron energies. If the quantum energy $h\nu$ of a line is equal to the energy acquired by an electron in falling through V volts or $V/300$ electrostatic units of potential, we have

$$\frac{Ve}{300} = h\nu = hc\tilde{\nu} = \frac{hc}{\lambda}.$$

Inserting $e = 4.803 \times 10^{-10}$, $h = 6.61 \times 10^{-27}$, $c = 2.998 \times 10^{10}$; and multiplying λ by 10^8 to convert to angstroms, we obtain the convenient conversion formulas¹

$$\tilde{\nu} = 8,079V, \quad \lambda = \frac{12,378}{V} \text{ A.} \quad (147a,b)$$

The quantity V is a measure of the energy of the radiant quantum, or of the energy lost by the atom in emitting it, in terms of the *electron-volt* as a unit.

To return to the experiment of Foote, Meggers, and Mohler, it was found that, as the voltage between filament and grid was

¹ If we employ the "spectroscopic" value of h , as described in Sec. 95, or $h = 6.624 \times 10^{-27}$, and also $e = 4.8025 \times 10^{-10}$, $c = 2.99776$, formulas (147a,b) become

$$\tilde{\nu} = 8,062V, \quad \lambda = \frac{12,404}{V} \text{ A.}$$

raised above 2.09 volts, the *D* lines continued to appear, but no others appeared until the voltage reached 5.12 volts, beyond which the complete arc spectrum was produced. Beyond 35 volts, the so-called "spark" spectrum was seen. The explanation of these facts cannot be quite the same as for the *D* lines. For electrons possessing energy equivalent to, say 3.0 volts, should, so far as *energy* is concerned, be able to communicate to sodium atoms sufficient energy to generate lines of wave length 4,130 angstroms or longer. In the spectrum of sodium is found a large number of lines of longer wave length than this, for example practically the entire sharp and diffuse series, the convergence wave length of which is of the order of 4,100 Å. Why do not 3-volt electrons excite these lines? Or, for that matter, the wave length of the second line of the sharp series is 6,158 Å or 2.01 electron-volts. Why is this line not produced at even lower voltages than the *D* lines, which correspond to 2.09 volts?

A little study of the energy-level diagram of sodium in Fig. 62 (b) above reveals the answers to these questions. A given line can be emitted only if atoms are present in the state which is the *initial* state for that line. To emit the line at 6,158 Å, the atom must start from a state with an energy much above the initial state for the *D* lines, this state being actually the *final* state for the former line; and 2.09-volt electrons cannot possibly put the atom into so high a state. Even 3-volt electrons cannot excite any state above 3^2P , from which the *D* lines start. On the other hand, since the convergence wave number of the principal series of sodium is 41,450, equivalent to 5.13 volts, an electron with more than 5.13 volts of energy can remove an electron from the atom to infinity, *i.e.*, it can *ionize* the atom. Subsequently, an electron will be picked up by the ionized atom, dropping from one discrete level into another with the emission of radiation until it reaches the lowest level and the atom is again in its normal state. Consequently, upon viewing the radiation from many different atoms, the entire principal, sharp, and diffuse series and all other lines of the arc spectrum of sodium may be seen.

The *spark* spectrum, appearing when the energy of excitation exceeds 35 volts, is ascribed to ionized atoms in which a *second* atomic electron has been raised to a higher level as the result of a second impact by an electron from the filament. This second atomic electron then causes the emission of radiation as it drops back into lower levels in the ionized atom.

The energy in volts necessary to raise an atom from its normal state into a given excited state is known as the *excitation potential* for that state. The first observation of an excitation potential was

made by Franck and G. Hertz in 1914.¹ They showed that electrons with an energy of 4.9 volts or more, but not less than 4.9 volts, would excite the resonance line at 2,536 Å in mercury vapor. They made this observation in the course of a long series of experiments on the energy gained by an electron when passing through a gas under the influence of an electric field. Observations of excitation potentials were important at that time because they furnished direct experimental proof of the existence of quantized energy levels.

100. Absorption and Reemission of Radiation, (a) *Absorption.*—On the classical theory, the characteristic frequencies emitted by an atom should be identical with the natural frequencies of the atom. Conversely, if light of one of these frequencies falls upon the atom, it should set the atom into resonance, some of the energy being thereby abstracted or "absorbed" from the incident beam. It follows that the *absorption spectrum* of a gas should be exactly the same as its emission spectrum.

For certain spectral lines this conclusion is substantiated by experiment in the well-known phenomenon of the *reversal* of spectral lines, of which phenomena the most conspicuous are the dark lines in the solar spectrum. The reversal of the *D* lines of sodium is a familiar laboratory or lecture demonstration. But many of the lines emitted by a gas in a discharge tube are not ordinarily observable in its absorption spectrum. All absorption lines are found to occur under suitable circumstances as emission lines, but not the reverse. For example, neon gas is highly transparent to visible light; it does not absorb the red wave lengths whose emission is responsible for the brilliance of neon signs.

The explanation of these facts on the Bohr theory is easily seen. The process of absorption will be just the reverse of the process of emission; a quantum of radiant energy is absorbed while simultaneously the atom jumps from one quantum state into a state of *higher* energy. The *possible* frequencies for the lines should, therefore, be the same in absorption as in emission. But in order to absorb a given line, the atom must already be in the state which is the lower state for that line. Since the atoms of a monatomic gas under ordinary conditions are in their normal states of lowest energy, they can absorb only those lines which start, in absorption, from the normal state. The restriction of the absorption spectrum to certain lines is thus accounted for. For neon, all lines that involve the normal state lie in the ultraviolet. Sodium vapor shows in absorption only the *principal series* [cf. diagram in Fig. 62(6) in Sec. 98]. From

¹ FRANCK and HERTZ, *G. phys. Ges., Verhandlungen*, vol. 16, p. 512 (1914).

the head of the principal series, however, a *continuous* absorption band extends toward higher frequencies. This is obviously caused by processes in which the energy of the incident photon $h\nu$ exceeds the ionization potential, so that the atom becomes ionized with the ejection of an electron at considerable speed. That is, we have here photoelectric emission from the sodium atoms, the electron being ejected with kinetic energy

$$\frac{1}{2}mv^2 = h\nu - eV_0,$$

where V_0 is the ionization potential of the atom and e the numerical electronic charge. The equation is a special case of Einstein's photoelectric equation.

When atoms in *excited states* are present, as in a discharge tube, additional lines may appear in the absorption spectrum.

The energy given to an atom in an absorption process may be lost subsequently in any one of several ways. Collisions with other molecules may cause the excited atom to return to its original quantum state, the energy of excitation being added to the kinetic energy of motion of the colliding particles. In this case the energy absorbed from the incident radiant energy is converted into heat. Collisions of this sort have been called "collisions of the second kind," the converse collisions, in which kinetic energy is converted into energy of excitation, being called collisions "of the first kind."

(b) *Resonance Radiation and Fluorescence*.—A second way in which an atom may lose an absorbed quantum of energy is by jumping into some other quantum state with the emission of a quantum of radiation. One possibility, which we shall discuss first, is that the atom may drop back into its normal state, from which it was removed by the process of absorption; in this latter case the radiation emitted will have the same frequency as the radiation originally absorbed. The net effect of such absorptions and reemissions is a powerful scattering of the incident light in all directions. To this phenomenon R. W. Wood gave the name *resonance radiation*.

Wood showed¹ many years ago that a bulb containing sodium vapor at very low pressure would, when irradiated by light from an intense sodium flame, emit the *D* lines, and those only. The resonance was very sharp. An examination of the reemitted *D* lines showed that they were very narrow, their width corresponding almost exactly to the width predicted by the Doppler effect at the temperature of the sodium vapor. The width of the lines of the exciting source was much greater, since the temperature of the source was much higher

¹ See WOOD, R. W., "Physical Optics."

than that of the vapor. On analyzing the light which had passed through the vapor, it was found that there was a narrow absorption line at the centers of the broad D lines.

This experiment of Wood's was in beautiful agreement with the classical theory. But (the present) Lord Rayleigh showed¹ that if sodium vapor is illuminated by the *second* line of the principal series, $X = 3,303 \text{ \AA}$, both that line and the D lines were emitted by the vapor. The emission of the D lines under such circumstances is a case of *fluorescence*, and, like fluorescence in general, is difficult to explain on the classical theory, whereas the explanation by the quantum theory is easy. The absorption of $X = 3,303$ leaves the atom in the excited state known as 4^2P [see Fig. 62(6) in Sec. 98]. From this state it cannot pass at once, for a reason described later (Sec. 131) into the 3^2P state, which is the initial state for the D lines, but it can jump into a 4^2S state, in a transition not shown on the diagram, with the emission of an infrared quantum, and then from this state it can jump into the 3^2P state, with the emission of a quantum of $X = 11,404$ or $11,382$, which is the first (doublet) line of the sharp series, and then finally into the normal state. Thus, the single quantum absorbed from the incident light is eventually reemitted as three fluorescent quanta. A second but less likely possibility is a jump from 4^2P to 3^2D , and then into 3^2P with emission of the first line of the diffuse series.

Excitation by the absorption of quanta is quite analogous to excitation by electron impact, except for this one important difference. The atom may become excited when colliding with an electron the energy of which *equals or exceeds* that required for the increase in energy level; whereas excitation by absorption takes place only when the energy of the incident quantum is *exactly equal* to that required to produce the particular change of state.

(c) *Excitation by Collision with Other Atoms.*—The atoms of a monatomic gas are continually colliding with each other due to thermal agitation. There is ample evidence to confirm the view that when an atom A in an excited state collides with an atom B in its normal state, both atoms may undergo a quantum jump. Thus B becomes excited, while A returns to its normal state, or drops into some intermediate state; and thereafter B may radiate one of its own characteristic frequencies. At ordinary temperatures the kinetic energy of the atom is relatively negligible; hence, the energy gained by B in such a process cannot appreciably exceed that which is available in the initial excitation of A !

¹ STRUTT, R. J., Bakerian Lecture, Roy. Soc, Proc, vol, 98, p, 272 (1916),

To illustrate phenomena involving the transfer of the energy of an excited state from one atom to another, mention may be made of the experiments on mercury vapor¹ excited to resonance radiation by the absorption of its own line $\lambda = 2,536$ A ($h\nu = 4.9$ volts). If with the mercury vapor is mixed the vapor of thallium, the characteristic lines of the latter element appear *in addition to the mercury resonance radiation* when the mixture is illuminated with the *mercury line* $X = 2,536$. The *mercury resonance radiation* is then *weaker than when no thallium is present*. Thallium vapor *alone* is not excited to resonance by the mercury line. This is explained by assuming that mercury atoms are first raised to an excited state by absorption of the line $X = 2,536$; some of these excited atoms by collision with thallium atoms then transfer their energy of excitation to the latter, which

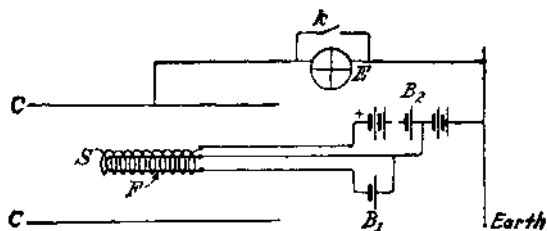


FIG. 64.

subsequently radiate their characteristic lines in returning to the normal state. The thallium lines so produced correspond to energy transfers in the thallium atom of *less than 4.9 volts*.

Similarly, Cario showed that the presence of argon in mercury vapor materially reduces the intensity of the resonance radiation when the mixture is illuminated by the $\lambda = 2,536$ line, but without exciting any argon radiation. Yet an examination of the light transmitted through the mixture shows that there is no diminution in *absorption* as a result of the presence of the argon. This effect is explained by assuming that collisions of the second kind take place between the excited mercury atoms and the argon atoms and that the *energy of excitation* of the former is transformed into *kinetic energy* of both atoms.

(d) *Electrical Methods of Observing the Excitation of Atoms.*—Referring to Fig. 63, it is readily seen that if the potential difference between grid and filament is sufficiently high, the electrons may be given sufficient energy to ionize by collision the molecules of the vapor in the space between the cylinder and the grid. In Fig. 64, the filament F is so connected to the battery B_2 that the potential

LORIA, *Phys. Rev.*, vol. 26, p. 573 (1925); CARIO, *Zeits. f. Physik*, vol. 10, p. 185 (1922); and CARIO and FRANCK, *Zeits. f. Physik*, vol. 17, p. 202 (1923).

difference between F and the grid S is less than, and in the opposite direction to, the potential difference between CC and S . Electrons accelerated from F toward S cannot, therefore, reach CC . The cylinder CC is connected to earth through the sensitive electrometer E and the grounding key k . As long as no positive ions are produced in the space between S and CC , the cylinder should acquire no charge. But when the velocity of the electrons accelerated from F to S becomes sufficient to ionize the vapor, the positive ions thus produced should be attracted to CC and an *ionization* current should be set up which is measurable by the rate at which the electrometer acquires a charge when the grounding key is open.

An effect of this kind was observed by Lenard many years ago. He found that with such gases as air, hydrogen, and carbon dioxide a

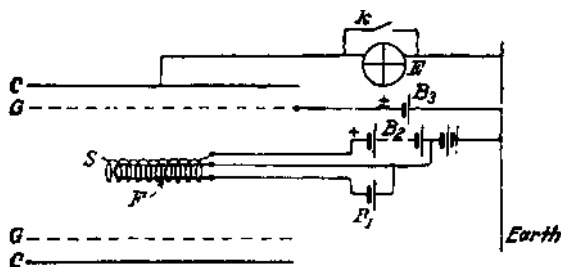


FIG. 65.—The apparatus of Davis and Goucher for studying ionization potentials.

current, which he thought to be a true ionization current, began to flow when the exciting electrons had fallen through a potential difference of some 11 volts. We have seen, however, from the experiments of Meggers, Foote, and Mohler, that, in the case of sodium vapor, *resonance radiation* is produced by 2.09-volt electrons and, in the case of mercury vapor, by 4.9-volt electrons. This resonance radiation, proceeding in all directions, falls on the inside walls of the cylinder, and, if the radiation is of sufficiently short wave length, *photoelectrons* will be expelled from the cylinder. Because of the direction of the field between CC and S , these photoelectrons will flow toward the grid, and we shall have in the circuit $CB2S$ a *photoelectric* current in exactly the same direction as if there were an ionization current. This effect will begin abruptly as soon as the resonance potential is reached. The current observed by Lenard may have been of this nature.

To separate the true ionization current from the photoelectric current, Davis and Goucher¹ introduced a coarse wire gauze G surrounding the grid and just inside the plate CC . By means of a battery

¹ *Phy. Rev.*, vol. 10, p. 101 (1917).

$B3$,¹ (Fig. 65), this gauze may be maintained at either a positive or a negative potential with respect to CC . When G is sufficiently negative with respect to CC , photoelectrons ejected from CC by the action of resonance radiation will not reach G and so cannot escape from CC . Photoelectrons ejected from G , however, may reach CC , and so a photoelectric current may result. If now $B3$ is reversed, so that G is positive with respect to CC , a photoelectric current in the opposite direction should result, due to photoelectrons ejected from CC . In either case the potential difference between CC and G is kept small as compared with that between G and S . The true ionization current cannot be reversed by reversing the direction of Ba , since the positive ions produced between S and G acquire sufficient velocity so that, passing through the gauze, they will reach CC in spite of any small opposing field between CC and G .

With this type of apparatus, Davis and Goucher obtained, for mercury vapor, curves of the type shown in Fig. 66, in which the abscissas are the accelerating potentials between filament and grid and the ordinates are the currents to the cylinder CC as measured by the rate of charge of the electrometer. Curve A was obtained when the gauze G was *positive* with respect to the cylinder; curve B when the gauze was *negative*. No current was observed until the accelerating voltage reached 4.9 volts. Beyond that voltage a current was observed which could be reversed by reversing the battery $B3$, indicating that the current was a photoelectric current. At point c , corresponding to 10.3 volts, a sudden change in the variation of the currents with increasing potential took place, indicating that the cylinder was beginning to collect positive ions. Up to point c , curve A is qualitatively the mirror image of curve B .

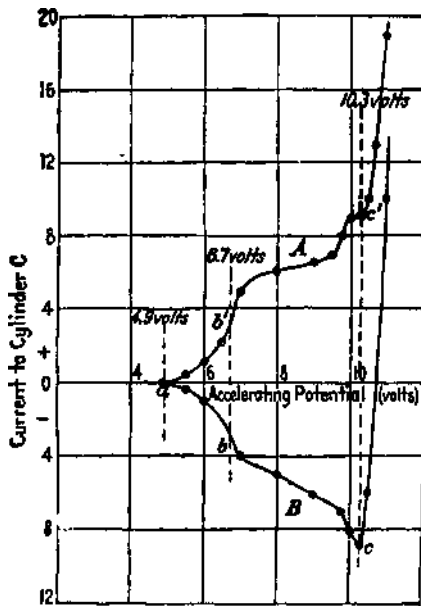


FIG. 66.—The resonance and ionization potentials of mercury vapor, as observed by Davis and Goucher.

With this type of apparatus, Davis and Goucher obtained, for mercury vapor, curves of the type shown in Fig. 66, in which the abscissas are the accelerating potentials between filament and grid and the ordinates are the currents to the cylinder CC as measured by the rate of charge of the electrometer. Curve A was obtained when the gauze G was *positive* with respect to the cylinder; curve B when the gauze was *negative*. No current was observed until the accelerating voltage reached 4.9 volts. Beyond that voltage a current was observed which could be reversed by reversing the battery $B3$, indicating that the current was a photoelectric current. At point c , corresponding to 10.3 volts, a sudden change in the variation of the currents with increasing potential took place, indicating that the cylinder was beginning to collect positive ions. Up to point c , curve A is qualitatively the mirror image of curve B .

¹ These "batteries" are really variable potentiometric sources of potential difference.

The beginning of the photoelectric currents at 4.9 volts is due to the excitation by electron impacts of the mercury line $X = 2,536$, which then acts photoelectrically on the cylinder and the gauze. The rapid increase in photoelectric current which occurs when the voltage reaches 6.7 volts is attributed to the generation, at that voltage, of the mercury line $X = 1,849$; whereas the sudden break in both curves at 10.3 volts is indicative of the beginning of the real ionization current. The quantities 4.9 and 6.7 volts are spoken of as *resonance potentials* of mercury vapor, and 10.3 volts as its ionization potential.

(e) *Ionization Potentials.*—As was remarked in Sec. 96, the ionization potential of an atom can also be calculated from the convergence limit of its "principal series," or the series for which the normal state is the final state in the emission of a line. The ionization potential is the equivalent of this convergence limit in volts. Ionization potentials calculated in this way from spectroscopic data agree well with those obtained by electrical methods. Higher ionization potentials can also be denned, the *second* ionization potential representing the energy necessary to remove a *second* electron from the atom, and so on. These higher ionization potentials can be calculated from the convergence limits of other series. Because of the greater precision obtainable, the spectroscopic method of determining ionizing potentials is usually employed.

For example, in the arc spectrum of helium there is a singlet series of lines due to transitions ending on the normal state, with a convergence frequency¹ $\nu = 198,298 \text{ cm.}^{-1}$. According to Eq. (147a) in Sec. 99, this corresponds to a first ionization potential of 24.54 volts. That is to say, it takes energy equal to 24.54 electron-volts to remove 1 of the 2 electrons from a helium atom. When this has been done, the ion behaves as a 1-electron atom. Equation (145) in Sec. 97 gives (with $n_1 = \infty$, $n_2 = 1$), as its ionization energy, $4R$, where R is the Rydberg constant for helium or $R = 109,722.26$. Thus $\nu_{\infty} = 4R$ or $\nu_{\infty} = 438,888$. The *second* ionization potential of helium is thus $438,888/8,079 = 54.32$ volts.

Values of many of the ionization potentials for the lighter elements are given in Table IV. They have been calculated from data in Bacher and Goudsmit's "Atomic Energy States," the numbers there given for the "absolute value of the lowest state" being divided by **8,079** to obtain values of the ionization potential in volts. First ionization potentials (for the neutral atom) are given under "I", second under "I I", and so on.

¹ BACHEK and' GOUDSMIT, "Atomic Energy States—as Derived from the Analyses of Optical Spectra," p. 220.

TABLE IV.—SOME IONIZATION POTENTIALS IN VOLTS

Element		I	II	III	IV	V	VI
H	1	13.58					
He	2	24.54	54.33				
Li	3	5.38	75.51	122.2			
Be	4	9 31	18.18	153.6	217 3		
B	5	8 30	25 1	37.87	258 9	339 5	
C	6	11.25	24 34	47.80	64 38	391.1	
N	7	14 52	29 57	47.54	77 3	97.7	
O	8	13.60	35.03	55 04	77.2	114.	137.9
F	9	18.7	34.7				
Ne	10	21.53	40.9	63.3			
Na	11	5.13	47.2				
Mg	12	7.63	15.01	80.			
Al	13	5.98	18.79	28 40			
Si	14	8.14	16.32	33 44	45.07		
P	15		19.9	30 1		64.92	
S	16	10.3	23.4	35 0	47.22	62.1	87.91
Cl	17	13.00	18.36	39.8			
Ar	18	15.73	27.82				
K	19	4 33	31.8	47.0			

Many beautiful experiments have been devised to extend the data and theories regarding the excitation of atoms by impact or resonance. The reader will find further study in this field, from original sources, very fascinating. The literature is extensive.

101. The Boltzmann Distribution Law.—In the last section, one other easy method of exciting spectral emission has not been mentioned. Many elements in gaseous form, especially the vapors of metals, emit line spectra upon being heated; thus sodium vapor heated by a Bunsen flame emits the familiar *D* lines, whether the vapor is mingled with the flame itself or is enclosed in a tube. In such cases, it is evident that the atoms or molecules are excited by *thermal agitation*.

If the substance is in thermal equilibrium, a simple theoretical formula can be given for the relative numbers of the atoms or molecules that are in each quantum state. It is deduced from statistical mechanics that, in the case of thermal equilibrium, the average number N_i in a quantum state in which the atomic or molecular energy is ϵ_i is

$$N_i = C e^{-\frac{\epsilon_i}{kT}} \quad (148a)$$

where k is **Boltzmann's constant** (Sec. 79), T is the absolute tempera-

ture, and C is a constant of proportionality. If N is the total number of atoms or molecules, we must have

$$N = \sum_i N_i = C \sum_i e^{-\frac{\epsilon_i}{kT}},$$

the sum extending over all possible quantum states. By eliminating C between these two equations we can also write

$$N_i = \frac{N e^{-\frac{\epsilon_i}{kT}}}{\sum_j e^{-\frac{\epsilon_j}{kT}}}.$$

It often happens, however, that a number of quantum states have the same energy, and then it may be more convenient to group these states into a multiple state. Suppose we have formed in this manner all the multiple states that we can, no two multiple states having the same energy. Let these multiple states be numbered off in a single series, and let the energy of multiple state number t be ϵ_t and the number of fundamental states composing it w_t ; the number w_t is often called the *statistical weight* of the multiple state. Then by Eq. (148a), when thermal equilibrium exists, the number N_i is the same for each of the fundamental states composing a multiple state, so that the number of atoms or molecules in multiple state number t is just $w_t N_i$. Calling this number N_t , we can write for it, replacing ϵ_i by ϵ_t in (148a),

$$N_t = C w_t e^{-\frac{\epsilon_t}{kT}}. \quad (148c)$$

These formulas constitute special cases of what is known as the *Boltzmann distribution formula* for quantized systems. We have already encountered an example of the formula in Eq. (115) in Sec. 83 for the distribution of a set of harmonic oscillators in thermal equilibrium. Maxwell's law constitutes an example of the corresponding formula in classical theory.

As an example, the sodium D lines result, as we have seen, from transitions between either of two excited levels lying close together and the normal level (c/. Sec. 98). For our purpose we may treat the upper two levels as a single composite level. Then it follows from the more modern theory of wave mechanics, to be described in the next two chapters, that this composite upper level is composed of six fundamental states, so that for it $w = 6$, whereas, for the normal level, $w = 2$. The two levels lie $h\nu = 3.36 \times 10^{-12}$ erg apart. Hence, indicating them by subscripts 1 and 0, respectively, we have for the

numbers of atoms in them,

$$N_1 = 6Ce^{-\frac{41}{kT}}, \quad N_0 = 2Ce^{-\frac{91}{kT}}.$$

In the case of sodium in a Bunsen flame at 1800°C , this gives

$$\frac{N_1}{N_0} = 3e^{-(41-91)/kT} = 3e^{-3.36 \times 10^{-12} / (2073 \times 1.38 \times 10^{-16})} = 2.3 \times 10^{-6}.$$

Thus only a very small part of the sodium atoms are excited in this manner at any given time. They suffice, however, to cause a considerable emission of sodium light. With further increase of temperature the higher members of the principal series and members of other series are found to appear.

102. The Extension of Bohr's Theory.—In his original paper, Bohr remarked that the orbit of the electron in a hydrogen atom might be an ellipse instead of a circle. A detailed theory of elliptical orbits was developed by Sommerfeld several years later. The geometrical ideas involved still possess a certain interest.¹

According to the laws governing planetary motion, or any motion under an inverse-square force, an elliptical orbit will have one of its foci at the nucleus and the energy of the system will depend only on the length of the major axis of the ellipse. The orbit will lie in a fixed plane, so that the motion can be described by means of two coordinates, for which we may take polar coordinates r, θ , with the origin at the nucleus. Then, as θ increases through 2π , r increases from its minimum value at one end of the ellipse to a maximum at the other end, after which it decreases again to a minimum. Thus r executes what is called a "libration" during each revolution of the electron. Sommerfeld assumed that the same generalized quantum condition which had been postulated for the harmonic oscillator and the circular orbit, as represented by Eq. (131) in Sec. 93, will hold for r ; i.e., he assumed that

$$\oint p_r dr = n'h,$$

where p_r is the momentum in the radial direction, the integral is taken throughout a libration, and n' is a positive integer or 0. For 0, he assumes with Bohr that the angular momentum is an integer k times $h/2\pi$. (In his books Sommerfeld writes *no* or just n for k and sometimes n_r for n') Sommerfeld showed then that the energy depends only on the "total" quantum number,

$$n = k + n',$$

¹ For a more complete account see A. Sommerfeld, "Atomic Structure and Spectral Lines," Methuen & Co., Ltd., London, 1929; also Ruark and Urey, "Atoms, Molecules and Quanta," p. 132.

and is, in fact, *the same function of n as in BoWs theory* [Eq. (136) in Sec. 94]. Thus up to this point nothing is gained except a greater variety in the possible types of orbit. For $n = 1$, we must have $n' = 0$, since k cannot be less than 1; *i.e.*, only the Bohr circle is possible. But, for $n = 2$, we can have either the Bohr circle, with $k = 2$ and $n' = 0$, or an ellipse with major axis equal to the diameter of the circle, corresponding to $k = 1$, $n' = 1$, and so on.

Sommerfeld then showed, however, that, if allowance is made for the known *variation of electronic mass with speed*, the energy of the elliptical motion is slightly different from that of the circular motion. The ellipse is also caused to precess slowly about the nucleus. In this way, he arrived at a splitting or fine structure of the levels of hydrogen, or of any other 1-electron atom. This splitting appeared to be in quantitative agreement with observation. Sommerfeld arrived, in fact, at the same expression for the energy as was obtained later from the wave-mechanical theory as modified to bring it into harmony with relativity. Sommerfeld's theory, however, included no consideration of the effects of electron spin, which was discovered much later; hence, we shall postpone all discussion of the fine structure of the hydrogen lines until we can take it up on a more adequate basis (Sec. 136).

In the meantime, Bohr and others were endeavoring to extend the theory to atoms containing more than 2 electrons. Little quantitative success was achieved in this direction. In particular, no plausible arrangement of orbits could be devised for the 2 atoms of helium which would give the correct value for the second ionization potential (54.3 volts). In a broader way, Bohr endeavored to understand the known variation of chemical and physical properties from atom to atom through the periodic table. His results were so unsatisfactory, however, that the old static atom of J. J. Thomson was revived for a time in a modified form, but without great success (see Sec. 123).

The truth is that two principles essential for the understanding of complex atoms were at that time unknown—the Pauli exclusion principle and the phenomenon of electron spin. Accordingly, we shall follow the old quantum theory no further and shall devote the next chapter to a description of the modern theory of wave mechanics. This theory furnished also a definite answer to another problem of spectroscopy that is second in importance only to that of spectral wave lengths, *viz.*, the relative intensities of spectral lines as they are observed under given conditions.

CHAPTER VII

WAVE MECHANICS

One must be prepared to approach the subject of this chapter philosophically, prepared to accept conclusions which are, at first thought, seemingly at variance with our senses and with a belief that has persisted almost unquestioned from the time of the Greeks, *viz.*, that matter is made up of particles. We have seen in Chap. I that light possesses both undulatory and corpuscular characteristics. But even so, we might say, light differs from matter; whereas we can determine the nature of light only by indirect observation, matter we can see. We observe directly that a handful of sand is made up of real particles. In ordinary experience they certainly do *not* exhibit wavelike characteristics. The particles of sand, which we see so clearly, are, however, made up of molecules and atoms and electrons and protons and neutrons, *none of which we can see directly any more than we can see light waves or photons*. It is with these *so-called* particles, evidence concerning which is almost as indirect as with photons, that the wave theory of matter is primarily concerned.

We shall attempt in this chapter to give merely a brief introduction to this most important subject. Our purpose will be to show how the concept of matter waves can be developed more or less naturally from the previous concepts of both classical and quantum physics and to summarize the pertinent experimental evidence. Certain features and results of the new mathematical theory known as *wave mechanics* will also be described, but no complete presentation of mathematical details will be attempted, as this would require mathematical tools and technique much beyond the scope of this book.

103. Matter Waves, (a) *Matter and Energy*.—With the discovery of the law of the conservation of energy toward the middle of the nineteenth century, it became obvious to physicists that the physical universe is made up of two great entities, *viz.*, matter and energy, each of which is "conserved." These two great conservation laws provided much of the foundation upon which classical physics was built. By 1900, the corpuscular nature of matter—molecules, atoms, electrons—had become firmly established; likewise the undulatory nature of light. By 1910, furthermore, Planck's quantum theory and the Einstein photoelectric equation together with various lines of

experimental evidence had made it clear that, impossible as it then seemed, physicists would have to accept the hypothesis that light itself possesses corpuscular characteristics. By 1920 the dual nature of radiant energy was generally recognized, although not understood, and, with the gradual accumulation of experience, physicists had almost become accustomed (perhaps "resigned" would be more nearly correct!) to using either the corpuscular or the undulatory theory of light, according to the experimental phenomena in hand.

During all this time, there was no suggestion that *matter* was anything but corpuscular. Matter was made up of atoms, which in turn were made up of electrons and nuclei. But, in 1924, Louis de Broglie¹ made the very bold suggestion that particles of matter, and in particular electrons, might possess certain undulatory characteristics, so that they, too, might exhibit a dual nature. He suggested also a way in which the undulatory characteristics of electrons might perhaps furnish a new basis for the quantum theory. He did not develop his ideas into an exact theory, however. His way of thinking was rather akin to that of the ancient Greeks. The reasoning used might almost be paraphrased as follows: "(1) Nature loves symmetry. (2) Therefore the two great entities, matter and energy, must be mutually symmetrical. (3) If (radiant) energy is undulatory and/or corpuscular (?), therefore matter must be corpuscular and/or undulatory (?)."

Had nothing further happened, de Broglie's speculations would doubtless soon have been forgotten. Reasoning of this sort, in the mind of a genius, may point the way to an advance, but, before definite scientific progress can be achieved, the new ideas must be precisely formulated and compared with experiment. The importance of doing this was overlooked by the ancient Greeks. De Broglie's loosely woven speculation set a German, E. Schrodinger, to thinking; and he discovered how to develop certain of de Broglie's ideas into a precise mathematical theory.² With additions made to it by Born, Heisenberg, Dirac, and others, this theory has become the highly successful quantum mechanics of the present day. In what follows, we shall endeavor to arrive at the basic assumptions of Schrodinger's theory by developing de Broglie's ideas in a natural and logical manner.

(b) *Matter Waves*.—In his first paper, de Broglie is concerned primarily with developing a theory of light in terms of "light quanta,"¹ DE BROGLIE, *Phil. M.* p. 22 (1925).

² SCHRODINGER, *Ann. d. Physik*, vol. 79, pp. 361, 489, 734; vol. 80, p. 437; vol. 81, p. 109 (1926).

or photons. If the energy of the light is concentrated in photons, how are the phenomena of interference to be understood? There must be waves of some sort associated with the photons, in order to account for the observed interference effects. We can no longer suppose that the energy is spread out over these waves, as in classical theory; nevertheless, the waves must somehow determine where, in an interference pattern, the photons can produce effects by being absorbed. The details of interference patterns depend largely upon the phase relations of the waves; hence, de Broglie called the latter *phase waves*. He assumed their frequency ν to be such that the energy in a photon equals $h\nu$.

But why should such waves be associated only with photons? A material particle carries energy; and "it would seem that the basic idea of the quantum theory is the impossibility of imagining an isolated quantity of energy without associating with it a certain frequency."¹ Material particles ought, therefore, like photons, to be accompanied by phase waves of some sort; and these waves ought, under suitable circumstances, to give rise to interference effects. Furthermore, to complete the analogy, the waves associated with a particle moving at speed v should have a frequency equal to the energy of the particle divided by h .

Regarding the energy of a material particle, however, there is a divergence between the assumptions of de Broglie and of Schrodinger. De Broglie includes in the energy the rest energy, m_0c^2 , the existence of which is suggested by the theory of relativity (Sec. 65). Schrodinger, on the other hand, prefers to construct first a nonrelativistic theory. He assumes the mass of the particle to be constant and equates just the ordinary energy, kinetic plus potential, to $h\nu$. In this respect we shall follow Schrodinger, reserving comment on the relativistic form of the theory, which must be used for high-speed particles, to a later section (Sec. 117). We shall assume, therefore, that any material particle of mass m moving at speed v has associated with it in some manner waves of frequency ν given by the relation

$$h\nu = \frac{1}{2}mv^2 + V, \quad (149)$$

where h is Planck's constant, m is the mass of the particle, and V is its potential energy, due to electrical fields or to other causes.

The physical nature of these waves was left indefinite by de Broglie. We cannot go to the extreme of supposing that a material particle is just a group of waves; for then its mass and energy, and also its charge if it has a charge, would of necessity be spread out over these waves and

¹ Translated from *Ann. de Physique*, vol. 3, p. 32 (1925).

would, in consequence, soon become scattered widely in space, contrary to the observed facts. It is of the nature of waves to diverge toward all sides. We should not be disturbed, however, by the impossibility of visualizing the waves. We should remember our experience in optics. Using classical theory, it was easy to picture light as wave motion; but, if we retain this concept, it is very difficult to picture a beam of light as a moving stream of photons. Similarly, as long as we retain the particle concept of an electron or proton, we cannot hope to form a concrete picture of the accompanying waves. Perhaps, even, they are only mathematical waves, so to speak, a device that we employ for the purpose of making calculations and predicting the results of observation. In adopting this standpoint, we are doing only what has been done, in various ways, many times previously in physical science. Strictly speaking, we have no very exact knowledge of the fundamental nature of a magnetic field, and yet we do not hesitate to use the symbol H with all due familiarity. In Sec. III will be described more precisely the extent to which it has been found possible to attach physical significance to matter waves.

104. Mechanics as Geometrical Optics of the Waves.—If waves of some sort are associated with all material particles and play a part in determining their motion, then there should be a parallelism between the laws of mechanics and the laws of wave motion. For the particles certainly obey the laws of mechanics in some cases, at least. Now it was pointed out long ago by Hamilton that there does exist a "close parallelism between the laws of mechanics and the laws of ordinary *geometrical optics*," De Broglie suggests, accordingly, that the familiar laws of particle mechanics may represent an approximation which is valid under such circumstances that the laws of geometrical optics hold for the matter waves.

Now, one of the fundamental laws of ordinary geometrical optics is that, in a homogeneous medium, light travels in straight lines or rays. This assumption is very nearly correct as long as the lateral dimensions of the beam are large compared with the wave length of the light. Under these conditions one might say that the "particle" characteristics of light appear to predominate, although we now know that the rectilinear propagation of light is entirely consistent with the wave theory.—When, however, the cross section of the beam is of the same order of magnitude as the wave length of the light, rectilinear propagation no longer holds, diffraction phenomena are observed, and undulatory characteristics of light predominate. If we carry the similarity over to mechanics, might we not expect that, for very small particles of matter, the ordinary laws of mechanics—found by Newton

to be applicable to the phenomena of "macromechanics"—would fail, and, by analogy with light, we should find that matter shows undulatory properties in the realm of "micromechanics"?

If we assume that the phenomena of ordinary mechanics constitute those of geometrical optics for matter waves, then, as de Broglie showed, we can find out a great deal about the properties that these waves must have in order that, under suitable conditions, the laws of mechanics may hold. Geometrical optics may be said to be based upon two laws, the law of reflection and the law of refraction. The law of reflection as applied to matter waves presents no difficulty, for it is essentially the same as the law governing the rebound of an elastic body from a hard wall. Refraction of the waves, on the other hand, obviously corresponds to the deflection of a particle by the action of forces. It will be worth while to study this latter phenomenon in greater detail.

105. Refraction of Matter Waves.—Let us first analyze a mechanical phenomenon which imitates a simple case of refraction in geomet-

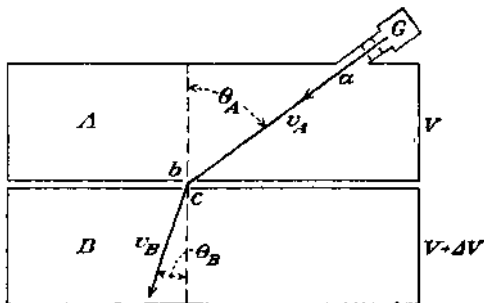


FIG. 67.—Showing "refraction" of a pencil of electrons.

rical optics. Let a pencil of electrons from a suitable gun *G* (Fig. G7), in which they are accelerated by a potential drop of V_0 electrostatic units, enter through orifice *a* an enclosed metal box *A*, the potential of which is V . Let these electrons emerge from *A* through orifice *b* into the space between box *A* and box *B*, the latter of which is maintained at a potential $+AV$ with respect to *A*. The electric field between the two boxes will give the electrons a component of velocity in a direction at right angles to the adjacent surfaces, and the electrons will enter box *B* through orifice *c* in a direction different from that in box *A*. This change of direction is closely analogous to that experienced by light in passing from one medium to another. Let v_A and v_B (Fig. 68) be the velocities of the electron beam in box *A* and box *B*, respectively; and let the corresponding kinetic energies be

$$\frac{1}{2}mv_A^2 = V_0e, \quad \frac{1}{2}mv_B^2 = (V_0 + \Delta V)e, \quad (150a,b)$$

e being the numerical charge on the electron. Since the electric field does not change the component of velocity in the x direction, we may write

$$\begin{aligned} v_A \sin \theta_A &= v_B \sin \theta_B, \\ \therefore \frac{v_B}{v_A} &= \frac{\sin \theta_A}{\sin \theta_B}, \end{aligned} \quad (151)$$

where θ_A and θ_B are the angles which the two beams make with the normal to the adjacent faces of the boxes. From Eqs. (150a,b) we

$$\frac{v_B^2}{v_A^2} = \frac{(V_0 + \Delta V)e}{V_0 e} = 1 + \frac{\Delta V}{V_0},$$

from which and from Eq. (151) we obtain

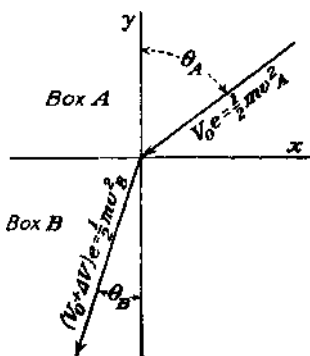


FIG. 68.

$$\frac{\sin \theta_A}{\sin \theta_B} = \frac{v_B}{v_A} = \sqrt{1 + \frac{\Delta V}{V_0}}. \quad (152)$$

Now this result might also be expressed by introducing an equivalent refractive index μ defined by the equation

$$\frac{\sin \theta_A}{\sin \theta_B} = \mu.$$

Then

$$\mu = \sqrt{1 + \frac{\Delta V}{V_0}}. \quad (153)$$

We see, therefore, that, for any given pair of values of V_0 and ΔV , the bending of the electron beam can be described by saying that the space B has a certain "electron index of refraction" with respect to space A . Or, we can proceed as in optics and introduce an index of refraction for each region. Let us assume as indices of refraction for electrons in spaces A and B , respectively,

$$\mu_A = k' \sqrt{V_0}, \quad \mu_B = k' \sqrt{V_0 + \Delta V},$$

in which k' is an arbitrary constant. Then

$$\mu = \frac{\mu_B}{\mu_A} = \frac{\sqrt{V_0 + \Delta V}}{\sqrt{V_0}} = \sqrt{1 + \frac{\Delta V}{V_0}}$$

as before. Clearly we can account for the motion of the electron along any path by assuming that, at any point where its potential is V , its refractive index is

$$\mu = k' \sqrt{V}.$$

It is important for the validity of this result that the arbitrary additive constant that is necessarily present in the potential V be so chosen that $V = 0$ at a point where the kinetic energy of the electron is zero. The kinetic energy at any other point is then

$$\frac{1}{2}mv^2 = eV.$$

If we eliminate V between the last two equations, we obtain

$$\mu = k''v,$$

where $k'' = k' \sqrt{m/2e}$ and represents a new arbitrary constant.

In this latter form, the result is applicable to the motion of any sort of particle in a field of force. For we could set up an electric field in which at each point the potential energy of an electron, $-eV$, would be the same as the potential energy of the other particle in its own force field; then a particle and an electron, projected with equal speed and in the same direction from corresponding points, would traverse similar trajectories. Thus *the refractive index necessary to account for the motion of a particle in a force field varies from point to point along the path in proportion to the velocity of the particle.*

The refractive index for waves of any sort, on the other hand, is inversely proportional to the wave speed u :

$$\mu = \frac{k'''}{u}.$$

This rule is a direct inference from Huygens' principle, which must hold for anything that we may call waves. Hence, if the waves, moving according to the laws of geometrical optics, are to follow the particle along its path, we must have at every point

$$u = \frac{b}{v} \quad (154)$$

in terms of a new constant of proportionality b , or the wave speed must be inversely proportional to the particle speed. We can also express this result in terms of the constant total energy E and the (perhaps variable) potential energy U of the particle. For

$$\frac{1}{2}mv^2 = E - U = h\nu - U,$$

if we introduce the assumption that the frequency ν of the waves is given by the Einstein relation, $E = h\nu$, h being Planck's constant. Then

$$v = \left[\frac{2(h\nu - U)}{m} \right]^{1/2}, \quad u = b_1(h\nu - U)^{-1/2}, \quad (155a,b)$$

$$\mu = \mu_0 \sqrt{h\nu - U}, \quad (156)$$

b_1 and μ_0 being new constants ($b_1 = b\sqrt{m/2}$, $\mu_0 = k'''/b_1$).

According to Eq. (156), in which U is a function of position in space, the refractive index for the waves must be supposed to vary from point to point, in somewhat the same way as it does for light waves in a nonhomogeneous piece of glass. At any given point, furthermore, u varies with v , i.e., the waves exhibit the phenomenon of dispersion. For matter waves this is true even in free space, where $U = 0$.

Parenthetically, it should be remarked, to avoid misunderstanding, that the well-known "electron microscope" does not depend for its operation upon the wave properties of the electrons. In this microscope the electrons move according to classical mechanics, although their motion is sometimes treated in terms of the equivalent

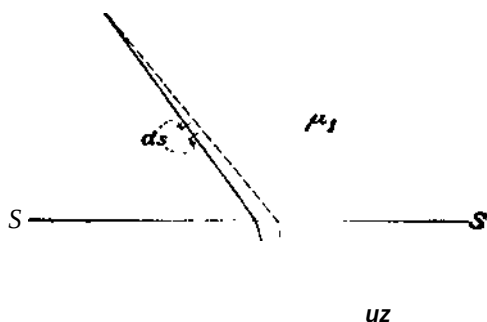


FIG. 69.—The solid line P_1P_2 represents the path taken by a beam of light; the dashed line represents a geometrically possible near-by path.

refractive index described above. The only connection that matter waves have with the electron microscope is that their wave length will ultimately set a limit to the possible resolving power that can be attained, just as the resolving power of ordinary microscopes is limited by the finite wave length of light. The wave length of the electron waves, however, we shall presently find to be so small that the resolving power of existing electron microscopes is limited by other factors.

106. Fermat's Principle and the Principle of Least Action.—For the benefit of those students who may be interested in the more abstract **formulations** of mechanics, an alternative and more complete treatment of the refraction of matter waves will also be given, along lines suggested by de Broglie.

One way of stating the law of the propagation of light in geometrical

optics is in terms of Fermat's *principle of least time*. This principle is a direct deduction from Huygens' principle and states that, of all possible paths which a ray of light might conceivably take between two given points, it chooses that path which makes the time required a minimum (more exactly, an extremum, a minimum if the points are not too far apart, otherwise, often a maximum). The principle is equally applicable to the propagation of waves of other sorts.

To express the principle in mathematical form, consider a surface SS (Fig. 69) separating two mediums of respective indices of refraction n_1 and n_2 . Let the full line represent the path actually taken by a monochromatic ray of light of frequency ν in passing from any point P_1 to any other point P_2 , the dashed line being a near-by geometrical path. Let ds represent an element of the path, and u the velocity of light of frequency ν in the medium in which ds is located. Then the time τ required for the light to travel from P_1 to P_2 is

$$\tau = \int_{P_1}^{P_2} \frac{ds}{u}.$$

Fermat's principle of least time is stated by the equation

$$\delta\tau = \delta \int_{P_1}^{P_2} \frac{ds}{u} = 0, \quad (157)$$

where $\delta\tau$ means "the difference between the value of τ for the natural path and for an infinitely near path." An alternative form of the last equation is obtained by multiplying Eq. (157) by c , the velocity of light in free space, and introducing the refractive index, $n = c/u$:

$$\delta \int_{P_1}^{P_2} \frac{c}{u} ds = \delta \int_{P_1}^{P_2} n ds = 0. \quad (158)$$

Now in form these equations are the same as those expressing the *principle of least action* in mechanics, which describes, or determines, the "natural" path taken by a body, regarded as a material particle, in its passage from one point to another under the action of conservative forces. Let the line AB (Fig. 70) represent the natural path of, say a projectile of mass m , air resistance being eliminated. Let ds be an element of length of the path and let v be the velocity of the projectile. Then the "action" A between any two points P_1 and P_2 on the natural path is defined as the line integral of the momentum along the path between those points. That is,

$$A_{P_1}^{P_2} = \int_{P_1}^{P_2} mv ds.$$

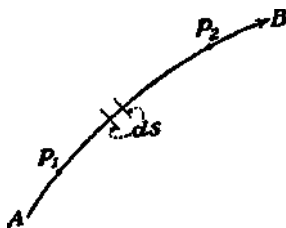


FIG. 70.

The principle of least action now states that, of all imaginable paths between P1 and P2, the projectile chooses the path which makes A a minimum (or at least an extremum) as compared with the value which A would have for any neighboring path that *might* be followed from P1 to P2. It must be assumed that all paths are traversed with the same total energy. Or, in mathematical form,

$$\delta A = \delta \int_{P_1}^{P_2} mv ds = 0,$$

where δA has a meaning analogous to that of δT in Eq. (157). Another form is obtained if we substitute for v in terms of the total energy E and the potential energy V ; since $\frac{1}{2} mv^2 = E - V$,

$$mv = \sqrt{2m(E - V)},$$

and (159) can be written

$$\delta \int_{P_1}^{P_2} \sqrt{2m(E - V)} ds = 0. \quad (160)$$

In this latter equation, E is to be kept constant as the path is varied, whereas V depends only on position in space.

These equations for the projectile become the same as those obtained previously for light waves if we suppose that

$$\frac{1}{u} = \frac{v}{b} = \frac{\sqrt{2m(E - V)}}{bm}$$

or

$$\mu = \frac{c}{u} = \mu_0 \sqrt{2m(E - V)}, \quad (102)$$

where $\mu_0 = c/(bm)$; with these substitutions, (157) becomes the same as (159), and (158) the same as (160), except for constant factors (m , b , u_0) which can be canceled out. Thus a ray of light can follow the same path as the projectile, provided the index of refraction varies along the path as given by Eq. (162). If we replace light waves by matter waves, we obtain Eqs. (154) and (156) above, in which U is written for V and the frequency of the waves is assumed to be such that $E = h\nu$; the arbitrary constant μ_0 in (156) replaces $1/4 \sqrt{2m}$ in (162).

107. The de Broglie Wave Length.—The most significant feature of matter waves will undoubtedly be their *wave length*. In general, it is this property of waves that controls the phenomena of diffraction and interference. Even in ordinary *wave optics*, the fundamental entity is the wave length; the frequency ν , which appears in the

expression $h\nu$ for the energy of a photon, which is the fundamental entity in *corpuscular optics*, is actually calculated from measured values of the wave length.

The wave length to be expected for matter waves can be inferred by developing another suggestion due to de Broglie. If we adopt the right assumption concerning the velocity or refractive index of matter waves, as we have seen, the waves can be caused to follow the same trajectory as does the particle with which they are associated, in so far as classical mechanics is valid for the particle and the laws of geometrical optics for the waves. But it follows from Eq. (154) that the waves will not actually keep up with the particle, since in general the waves and the particle move with different velocities. It is well known, however, that, whenever the wave velocity varies with the frequency (*i.e.*, whenever dispersion exists), a finite train or group of waves moves with a velocity different from that of the individual waves. This phenomenon is easily observed on water. Close inspection of a group of waves advancing over a water surface will show that the individual waves advance about twice as fast as the group as a whole; new waves continually arise at the rear of the group, pass through it, and die out at the front. In the case of matter waves, de Broglie showed that their group velocity could be assumed to be equal to the velocity of the associated particle, so that a group of them is able to accompany the particle in its motion.

It is readily shown¹ that, if u is the *wave velocity*, a finite group of waves having almost the form of sine waves with wave length X advances with a *group velocity* u' of magnitude

$$u' = u - \lambda \frac{\partial u}{\partial \lambda}. \quad (163)$$

Here a partial derivative of u is written because u may vary with the refractive index from point to point as well as with X . If there is no dispersion, as in the case of sound or of light *in vacuo*, $du/d\lambda = 0$ and $u' = u$. In the presence of dispersion, however, the group velocity and the wave velocity are not equal. In the latter case, the group velocity is obviously the velocity at which the energy travels; the individual waves function essentially as "phase waves," serving to determine, for example, where the energy shows up in an interference pattern.

Let us now postulate the equality of group velocity and particle velocity for matter waves and see to what conclusions we are led.

¹ See, *e.g.*, WOOD, R. W., "Physical Optics"; or HOUSTON, R. A., "Treatise on Light."

Then, replacing u in (163) by the particle velocity v , we have

$$v = u - \lambda \frac{\partial u}{\partial \lambda}. \quad (164)$$

But, for any type of waves, $u = \nu\lambda$; and the frequency ν of matter waves we have assumed to be such that

$$h\nu = E = \frac{1}{2}mv^2 + U, \quad (165)$$

U being the potential energy and E the total energy of the particle. Differentiating this latter equation with respect to λ , we have¹

$$h \frac{\partial \nu}{\partial \lambda} = mv \frac{\partial v}{\partial \lambda}.$$

But substitution of $u = \nu\lambda$ in Eq. (164) gives

$$v = \nu\lambda - \lambda \frac{\partial}{\partial \lambda}(\nu\lambda) = -\lambda^2 \frac{\partial \nu}{\partial \lambda},$$

after carrying out the differentiation. Elimination of $\partial \nu / \partial \lambda$ between these last two equations gives

$$v = -\frac{\lambda^2 mv}{h} \frac{\partial v}{\partial \lambda}; \quad \therefore \frac{\partial v}{\partial \lambda} = -\frac{h}{m\lambda^2}.$$

Integrating, we have the general relation

$$v = \frac{h}{m\lambda} + \text{const.}$$

If we now set the constant of integration equal to zero in this last equation, we have the product $mv\lambda$ equal to a universal constant. Then along any particular path, since $X = u/v$, mvu/v is constant, and, since m and v are constant, vu is also constant. This result is in agreement with Eq. (154). It can be shown, furthermore, that no other assumption concerning the constant of integration in the last equation can be harmonized with Eq. (154). Hence, we can write the last equation in the form

$$\lambda = \frac{h}{p}, \quad p = mv. \quad (166a, b)$$

Here p represents the ordinary momentum of the particle.

Equations (165) and (166a, b) fix the frequency ν and the wave length λ of the waves in terms of the energy E and momentum p of the particle with which they are associated. The wave velocity, if

¹ U contains "as variables x, y, z ; these are to be kept constant in the differentiation.

wanted, can then be calculated as $u = v\lambda$. Wave lengths given by Eq. (166a, b) are known as *de Broglie wave lengths*. It may be remarked in passing that Eq. (166a) holds for photons as well as for material particles; for the momentum of a photon is $h\nu/c = h\lambda$ [cf. Eq. (17) in Sec. 33]. Equation (166a) combines corpuscular and undulatory concepts in a very intimate way; for λ has a clean-cut meaning only in connection with a wave theory, and p , the momentum, is most naturally associated with a moving *particle*.

We can now compute the wave lengths to be expected for electron waves, atom waves, or molecule waves. For an electron moving at a velocity much below that of light with kinetic energy equal to V electron-volts (Sec. 96), we have

$$\frac{1}{2}mv^2 = \frac{eV}{300}, \quad p = mv = \sqrt{\frac{2meV}{300}},$$

e being the numerical electronic charge in electrostatic units. Hence by (166a) its wave length is¹

$$\lambda_e = h\sqrt{\frac{150}{meV}} = \frac{12.24}{\sqrt{V}} \times 10^{-8} \text{ cm.} = \frac{12.24}{\sqrt{V}} \text{ \AA}, \quad (167)$$

if we introduce $m = 0.9107 \times 10^{-27}$ gram, $e = 4.803 \times 10^{-10}$ e.s.u.

c being the speed of light. From this equation, if we write

$$\frac{eV}{300m_0c^2} = \rho,$$

we find

$$\left(1 - \frac{v^2}{c^2}\right)^{-1/2} = 1 + \rho, \quad v = c\sqrt{1 - (1 + \rho)^{-2}}.$$

Substituting these values in the relativistic equation for momentum, Eq. (60) in Sec. 63, we have for the momentum

$$p = \frac{m_0v}{(1 - v^2/c^2)^{1/2}} = m_0c\sqrt{(1 + \rho)^2 - 1} = m_0c\sqrt{\rho(2 + \rho)}.$$

Hence, by (166a), which is readily shown to hold in the relativistic case also, the electronic wave length is

$$\lambda_e = h\sqrt{\frac{150}{m_0eV}}\left(1 + \frac{1}{2}\rho\right)^{-1/2}.$$

Here ρ represents the ratio of the electron's kinetic energy to its relativistic rest

For 100-volt electrons, $X_c = 1.22$ angstroms; for 10,000-volt electrons, $X_e = 0.122$ angstrom.

Wave lengths can be calculated in a similar way for molecules or, for that matter, even for large masses such as billiard balls. The larger the mass, the shorter is the wave length at given speed.

It would be logical now, following the historical order, to describe at once the complete mathematical theory of matter waves, as worked out by Schrodinger. All that we need, however, for a discussion of experiments on the diffraction and interference of these waves is their wave length. It may relieve the tedium of so much abstract discussion if we describe next the experiments which have shown that electrons and even molecules actually do possess certain wave properties.

108. Experiments on Electron Waves, (a) *Reflection from a Crystal*.—The first experiments on electron diffraction were reported¹ by Davisson and Germer 3 years after de Broglie's first paper appeared. Davisson and Germer were studying the reflection of electrons from a nickel target and accidentally subjected a target to such heat treatment that it was transformed into a group of large crystals. Anomalies then appeared in the reflection from it. Following up this lead, they prepared a target consisting of a single crystal of nickel and bombarded its surface at normal incidence by a narrow pencil of low-voltage electrons; by means of a suitable "collector" of small aperture, they studied the distribution in angle of the electrons reflected from the crystal. In this reflected beam, they found striking maxima and minima, which they were able to explain in terms of diffraction of the electron waves.

The diffraction of such waves by a crystal is very similar to the diffraction of X-rays, which has been studied extensively, and is described at length in a later chapter (*cf.* Sec. 174). Here we shall discuss the structure of crystals, and the diffraction of waves by

energy mcv^2 , and the parentheses represent the effect of the variation in mass. For $p \ll 2$, i.e., $V \ll 10^6$ volts, we can write

$$\lambda_e = h \sqrt{\frac{150}{m_0 e V}} \left(1 - \frac{p}{4} + \frac{3}{32} p^2 \dots \right).$$

The correction to formula (167) is about 2.5 percent for 50,000-volt electrons and about 8.5 percent for 200,000-volt electrons. For $p \gg 2$, i.e., $V \gg 10^6$, we can use the expansion in powers of $1/p$:

$$\lambda_e = \frac{300 ch}{e V} \left(1 - \frac{1}{p} + \frac{3}{2} \frac{1}{p^3} \dots \right).$$

¹ DAVISSON and GEBMER, *Phys. Rev.*, vol. 30, p. 705 (1927).

them, only as far as is necessary in order to understand the significance of the experiments on matter waves.

Crystallographic studies by means of X-rays show that the nickel crystal is of the "face-centered cubic" type, as shown in Fig. 71,¹ i.e., the crystal can be imagined to be constructed of cubical unit cells each having an atom at each corner and one in the center of each face, with none inside the cube. The atoms are indicated by circles in the figure, certain ones being joined by lines in order to outline the unit cells. The length a_0 of the edge of the unit cube is 3.51 angstroms. Figure 71(b) shows a face cut on the crystal at right angles to one of the diagonals of the cube. In Fig. 71(c) the direction of the incident electrons is shown by the heavy arrow. The lighter arrows show,

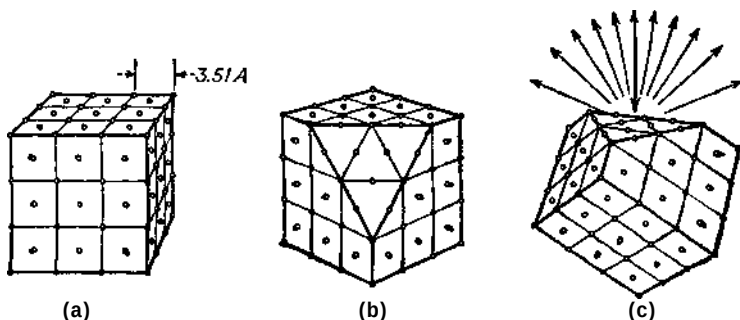


FIG. 71.—(a) A nickel crystal, showing face-centered cubic structure, (b) Same, with a face cut at right angles to a diagonal, (c) An incident beam of electrons (heavy arrow) is scattered in all backward directions.

schematically, that these electrons are reflected in all (backward) directions. Considering only the surface layer of atoms in the triangular face of Fig. 71(c), it is readily seen that these atoms are arranged in rows parallel to one side of the triangle, and it is easily shown that the distance d between these rows is

$$d = \sqrt{\frac{3}{2}} a_0 = 2.15 \text{ angstroms.}$$

Now it is fairly obvious that we may regard these rows of atoms as equivalent to the lines of a plane grating of grating space $d = 2.15$ angstroms. Radiation of wave length λ incident normally on such a grating, the plane of incidence being taken normal to one side of the triangle, should be diffracted, as is light from a reflection grating, according to the well-known law

$$n\lambda = d \sin \theta, \quad (168)$$

¹ This and certain other figures are used by permission of Dr. Davisson.

where θ is the angle between the (normally) incident beam and the diffracted beam and n is the order of diffraction [Sec. 174(6)]. These relations are shown schematically in Fig. 72. It will be observed, from Fig. 71(c) and from Fig. 72, that the crystal is in reality equivalent to a series of such plane gratings, piled one above the other and separated by a distance $b = a_0/\sqrt{3}$. Radiation penetrating to and diffracted backward from any one of these underlying layers will be combined with that diffracted from other layers, with the result that, for a given wave length λ of the (normally) incident radiation,

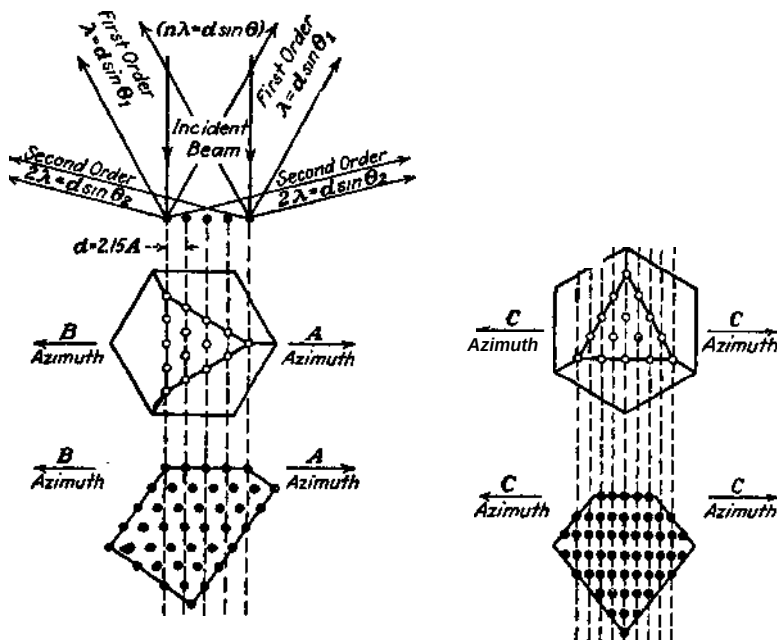


FIG. 72.—Schematic representation of reflection of electron beams from a nickel crystal.

we shall expect to find a diffracted beam at the angle θ defined by Eq. (168), unless the beams from the several layers destructively interfere. If conditions are such that destructive interference takes place, the beam to be expected at angle θ will be wholly or partially suppressed.

The apparatus used by Davisson and Germer is represented diagrammatically in Fig. 73. The "electron gun" consists of a heated filament emitting electrons, an accelerating field to give the electrons any desired velocity, and a series of collimating apertures to produce a (nearly) unidirectional beam. This monovelocity beam of electrons strikes the surface of the crystal at normal incidence,

and the electrons are reflected, or scattered, in all directions. A collector for measuring the reflected electrons is so arranged that it can be adjusted to any angular position with respect to the crystal, and thus the intensity of the electron beam reflected within a small solid angle $d\Omega$ may be determined. The collector has two walls insulated from each other. A retarding potential is applied between the inner and the outer wall so that only the fastest electrons—those possessing nearly the incident velocity—may enter the inner chamber and be measured by the galvanometer. The crystal may be turned about an axis parallel to the axis of the incident beam, and thus

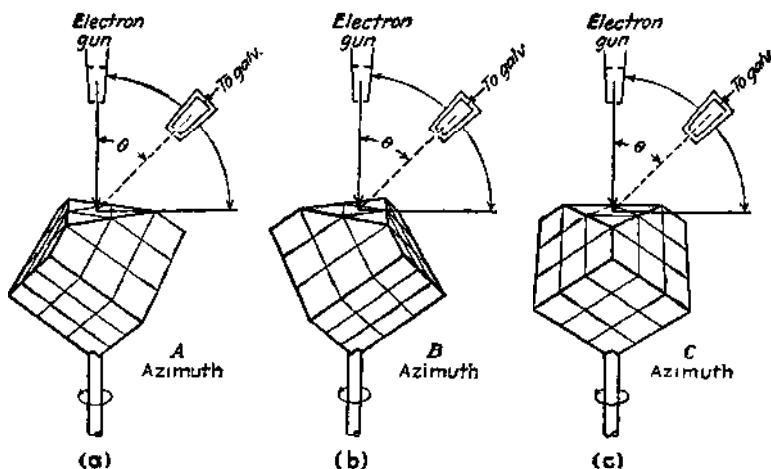


FIG. 73.—Schematic representation of apparatus used by Davisson and Germer in their discovery of the diffraction of electrons.

any azimuth of the crystal may be presented to the plane defined by the incident beam and the beam entering the collector. In what follows we shall be concerned mainly with the *A* and the *B* azimuths.

If a beam of low-voltage electrons is incident on the crystal, turned at any arbitrary azimuth, and the distribution of the scattered beam is measured as a function of the colatitude—the angle between the incident beam and the beam entering the collector—a curve similar to that in Fig. 74(a) is obtained, which refers to incident 36-volt electrons. If now the crystal is turned to the *A* azimuth, the distribution curve for 40-volt electrons [Fig. 74(6)] shows a slight "hump" at about colatitude 60° . With increasing voltage this hump moves upward and develops into a spur which becomes most prominent at 54 volts [Fig. 74(e)], at which voltage the colatitude of the spur is 50° . At higher voltages the spur gradually disappears.

The spur in its most prominent state of development offers convincing evidence for the existence of electron waves. According to Eq. (167), the wave length λ_e of a beam of 54-volt electrons should be

$$\lambda_e = \frac{12.24}{\sqrt{54}} = 1.67 \text{ angstroms.}$$

Since a diffracted beam is observed at 50° , it is obvious that, irrespective of how the underlying layers have cooperated, constructive interference has occurred among the beams diffracted from the

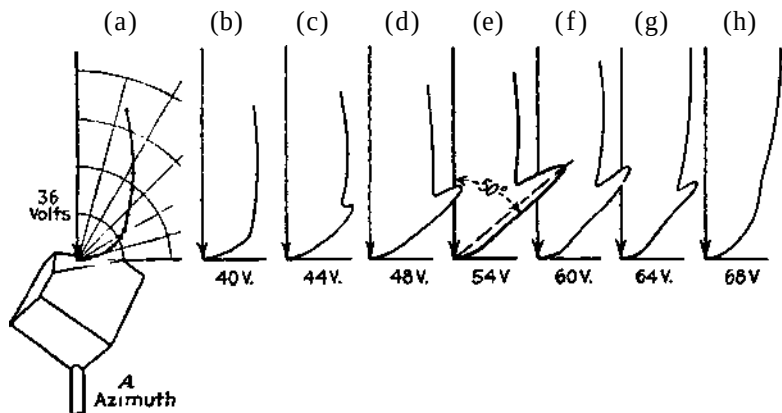


FIG. 74.—The development of the diffraction beam in the *A* azimuth, showing maximum length of "spur" for 54-volt electrons, at colatitude 50° .

several plane gratings; hence, we may use Eq. (168) to determine the wave length. We find

$$\lambda_e = d \sin \theta = 2.15 \sin 50^\circ = 1.65 \text{ angstroms}$$

This observed value of λ_e agrees excellently with that computed above from de Broglie's formula.

A similar agreement was found for other beams. In the *B* azimuth, a spur was strongly developed¹ with 65-volt electrons at a colatitude 44° . The wave length of 65-volt electrons is found in a similar way to be 1.52 Å, whereas at 44° we have $\lambda_2 = 2.15 \sin 44^\circ = 1.49$ Å. In the *A* azimuth, a spur was also found at 55° colatitude with 181-volt electrons. This was interpreted as a second-order beam. For

¹ The difference in the colatitude at which the spur is most strongly developed for the *A* and the *B* azimuth, respectively, is due to the fact that the displacement of each plane grating with respect to the one immediately above is different for the two azimuths, and hence the angle at which greatest reinforcement occurs is different in the two cases.

181-volt electrons, $\lambda_e = 0.91 \text{ \AA}$, whereas second-order diffraction gives, for colatitude 55° ,

$$2\lambda_e = 2.15 \sin 55^\circ, \quad \therefore \lambda_e = 0.88 \text{ \AA}.$$

In the C azimuth,¹ for which $d = 1.24 \text{ \AA}$, a strong beam was found at 56° colatitude for 143-volt electrons. The wave length corresponding to 143-volt electrons is $1.02 \text{ \AA}$, whereas $1.24 \times \sin 56^\circ = 1.03 \text{ \AA}$ —again an excellent agreement. Over 20 such beams were reported by Davisson and Germer for electron energies up to 370 electron-volts. In each case, the agreement between the observed and the computed wave lengths was within the limits of error of the observations. Subsequently, with improved apparatus, they reported measurements in which the agreement between observed and computed values of wave length was better than 1 percent.²

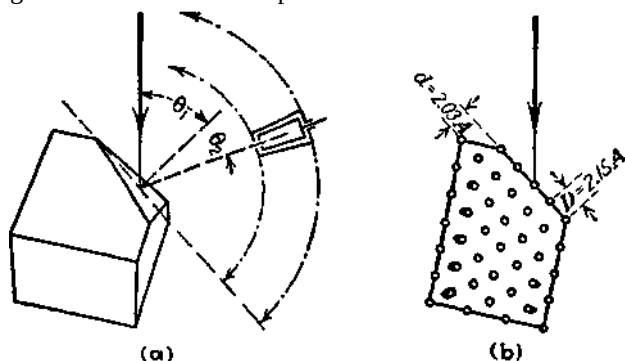


FIG. 75.—(a) Schematic representation of apparatus for studying reflection of a beam of electrons incident at angle θ_1 on the (cut) face of the nickel crystal. (b) Cross section of the crystal showing the "cooperating" planes of atoms.

The excellent, quantitative agreement between such directly observed values of λ_e with values computed from de Broglie's formula is convincing proof that the beam of electrons proceeding from filament to crystal does have wavelike characteristics as definite as those of light or X-rays; and, furthermore, that the length of the electron waves is correctly given by de Broglie's formula, Eq. (167).

(6) *Refraction in a Crystal*.—When Davisson and Germer let the electrons fall *obliquely*, instead of perpendicularly, upon the nickel crystal, they observed new effects which did not seem to agree entirely with the theory.³ Their figures illustrating this case are reproduced in Fig. 75. With a fixed angle of incidence θ_1 , electrons were observed

¹ Plane of incidence and reflection parallel to a side of the triangle.

² *Nat. Acad. Sci., Proc.*, vol. 14, p. 317 (1928).

³ DAVISSON and GERMER, *Nat. Acad. Sci., Proc.*, vol. 14, pp. 317, 619 (1928).

to be reflected at all angles θ_2 , as is illustrated in Fig. 76, which is also taken from their paper. For any given angle of incidence, however, at certain electron speeds, *selective regular reflection* occurred, as is shown by the "spurs" in Fig. 76(6), (c), and (d).

It was assumed that this selective reflection was due to constructive interference by beams reflected from successive layers of atoms in the crystal. The phenomenon is well known in the case of X-rays, which are very weakly reflected unless the conditions are met for such constructive interference. The condition for strong reflection will be deduced later for use in the more important case of X-rays; it is expressed to a high degree of approximation by Bragg's formula:

$$n\lambda = 2d \sin \theta_g, \quad (169)$$

where λ is the wave length, θ_g the "glancing angle" or angle between the incident ray and the plane of reflection (complement of the angle

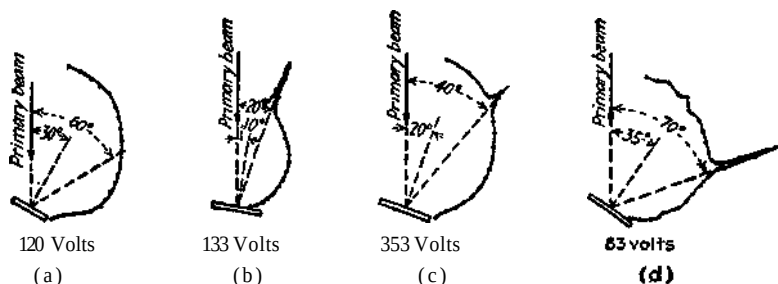


FIG. 76.—(a) Nonselective reflection of an electron beam, (6, c, d) Selective reflection at various angles of incidence

of incidence), d the distance between layers [shown as 2.03 Å in Fig. 75(6)], and n a positive integer. (See Sec. 174.) If we may assume that the electron waves penetrate a short distance into the crystal, so as to be reflected from a number of layers of atoms, the same law should hold for the selective reflection of electrons. (In this case, the spacing in a given layer, shown in Fig. 75(6) as $D = 2.15$ Å, is of no consequence, since reflection from each layer is regular.)

When Davisson and Germer calculated electron wave lengths from their data by means of the Bragg formula, however, a discrepancy was found between them and the de Broglie wave lengths. Thus, for 83-volt electrons, selective reflection was observed at an angle of incidence of 35° [Fig. 76(d)]; putting $\theta_g = 90 - 35 = 55^\circ$, $d = 2.03$ Å, in Eq. (169), we find

$$\lambda_e = \frac{3.33}{n} \text{ Å.}$$

For 83-volt electrons, Eq. (167) gives a de Broglie wave length of

$$\lambda_e = 1.34\text{A}.$$

It is impossible to find an integral n (1, 2, $\dots$) that will bring these two equations into agreement (if $n = 2$, $\lambda_e = 1.66\text{ A}$; if $n = 3$, $\lambda_e = 1.11\text{ A}$).

It was suggested by Eckart¹ and by Bethe² that this discrepancy would be explained if it were assumed that the electron beam suffers *refraction* as it enters the crystal. In the case of X-rays, the refractive index is so nearly unity that it need be taken into consideration only in very precise work (Sec. 187). In the latter case, Eq. (169) must be replaced by

$$n\lambda = 2d\sqrt{\mu^2 - \cos^2\theta_e}, \quad (170)$$

u being the refractive index for rays entering the crystal.

Equation (170) will presumably hold also for electron waves. Since we do not know u , we cannot use it to secure another check of the de Broglie relation; but, if we assume the latter relation, we can use the formula to calculate u . It is necessary, however, to make some assumption concerning the values of n . The right assumption to make might be discovered by endeavoring to secure consistency between values of u obtained from many different reflections; but it is more satisfactory to be guided by physical considerations based upon the probable cause of the refraction. In Sec. 105, we saw that electron waves should be refracted upon entering a region of different electric potential; if the potential suddenly increases by an amount AV , so that the potential energy of the electron decreases by $e AV$, the index of refraction is

$$\mu = \sqrt{1 + \frac{\Delta V}{V_0}}, \quad (171)$$

VO being the energy of the electrons before entering the region, expressed in equivalent potential units. Now photoelectric and thermionic phenomena indicate that matter *attracts* electrons; the potential energy of an electron decreases when it enters a metal, by an amount that was formerly identified with the work function but is now estimated at 10 to 20 electron-volts (Sec. 54). Hence u should be greater than unity for electron waves entering a crystal; and AV as calculated from Eq. (171) should be of the order of 10 to 20 volts.

Using the above data, i.e., $\lambda = 1.34\text{ A}$, $d = 2.03\text{ A}$, $\theta_e = 55^\circ$, if we take $n = 2$, we find, from (170), $u = 0.76$. This is less than unity

¹ ECKART, *Nat. Acad. Sci., Proc.*, vol. 13, p. 460 (1927).

² BETHE, *Naturwiss.*, vol. 15, p. 787 (1927).

$$V_0 = 83 \text{ volts, } \Delta V = 26$$

$$\approx \theta_1 = 10^\circ (\theta_0 = 80^\circ)$$

V , volts	λ , angstroms	Order of diffraction, n (assumed)	μ [from Eq (170)]	$\mu = \sqrt{1 + \frac{\Delta V}{V_0}}$ for $\Delta V = 18$ volts
64	1.52	3	1.14	1.132
130	1.07	4	1.07	1.068
215	.83	5	1.04	1.042
327	.67	6	1.02	1.027
450	.58	7	1.01	1.020
585	.50	8	1.008	1.015

Davisson and Germer. A summary of the data is shown in the first four columns of Table I. The values of the index of refraction computed by Eq. (171) for an assumed inner potential ΔV of 18 volts for the nickel crystal are shown in column 5 and are seen to be in reasonable agreement with the directly observed values of u as given in column 4.

The phenomena of the refraction of electron waves are somewhat more complex, however, than might be expected from the simplicity of Eq. (171) and the agreement of the data. For further information, the reader is referred to the literature.¹

(c) *Transmission through a Crystal.*—Electrons, if given sufficient energy, are known to pass readily through thin films of matter such as metal foil or thin mica. If an electron beam has wave properties, then in its passage through matter we should observe some or all of the phenomena characteristic of the similar passage of X-rays.

Soon after the discovery of electron waves, Kikuchi² succeeded in obtaining electron diffraction patterns by passing a pencil of electrons through a thin mica crystal. Figure 77 shows the pattern obtained by passing 68,000-volt electrons through mica of the order of 10~5 cm. thick. These observations are the exact analog of the

¹ See, for example, DAVISSON and GERMER, *Nat. Acad. Sci., Proc.*, vol. 14, p. 619 (1928); DAVISSON, *Bell System Techn. J.*, vol. 11, p. 546 (1932).

² *Japan. Jour. Phys.*, vol. 5, p. 83 (1928).

first experiments on the diffraction of X-rays by Friedrich, Knipping, and Laue [Sec. 174(6)].

Studies of metal foils with X-rays show that with appropriate treatment such foils are composed of polycrystalline material with a random orientation of crystal axes. Crystallographically, such foils are similar to crystalline powders. If in the Friedrich, Knipping, and Laue experiment with X-rays (Fig. 14G) a powder or a foil is substituted for the single crystal, the observed diffraction pattern will consist of a series of concentric rings instead of spots. Such a pattern produced by passing a pencil of X-rays through gold, is shown in



FIG. 77.—Diffraction of 68,000-volt electrons by a thin sheet of mica. [After Kikuchi, *Japan. Jour. Phys.*, vol. 5 (1928).]

Fig. 78(a). In Fig. 78(6) is shown the diffraction pattern obtained by passing 30,000-volt electrons through a film of gold; whereas Fig. 78(c) shows the pattern for 48,000-volt electrons, passing through a film of silver. Qualitatively, the similarity between the effect of X-rays and of electrons is so striking as to prove conclusively the wave nature of electrons. From the diameter of the rings in Fig. 78(6) and the dimensions of the apparatus, it is possible, knowing the length of the side of the unit cube of gold from X-ray data, to measure the wave length of the electron beam producing the pattern.

109. Diffraction of Molecule Waves.—Even *molecules* should exhibit wave properties under suitable conditions, according to the new theory. This, too, has been verified by experiment. A molecular beam is easily formed by allowing molecules of a gas to stream out of an enclosure through a small hole or slit into an evacuated chamber. Often the enclosure is heated, in order to vaporize the substance to

be studied; it is then called an "oven." Two difficulties have to be overcome, however, which do not arise in working with electrons. The molecules issue with a maxwellian distribution of velocities, whereas for diffraction experiments a beam of uniform velocity is desirable, corresponding to monochromatic waves. Then, too, neutral molecules are very much harder to detect than are charged particles.

Especially interesting is an experiment¹ performed by Estermann, Frisch, and Stern. They managed to select a beam of helium molecules or atoms having fairly uniform velocities by passing the beam through narrow slits in two parallel circular disks placed 3 cm. apart,

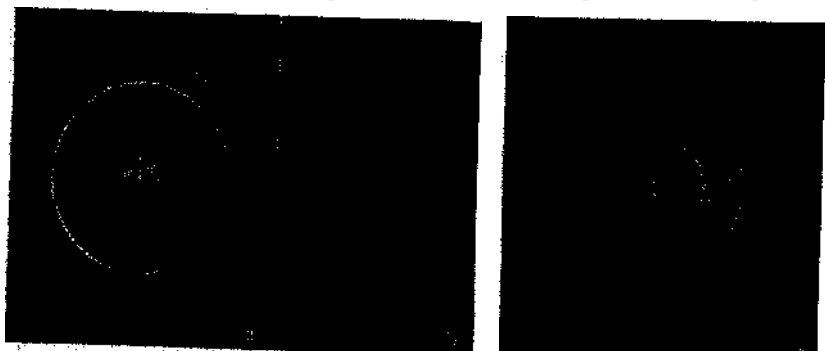


FIG. 78.—(a) Diffraction pattern produced by passing X-rays through thin **gold foil**. (Kindness of Dr. I. Fankuchen). (b) Diffraction pattern produced by passing a pencil of 30,000-volt electrons through gold foil. (From Thomson: "Wave Mechanics of Free Electrons" Fig. 196.) (c) Same for 48,000-volt electrons and a silver foil 500 Å thick. (Courtesy of Dr. L. H. Germer, Bell Telephone Laboratories, Inc.)

rigidly connected together and rotated about their common axis. The slits in the disks were adjusted so as to be opposite each other. An atom, after passing through a slit in the first disk, would arrive at the second disk too late to pass through the corresponding slit in that disk; but, if its velocity were just right, it would be in time to pass through the *next following slit*. Atoms moving faster or more slowly would arrive too soon or too late and would be stopped by the disk. After leaving the second disk, the beam fell upon the surface of a lithium fluoride crystal, by which it was reflected or diffracted. To measure the intensity of the beam diffracted in a given direction, the helium atoms were allowed to pass through a small hole into a chamber where they accumulated until a certain pressure of helium was reached (of the order of 10^{-6} cm. Hg). This pressure was measured by the cooling effect of the helium on an electrically heated metal strip, owing to conduction of heat through the helium; the electrical resist-

¹ ESTERMANN, FRISCH, and STERN, *Zeits. f. Physik*, vol. 73, p. 348 (1931).

ance of the strip was measured with a Wheatstone bridge as an indication of its temperature.

A strong diffracted beam of helium atoms was observed in addition to the regularly reflected beam. In the most precise measurement, the maximum in the diffracted beam was found at 19.45° , corresponding to a wave length of $\lambda = 0.600 \times 10^{-8}$ cm. as calculated from the usual formula for diffraction by a crystal. From the dimensions and rate of rotation of the disk, the velocity of the helium atoms was calculated to be 1.635×10^6 cm. sec.⁻¹; the corresponding de Broglie wave length is, by Eq. (166a,b),

$$\lambda = \frac{h}{mv} = \frac{hN_0}{Mv} = \frac{6.61 \times 10^{-27} \times 6.02 \times 10^{23}}{4.00 \times 1.635 \times 10^6} = 0.608 \times 10^{-8} \text{ cm.}$$

(h = Planck's constant, N_0 = Avogadro's number, M = molecular weight.) The agreement is within 1.5 percent. This result is particularly interesting, not only because it refers to an atom, but also because the velocity was measured in simple mechanical fashion instead of in some indirect manner.

110. Schrodinger's Wave Equation.—We shall return now to the discussion of the theory of matter waves. In Secs. 103 and 104, we have described the line of thought which led de Broglie to the conception of such waves. In his first paper, he managed to apply these ideas to the motion of an electron in the field of a nucleus and actually arrived at Bohr's formula for the energies of the quantum states. In order to obtain discrete states, he replaced Bohr's quantum condition by the requirement that the orbit must contain an integral number of wave lengths of the waves, so that "resonance" may occur. This is not a very satisfactory statement of the condition, however, because in such cases the wave length is so large relative to the scale of the motion that diffraction effects would be enormous. Schrodinger, refining de Broglie's idea, suggested that each quantum state represents a system of *standing waves*, or a *normal mode* of harmonic vibration, somewhat like the vibration of a violin string when sounding its fundamental or one of its overtones.

In order to find out what the normal modes of oscillation would be for the matter waves accompanying an electron, we require a more complete statement of the law governing their motion. Now the usual basis of the mathematical theory of wave motion is furnished by a *differential equation*. Thus, for a violin string vibrating in a fixed plane, we have the equation

$$d^2 \psi / dx^2 + k^2 \psi = 0$$

in which n is the displacement of any point of the string at a time t , x is the coordinate of this point measured from one end of the string, and g is a constant depending on the tension in the string and its mass per unit length. In addition to this differential equation, we have also the *boundary condition* that $n = 0$ at all times at each end of the string. To cite a second example, in the case of sound waves moving in three dimensions, the differential equation for the pressure p in terms of Cartesian coordinates x, y, z and the time t is

$$\frac{\partial^2 p}{\partial t^2} = f^2 \left(\frac{\partial^2 p}{\partial x^2} + \frac{\partial^2 p}{\partial y^2} + \frac{\partial^2 p}{\partial z^2} \right),$$

f denoting another constant. These differential equations are derived from the laws of mechanics; they are obtained by applying Newton's laws of motion to the particular medium under discussion.

We cannot follow an analogous method in arriving at the differential equation for the matter waves; for we know nothing of any medium to convey them. The best we can do is to try to guess a mathematical wave equation that will be in harmony with those properties which we have found the waves to have, and then to test our guess further by means of comparison between observations and other results deduced from this wave equation. Limitations of space and of the mathematical tools required for this book make it impossible for us to attempt here any extensive treatment of wave mechanics.¹ It may be worth while, however, to see how the properties that we have assigned to matter waves lead rather naturally to Schrödinger's differential equation for them and then to describe in general terms the application of the equation to a few important problems.

For plane harmonic matter waves of frequency ν and wave length λ moving toward $+x$, we might expect to be able to write

$$\Psi = A \sin 2\pi \left(\nu t - \frac{x}{\lambda} \right),$$

r being the quantity, whatever it is, that oscillates in the waves. Since, according to Eqs. (165) and (166a),

$$\nu = \frac{W}{h}, \quad \lambda = \frac{h}{p},$$

where W and p are, respectively, the energy and the momentum of the associated particle, we can also write the last equation in the form

$$\Psi = A \sin \frac{2\pi}{h} (Wt - px), \quad (172)$$

¹For a simple introduction, see DUSHMAN, S., "Elements of Quantum Mechanics," 1938.

in which the constants are connected by the relation,

$$W = \frac{p^2}{2m} + V;$$

here V is the potential energy of the particle, and

$$\frac{p^2}{2m} = \frac{(mv)^2}{2m} = \frac{1}{2}mv^2,$$

its kinetic energy. Now, whatever the correct differential equation for Ψ may be, it should be such that any allowable function Ψ expressed in the form of Eq. (172) is a solution of it regardless of the value of W or p . Let us endeavor, therefore, to find by trial a differential equation such that, when the expression just given for Ψ is substituted in it, W and p cancel out in consequence of the known relationship expressed in Eq. (173).

A few derivatives of M as obtained from Eq. (172) are:

$$\begin{aligned} \frac{\partial \Psi}{\partial t} &= \frac{2\pi}{h} W A \cos \frac{2\pi}{h} (Wt - px) & \frac{\partial \Psi}{\partial x} &= -\frac{2\pi}{h} p A \cos \frac{2\pi}{h} (Wt - px) \\ \frac{\partial^2 \Psi}{\partial t^2} &= -\frac{4\pi^2}{h^2} W^2 A \sin \frac{2\pi}{h} (Wt - px) & \frac{\partial^2 \Psi}{\partial x^2} &= -\frac{4\pi^2}{h^2} p^2 A \sin \frac{2\pi}{h} (Wt - px) \end{aligned}$$

The only sort of combination of these derivatives that would enable us to use Eq. (173) seems to be one in which the *first* derivative with respect to t is combined with the *second* derivative with respect to x . We can get a term in V by using Ψ itself, undifferentiated, multiplied by V . We are thus led to try as a wave equation

$$\frac{\partial \Psi}{\partial t} = a \frac{\partial^2 \Psi}{\partial x^2} + b V \Psi, \quad (174)$$

where the coefficients a and b remain to be determined.

If we substitute Ψ from (172) in this equation, however, we encounter a difficulty, since $\partial \Psi / \partial t$ contains a cosine, whereas $\partial^2 \Psi / \partial x^2$ and Ψ itself both contain a sine. However, we could equally well have started with a cosine instead of a sine; or we could use a *combination of cosine and sine*, such as

$$\Psi = A \sin \frac{2\pi}{h} (Wt - px) + B \cos \frac{2\pi}{h} (Wt - px),$$

where A and B are independent constants. The student may already be familiar with the use of such combinations as the general solution of a differential equation of the second order. If we substitute this

Ψ

$$\begin{aligned} -\frac{2\pi}{h}WB &= -a\frac{4\pi^2}{h^2}p^2A + bVA, \\ \frac{2\pi}{h}WA &= -a\frac{4\pi^2}{h^2}p^2B + bVB, \end{aligned}$$

or

$$W = a\frac{2\pi}{h}p^2\frac{A}{B} - \frac{bh}{2\pi}V\frac{A}{B} = -a\frac{2\pi}{h}p^2\frac{B}{A} + \frac{bh}{2\pi}V\frac{B}{A}.$$

These equations will agree with (173), provided

$$\begin{aligned} a\frac{2\pi}{h}\frac{A}{B} &= \frac{1}{2m}, & -\frac{bh}{2\pi}\frac{A}{B} &= 1, \\ -a\frac{2\pi}{h}\frac{B}{A} &= \frac{1}{2m}, & \frac{bh}{2\pi}\frac{B}{A} &= 1. \end{aligned}$$

Eliminating either a or b from these equations, we obtain

$$\frac{A^2}{B^2} = -1, \quad \therefore \frac{A}{B} = \pm i; \quad i = \sqrt{-1}.$$

To agree with common practice, let us take

$$\frac{A}{B} = -i.$$

Then we obtain (since $1/i = -i$):

$$-ia\frac{2\pi}{h} = \frac{1}{2m}, \quad a = \frac{ih}{4\pi m}; \quad i\frac{bh}{2\pi} = 1, \quad b = \frac{2\pi}{ih}.$$

Thus (174) becomes

$$\frac{\partial\Psi}{\partial t} = \frac{ih}{4\pi m}\frac{\partial^2\Psi}{\partial x^2} + \frac{2\pi}{ih}V\Psi$$

or, after multiplying through by $-h/(2\pi i)$,

$$-\frac{h}{2\pi i}\frac{\partial\Psi}{\partial t} = -\frac{h^2}{8\pi^2 m}\frac{\partial^2\Psi}{\partial x^2} + V\Psi.$$

For three-dimensional waves this will read

$$-\frac{h}{2\pi i}\frac{\partial\Psi}{\partial t} = -\frac{h^2}{8\pi^2 m}\left(\frac{\partial^2\Psi}{\partial x^2} + \frac{\partial^2\Psi}{\partial y^2} + \frac{\partial^2\Psi}{\partial z^2}\right) + V\Psi$$

or, in a condensed notation often used,

$$-\frac{\hbar}{2\pi i} \frac{\partial \Psi}{\partial t} = -\frac{\hbar^2}{8\pi^2 m} \nabla^2 \Psi + V\Psi, \quad (175)$$

∇^2 (or **A**) standing for the "Laplacian operator" or the sum of the three space derivatives of the second order.

Equation (175) is Schrodinger's famous *wave equation containing the time*. The agreement of results deduced from this equation justifies the belief that the equation is valid for the matter waves associated with a particle of mass m as long as relativistic effects can be neglected.

Schrodinger's equation is simple in form, containing in its coefficients only the potential-energy function for the particle, $V(x,y,z)$, and the universal constant $\hbar$. It is remarkable among the differential equations of mathematical physics in that it contains But for this factor it would resemble the equation for the flow of heat in a solid body, which likewise contains the first derivative with respect to the time. The motion of matter waves combines, in fact, some of the features of heat flow in a solid with other features resembling the propagation of mechanical disturbances, such as sound waves.

The *solutions* of the equation are likewise peculiar. As we have seen, no solution can have the form of a simple sine or cosine function, such as can be used to represent a train of sound waves in the air [cf. Eq. (172)]. The combination of sine and cosine that we found ourselves compelled to use is equivalent to a complex exponential, since $e^{\pm i x} = \cos x \pm i \sin x$.

In case V is a constant, it is easily seen by direct substitution that a solution of (175) is

$$\Psi = C e^{-\frac{2\pi i}{\hbar}(Wt - px)} \quad (176)$$

provided W and p are constants satisfying Eq. (173). Here C is an arbitrary constant, representing the wave amplitude. If $W > 0$ and $p > 0$, this solution of the wave equation represents plane waves of V traveling toward $+x$; for, at a point moving with velocity W/p , the value of SF remains constant.¹ W and p represent energy and momentum of the associated particle, respectively; the frequency ν and the wave length λ of the waves are easily seen to be $\nu = W/\hbar$, $\lambda = \hbar/p$, in accordance with conclusions reached previously. If $p < 0$ but $W > 0$, both the particle and the waves are moving toward $-x$.

If $W < 0$, as may happen if $V < 0$, the waves and the particle travel in opposite directions. We can always make this so, if we wish,

¹ If a point moves with velocity W/p , in time dt it moves a distance $dx = W dt/p$; hence, $d(Wt - px) = W dt - p dx = 0$, and so $d = 0$.

by adjusting the arbitrary additive constant that occurs in V , as in any potential energy. The possibility of reversing the direction of motion of matter waves by a mere mathematical change of this sort is a clear indication that they cannot be ordinary real waves.

111. Physical Significance of W .—The question at once arises: what is the physical significance of the mysterious Λ that appears in the mathematical theory?

An answer sometimes given to this question is that Λ is merely an auxiliary mathematical quantity that is introduced to facilitate computations relative to the results of experiment. For example, in the experiments described above on the diffraction of electrons by a crystal, the experimenter sets up an electron gun that fires electrons of a certain energy at a crystal. A detector placed at a certain angle gives indications which are taken to mean that electrons are being received. In order to develop a quantitative theory for such observations, we *assume* that a beam of electron waves of frequency $\nu = W/h$ falls on the crystal, and we *calculate* the intensity of the waves scattered in the direction of the detector. From this mathematical result we infer, following certain rules, what the indication of the detector should be. In order to make such calculations, it is not really necessary to attach any physical significance at all to the mathematical symbol Λ .

There is much to be said for such a view. After all, observational results are the primary material of physics; the purpose of theories is to correlate these results, to group them into those regularities of experience that we call laws, and to predict the results of new experiments. Even the motion of the electron as a particle is only an auxiliary concept, introduced for convenience in describing and interpreting observational results; the evidence for the particle nature of the electron is in reality very indirect. If we carefully examine the history of the concept of the electron as a particle, we shall realize that this concept owes its origin to the well-established beliefs in the corpuscular nature of matter. The electron was regarded as one of the building blocks of matter and was, therefore, visualized as a particle.

Such auxiliary concepts as particles constitute very convenient aids to thinking, and most physicists find it advantageous to make use of them. Whereas some experimental results are conveniently understood in terms of the electrons as particles, however, others can be understood only **in** terms of the interference of waves. It is worth while, therefore, to* go as **far as** we can toward assigning some physical significance to these waves, or to **the quantity** Λ in terms of which the

mathematical theory of the waves is expressed. It turns out that significance can be attached to the waves, according to circumstances, in two rather different ways.

In dealing with an experiment on scattering such as that just described, a beam of waves of indefinite total length is commonly assumed, represented by a Ψ like that in Eq. (176). The *square of the absolute value of Ψ , $|\Psi|^2$, is then taken to be proportional* (for given particle energy) *to the number of particles in the beam that cross unit area per second*, the unit area being taken perpendicular to the direction of motion.¹ Thus, in an experiment on the diffraction of electrons, let n_0 electrons cross unit area per second in the incident beam and let Ψ_0 denote the mathematical expression assumed for Ψ in the waves representing this beam. - Then, if calculation by means of the wave equation, Eq. (175), gives Ψ_c as the value of Ψ in a certain direction of diffraction, the theoretical value for the intensity of the diffracted beam in that direction, defined as the number of diffracted electrons crossing unit area per second, is

$$n_d = \frac{|\Psi_d|^2}{|\Psi_0|^2} n_0$$

[provided both the incident and the diffracted beams move in regions where V in (175) has the same value].

This procedure corresponds exactly to the method of handling similar problems in optics. The intensity of a beam of light, which might be defined as the number of photons crossing unit area per second, is proportional (in a given medium) to E^2 , the square of the electric vector (Sec. 32). The principal difference here is that, *if* being a complex number, we must use $|\Psi|^2$ instead of Ψ^2 .

There are other cases, however, as in the photoelectric effect, in which one wishes to follow the flight of a single electron or other particle. Then SF is taken to refer to a single particle. Usually, in a given case, values of r appreciably different from zero occur only within some finite region. A solution of the wave equation of this latter type is sometimes called a *wave packet*, especially if the region in which it has nonzero values changes position in space as time goes on. In such a case it is natural to ask, where is the particle in relation to the wave packet? The accepted answer may be stated in terms of probabilities. The *position of the particle* is usually considered *not to be defined any more closely than is indicated by the values of SF*. At any given instant, if a suitable observation were made, the particle might

¹ Ψ is a complex number; by the absolute value of any complex number, $x + iy$ (where x and y are real), is meant $(x^2 + y^2)^{1/2}$ written $|\mathbf{x} + i\mathbf{y}|$.

be found at any point where Ψ is different **from zero**; the *probability* of finding it in the neighborhood of a given point is proportional to the value of $|\Psi|^2$ at that point. More exactly, the *probability density* at any point is represented by $|\Psi|^2$; the *probability of finding the particle within any element of volume* $dx dy dz$ is

$$|\Psi|^2 dx dy dz. \quad (177)$$

When interpreted in this way, Ψ is sometimes called a *probability amplitude* for position of the particle.¹

The interpretation just stated imposes upon Ψ a certain mathematical requirement. For the *total probability* of finding the particle *somewhere* is, of course, unity. Hence Ψ must satisfy the condition that

$$\iiint |\Psi|^2 dx dy dz = 1, \quad (178)$$

the triple integral extending over all possible values of x , y , and z . A Ψ satisfying this requirement is said to be *normalized* (to unity). Any solution of Eq. (175) that gives a finite value to the integral in Eq. (178) can be normalized; for, because the wave equation is linear and homogeneous in Ψ , any solution of this equation can be multiplied by an arbitrary constant without ceasing to be a solution, and the arbitrary constant can then be adjusted so as to make Eq. (178) true.

There are also interesting relations between a wave packet and the *velocity* or *momentum* of the particle. A detailed study, for which we have not space, brings out the following features.

Suppose the particle is in free space (i.e., $V = 0$). Then a possible solution of the wave equation is given by Eq. (176) or

$$\Psi = C e^{-\frac{2\pi i}{h}(Wt - px)}, \quad (179)$$

where C and p are constants and $W = p^2/m$. This solution, representing an infinite train of waves, is considered to correspond to a particle moving with energy W , momentum p , and velocity $v = p/m$. It cannot be treated as a wave packet, however, for it makes

$$\iiint |\Psi|^2 dx dy dz = \infty$$

(unless $C = 0$!). We may conclude that *it is not possible for a particle to have a perfectly definite momentum or velocity*. We can, however, have as close an approach as we please to a definite momentum, for Ψ may be assumed to be indistinguishable from (179) over as large a region as

¹ The word "amplitude" is not used here in quite the same sense as in elementary physics.

desired, sinking to zero outside the boundary of this region; the constant C can then be chosen small enough so that $\iiint |\Psi|^2 dx dy dz = 1$. A wave packet formed in this way will retain its form for a long time, traveling along with a velocity roughly equal to the group velocity of waves of frequency $\nu = W/h$ or with a velocity equal to p/m .

In general, any wave packet can be expanded in terms of wave trains like (179), just as any patch of light waves can be resolved into monochromatic trains. Essentially, this amounts to representing Ψ by a Fourier integral. When this has been done in a suitable way, the coefficients of the various wave trains in the expansion constitute a probability amplitude for momentum; i.e., the square of the absolute value of any coefficient gives the probability that a suitable observation would reveal the particle as moving in the direction, and with the momentum or velocity, that is associated with the corresponding wave train. Thus, in general, according to wave mechanics, a particle does not have either a sharply defined position or a sharply defined momentum or velocity.

112. The Indeterminacy Principle.—The indefiniteness that we have just found to exist in the values of certain mechanical magnitudes associated with a particle, such as its position or momentum, is a fundamental feature of wave mechanics. The indefiniteness as to position can be minimized, to be sure, by making the wave packet very small (Δx practically zero except within a very small region); but in that case it can be shown that the packet will spread rapidly. Consequently, if we were to observe the position of the particle a little later and then calculate its velocity by dividing the distance covered by the time that has elapsed, any one of a wide variety of results might be obtained. Thus, a small packet means a large indefiniteness in momentum and velocity. On the other hand, if we give to Δx a form like that in Eq. (179) over a large region, in order to fix the velocity and momentum of the particle within narrow limits, there is a large indefiniteness in the position. In general, it can be shown that, if Δx denotes the effective range in the possible values that might be found by observation for the coordinate x of the particle, and if Δp indicates the similar range for momentum, then very roughly

$$\Delta p \Delta x \geq h, \quad (180)$$

where h is Planck's constant. This principle was first enunciated by Heisenberg in 1927,¹ who called it, in German, the principle of "Unbestimmtheit." This term has been variously translated as "indeterminacy," "indefiniteness," "uncertainty."

¹ HEISENBERG, *Zeits. f. Physik*, vol. 43, p. 172 (1927).

The principle can be said to have its basis in the *wave properties* of matter. It even has an analog in the field of optics. A single sinusoidal wave of light of wave length λ represents a certain amount of energy that is closely localized in space, but it does not constitute monochromatic light; for, upon passing through a spectroscopic slit, it will spread widely in the spectrum. To have an approach to monochromatic light, we must have a train of many waves; and this means a corresponding dispersal of the energy in space.

The conclusion that a wave packet cannot represent a particle as having at the same time a definite position and a definite momentum might seem to be in conflict with the fact that both position and momentum are capable of precise measurement. In ordinary physics, we have no difficulty in determining both of these quantities simultaneously; e.g., from two snapshots of a rifle bullet, its position and velocity at a given instant can both be calculated. Heisenberg pointed out, however, that this could be done only because, on the scale of observation used in ordinary physical measurement, the indeterminacy required by Eq. (180) is so minute as to be lost in the experimental errors. It is quite otherwise for an electron, or a molecule

Consider, for example, how an electron might be located with *atomic* precision. We might use a microscope; but then we should have to use light of extremely short wave length in order to secure sufficient resolving power. To distinguish positions 10^{-9} cm. apart, for example, we should have to use γ -rays. Under these circumstances, the effect of the light on the electron cannot be neglected. If we are to "see" the electron, at least 1 photon must bounce off it and enter the microscope. In rebounding from the electron, however, this photon will give it a strong Compton kick (see Sec. 186). Thus at the instant at which we locate the electron, its momentum undergoes a discontinuous change. Furthermore, there is an indefiniteness about the magnitude of this change, for it will vary according to the direction in which the scattered photon leaves the scene of action. We cannot limit closely the range of possible directions for the scattered photons that enter the microscope, by stopping down the aperture, without a serious loss of resolving power. A quantitative analysis of such observations leads again to Eq. (180).

It appears, then, that we cannot *at the same time* assign to an electron or other small particle, in terms of actual or possible observations, a definite position and a definite momentum (or energy). Thus, an assertion that both the position and the momentum of a particle have simultaneously certain precise values is a statement devoid of physical meaning; for, since 1900, it has become increasingly accepted

as a principle of physics that only those magnitudes which can be observed, directly or indirectly, have physical significance. Our classical notion of a particle as something that can move along a sharply defined path, having at each instant a definite position and velocity, is therefore not applicable to electrons or protons or atoms or molecules. These small bits of matter may be said to have some particle properties, but they also possess certain wave properties, so that, in the classical sense of the words, they are neither true particles nor true waves. Darwin proposed to call them "wavicles."

After this discussion of the physical significance of Ψ , we shall now return to a further consideration of the mathematical theory. Whatever changes may occur in the future in the physical ideas that are considered to underlie wave mechanics, it seems probable that the mathematical theory is here to stay.

113. Stationary or Quantum States.—So far we have discussed the motion of free particles only. Suppose now an electron or other particle is in a region where, because of a force field, it has a potential energy V depending on its position. To treat such cases, as has already been indicated, Schrodinger suggested seeking solutions of the wave equation which represent standing waves.

The distinguishing characteristic of standing waves is that the phase of the vibration in such waves is everywhere the same, whereas in running waves there exists at any moment a progression of phase along the wave train. In a mathematical expression for Ψ representing standing waves, therefore, the time must occur in a separate factor. The analogy of Eq. (176) suggests that we seek a solution of Eq. (175) of the form

$$\Psi(x, y, z, t) = e^{-\frac{2\pi i}{h} W t} \psi(x, y, z). \quad (81)$$

Here W is a constant; $\hbar = \sqrt{-1}h$ is Planck's constant, t is the time, and x , y , and z are Cartesian coordinates for the particle; Ψ , as indicated, is a function of the four variables, x , y , z , t , whereas ψ is a function of only x , y , and z . The frequency of oscillation of such a Ψ is $\nu = W/h$. For, if t is increased by h/W , the exponent of e is thereby increased by $-2\pi i$, which multiplies Ψ by

$$e^{-2\pi i} = \cos 2\pi - i \sin 2\pi = 1$$

and so leaves Ψ unchanged; hence h/W is the period of oscillation of V , or W/h is the frequency. Since by assumption the frequency of harmonic matter waves always represents the energy divided by h (Sees. 103, 104, 105), it follows that W represents the energy of the

particle. It is easily shown that Eq. (181) defines the only possible form of Ψ in which t occurs in a separate factor.

Differentiating, we find from (181) that

$$\frac{\partial \Psi}{\partial t} = -\frac{2\pi i}{h} W e^{-\frac{2\pi i}{h} W t} \psi, \quad \frac{\partial^2 \Psi}{\partial x^2} = e^{-\frac{2\pi i}{h} W t} \frac{\partial^2 \psi}{\partial x^2}, \text{ etc.}$$

Thus the operator ∇^2 affects only the factor ψ . Hence, if we insert the above expression for Ψ in Eq. (175), we obtain

$$W e^{-\frac{2\pi i}{h} W t} \psi = -\frac{h^2}{8\pi^2 m} e^{-\frac{2\pi i}{h} W t} \nabla^2 \psi + V e^{-\frac{2\pi i}{h} W t} \psi.$$

Canceling $e^{-\frac{2\pi i}{h} W t}$, we can write this result as

$$-\frac{h^2}{8\pi^2 m} \nabla^2 \psi + V \psi = W \psi$$

or

$$\nabla^2 \psi + \frac{8\pi^2 m}{h^2} (W - V) \psi = 0. \quad (182)$$

The differential equation thus obtained for the amplitude factor $\psi(x, y, z)$ is known as Schrodinger's *amplitude equation*, or, more commonly, just as *Schrodinger's wave equation*.

Not every solution of Eq. (182) can be used however. As explained in Sec. I 11, we must be able to make $\iiint |\Psi|^2 dx dy dz$ finite. Now, if Ψ has the form shown in (181), its absolute value is

$$|\Psi| = |e^{-\frac{2\pi i}{h} W t}| |\psi| = |\psi|$$

so that $|e^{-\frac{2\pi i}{h} W t}| = 1$.^{*} Thus only those solutions of (182) can be used to represent standing waves for which

$$\iiint |\psi|^2 dx dy dz$$

is finite. (The statement usually made is that ψ itself must remain finite even at infinity.) In general, this condition can be met only for certain values of W ; these are known as *characteristic values* (or *eigenvalues*). These values of W are the allowed values of the energy corresponding to the stationary states, or quantum states, of the particle in the given force field. Corresponding to each allowed value of W there are one or more different functions, obtained as solutions of Eq. (182); these are **known as characteristic functions** or *wave functions* (or *eigenfunctions*).

According to the physical significance that was assigned to Ψ in

^{*} $|e^{-is}| = |\cos x - i \sin x| = (\cos^2 x + \sin^2 x)^{1/2} = 1.$

Sec 111, the *probability* of finding the particle in a given element of space, $d\tau$, when it is in such a stationary or quantum state, is given either by $|\Psi|^2 d\tau$ or by $|\psi|^2 d\tau$, since, as we have seen, $|\Psi| = |\psi|$.

The problem of determining the allowed energy levels for a particle in a given force field thus reduces to the mathematical problem of finding the allowed solutions of a wave equation. Limitations of space and of the mathematical tools required for this book compel us to leave it to the reader to familiarize himself, by study elsewhere,¹ with the further mathematical developments of the theory. We have space only for a few descriptive remarks concerning the two cases that have already been treated in previous chapters in terms of the old quantum theory.

114. The Harmonic Oscillator.—Let x be the coordinate of a mass m moving in one dimension with a potential energy given by $V = \frac{1}{2}\beta x^2$, 0 denoting a x constant. For such a particle, the (amplitude) wave equation (182) becomes simply

$$\frac{d^2\psi}{dx^2} + \frac{8\pi^2m}{h^2}\left(W - \frac{1}{2}\beta x^2\right)\psi = 0. \quad (183)$$

A mathematical study of this equation shows that the characteristic values of W can be written

$$W_n = (n + \frac{1}{2})h\nu_c, \quad n = 0, 1, 2, \dots; \quad \nu_c = \frac{1}{2\pi}\sqrt{\frac{\beta}{m}}. \quad (184a,b,c)$$

The constant ν_c is the frequency with which the mass m would vibrate according to classical theory; but it has no special physical significance here.

The normalized wave functions for the first three quantum states, corresponding to $n = 0, 1$, or 2 , are found to be as follows, stated in terms of

$$\xi = 2\pi x \sqrt{\frac{m\nu_c}{h}}, \quad C_0 = \left(2\sqrt{\frac{\pi m\nu_c}{h}}\right)^{1/2}.$$

$$\text{For } n = 0: \quad \psi_0 = C_0 e^{-\xi^2/2}; \quad n = 1: \quad \psi_1 = \sqrt{2} C_0 \xi e^{-\xi^2/2};$$

$$n = 2: \quad \psi_2 = \frac{C_0}{\sqrt{2}} (2\xi^2 - 1) e^{-\xi^2/2}.$$

The reader can readily verify that each of these expressions for

¹ E.g., in DUSHMAN, S., "Elements of Quantum Mechanics," 1938; PAULINO, L., and WILSON, "Introduction to Quantum Mechanics," 1935; ROJANSKY, V., "Introductory Quantum Mechanics," 1938.

² Ibid.

is a solution of Eq. (183), if in the latter equation the corresponding value of W is inserted from (184a), and that they are normalized so that $\int_{-\infty}^{\infty} \psi^2 dx = 1$. (Hint: Change the variable from x to ξ even in the differential equation! Also, remember that $\int_{-\infty}^{\infty} e^{-\xi^2} d\xi = \sqrt{\pi}$.) A graph of these three wave functions is shown in Fig. 79.

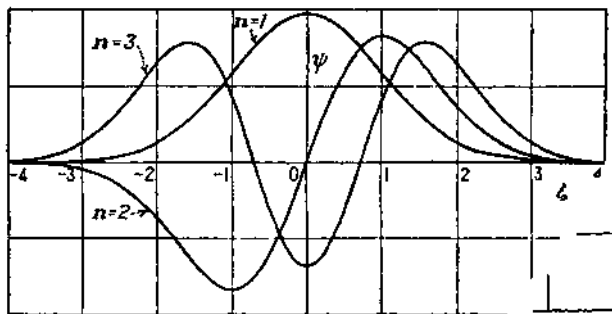


FIG. 79.—Characteristic functions for $n = 1, 2$, and 3 for a harmonic oscillator.

115. The One-electron Atom, (a) *Energy Levels and Wave Functions*.—An electron of numerical charge e , in the field of a fixed nucleus of charge Zc , has a potential energy

$$V = -\frac{Ze^2}{r}$$

at a distance r from the nucleus. Hence the wave equation (182) for an electron in such a field is

$$\nabla^2 \psi + \frac{8\pi^2 m}{h^2} \left(W + \frac{Ze^2}{r} \right) \psi = 0, \quad (185)$$

being the electronic mass (more accurately, the reduced mass—Sec. 94). This equation is more complicated to handle than that for the harmonic oscillator, because it involves three variables (x, y, z). Only a few mathematical details will be given here.

The variables can be "separated" in terms of polar coordinates, r, θ, ϕ ; solutions can then be found of the form

$$\psi_{nl\lambda} = C_{nl\lambda} e^{i\lambda\phi} P_{nl}(\cos\theta) R_{nl}(r). \quad (186)$$

Here $R_{nl}(r)$ is a certain function of r , P_{nl} denotes a certain function of $\cos\theta$, and $C_{nl\lambda}$ is a normalizing factor. [Sometimes $R(r)/r$ is written in place of $R(r)$.] The symbols n, l , and λ denote three integers which appear to occur in the customary expressions for the functions; these integers may have any values such that

$$|\lambda| \leq l < n.$$

Thus the possible values of n are $n = 1, 2, \dots$; for each n , $l = 0, 1, \dots, n - 1$; for each l , $\lambda = -l, -l + 1, -l + 2, \dots, l - 1, l$. (More commonly the letter m , or m_l , is used in place of λ . The notation used here has certain advantages which will appear later in the study of many-electron spectra.) The numbers n , l , and λ serve as convenient labels for the wave functions and are commonly called "quantum numbers."

The corresponding values of the energy turn out to be the same as those found by Bohr (Sec. 94):

$$W = -\frac{2\pi^2 me^4 Z^2}{h^2 n^2}.$$

This problem illustrates an important feature that is frequently met with called *degeneracy* of the wave functions or quantum states. For any value of n above 1, there are several independent wave functions for each allowed value of the energy. Corresponding to given n , the number λ ranges over $2l + 1$ different values from $-l$ to l ; and, for given n , l ranges from 0 to $n - 1$; and for each combination of a value of l with an allowed value of λ there exists a distinct wave function.

(b) *Relations with Angular Momentum.*—When the atom is in a definite quantum state, the probability density, $|\Psi|^2$ or $|\psi|^2$, is everywhere constant in time. Thus it might be said in one sense that the electron is not in motion. There is at least no suggestion of motion along an orbit. Certain phenomena suggest, however, that, provided $l > 0$, the atom should be regarded as possessing moment of momentum or *angular momentum* about the nucleus; and this momentum is associated with magnetic effects in the same way as is that of the classical motion in an orbit. Accordingly, for convenience, we shall speak of the electron as "moving" in the field of the nucleus, and we may sometimes even refer to this motion as "orbital."

The value of the angular momentum furnishes a useful means of distinguishing between states that have the same energy. Before proceeding further, the treatment of angular momentum in classical theory may be recalled.

The student is doubtless familiar with the interpretation of angular momentum as a vector quantity. When a rigid body is rotating about a fixed point, like a top about its point, there will be a certain axis through the fixed point about which the angular momentum of the body is a maximum. About any other axis, inclined at an angle θ to the line of maximum angular momentum, the angular momentum is equal to the maximum value multiplied by $\cos \theta$. The angular

momentum can be represented by a vector drawn in the direction of the axis about which it is a maximum; the length of this vector can be made to represent the magnitude of the maximum angular momentum, and its direction, indicated by an arrowhead, is commonly taken in the direction in which a right-handed screw would advance along the axis while turning about it in the same direction as the rotating body. The angular momentum about any other axis is then represented by the component of this vector in the direction of that axis. The vector may or may not coincide with the instantaneous axis of rotation of the body.

A similar treatment can be given to the angular momentum of a moving particle. In classical theory, the electron can be regarded

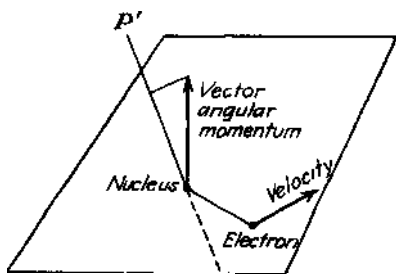


FIG. 80.—Illustration of the vector angular momentum of an electron due to its motion about the nucleus.

as moving at every instant in a certain plane drawn through the nucleus; the vector representing its angular momentum about the nucleus is commonly drawn perpendicular to this plane and in the direction specified in the last paragraph (see Fig. 80). The angular momentum about any oblique axis such as PP' , being equal to the component in that direction of the vector angular momentum, is represented by the projection on the

given oblique axis of the directed line that represents the vector angular momentum itself.

Now, it can be shown that a given wave function $\psi_{n\lambda}$ represents a special state of the electron in which its angular momentum about the axis of polar coordinates has the definite value

$$\lambda \frac{h}{2\pi}.$$

Thus, in the set of $2I + 1$ quantum states that exists for a given n and I , the angular momentum about the polar axis ranges by integral steps from a maximum of I units, each of magnitude $\frac{h}{2\pi}$, in one direction through zero to I units in the opposite direction. Such a variation of the angular momentum from one state to another is represented if we imagine that the angular momentum of the electron has a fixed value of I units but is inclined at various fixed angles to the axis. This is illustrated in Fig. 81, in which the vector representing the

angular momentum is assumed to have a length of 2 units and is shown in five alternative positions, corresponding to $\lambda = -2, -1, 0, 1, 2$. Only for $\lambda = \pm 2$ does the vector coincide with the polar axis. A *vector diagram* so constructed serves vividly to illustrate the relation between the quantum number l and the various allowed values of λ .

Actually, however, according to wave mechanics, the total angular momentum of the electronic motion about the nucleus (more exactly, about the center of mass of the atom), is not equal to $\hbar/2\pi$ but has the rather curious value $\sqrt{l(l+1)} \hbar/2\pi$. Furthermore, it has no fixed vector direction, as it has in classical theory. The vector angular momentum can be regarded, to be sure, as making a fixed angle with the axis, for, in wave mechanics, the magnitude of its component perpendicular to the axis has a fixed value. But, if we draw a fixed line PP' perpendicular to the axis of polar coordinates and consider the component of the angular momentum in the direction of this fixed line, then this component has no definite value.

The component of angular momentum in the direction of such a line PP' is susceptible of measurement, according to the accepted view, and the result of such a measurement might be the value 0, or it might be a positive multiple of $\hbar/2\pi$, or a negative multiple. The average of many such measurements would almost certainly be close to zero, and closer as the number of measurements is increased; this fact is expressed by saying that the value of the component of momentum in this direction has a mean expectation of zero. An analogous situation would exist in classical mechanics if the vector representing the angular momentum were to revolve at uniform velocity about the chosen axis of polar coordinates; then its component in the direction of the line PP' , drawn perpendicular to this axis, would have an average value of zero. Nothing in classical or wave mechanics suggests such a revolution of the vector, however; rather, the revolution furnishes a means of visualizing the fact that an assignment of the magnitude of the angular momentum and of its component along a given axis (the axis of polar coordinates) does not completely fix the vector angular momentum but, in classical mechanics, leaves it free to assume any azimuth about the axis and, in wave mechanics, leaves it indeterminate. Any given quantum state of the electron thus corresponds,

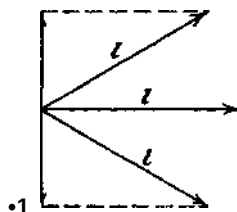


FIG. 81.—Illustrating the quantization of angular momentum for $l=2$.

not to a single classical motion with the vector angular momentum in a definite direction, but to a sort of average of many motions, each associated with a different position of the vector momentum in azimuth around the axis.

Sometimes a little more information is put into the vector diagram by drawing the vector of a length to represent $\sqrt{l(l+1)}$ units of magnitude instead of l units; then the vector necessarily remains somewhat inclined to the axis even when $\lambda = \pm 1$, since in this case the component along the axis represents only l units. This is a complication in the diagram that is of small value, however. The most important use of the diagram is to exhibit the relation between the quantum numbers l and X , and this relation is adequately exhibited when the vector is drawn of a length to represent l itself.

A separate diagram is to be drawn for each value of l . The relationships involved here are sometimes referred to as *space quantization*.

No analogous interpretation can be given for the quantum number n , which is sometimes called the *principal* quantum number. It should be remarked also that in the simplified theory that is here described no allowance has been made for electronic "spin" (see Sees. 117 and 135).

(c) *Degeneracy of the Wave Functions*.—The axis for the polar coordinates can, in general, be drawn in any desired direction. Thus the set of quantum states and of associated wave functions is in some degree a matter of arbitrary choice. This arbitrariness is one aspect of the degeneracy of the wave functions, which deserves more detailed discussion.

When several wave functions are associated with the same value of the energy W , then any linear combination of them is another possible wave function for the same energy. Thus, let $\psi_1, \psi_2, \psi_3 \cdots \psi_k$ be solutions of Eq. (185), for $W = W_l$. Then the linear combination ψ' where

$$\psi' = c_1\psi_1 + c_2\psi_2 + c_3\psi_3 \cdots + c_k\psi_k, \quad (188)$$

$C_1, C_2, C_3, \dots, C_k$ being any constants, is also a solution of Eq. (185), as may be verified by substitution. Thus ψ' is as valid a wave function, and represents just as valid a quantum state for the energy W , as do $\psi_1, \psi_2, \dots, \psi_k$. The possible wave functions for W_l are thus infinite in number. It is possible, however, to find a number k of them in terms of which all others can be written as linear combinations. For Eq. (185), as already noted, the number of independent wave functions associated with a given value of W , or with a given value of n , is n^2 . The n^2 independent wave functions can be chosen in many ways, but their number is always the same.

If the axis of polar coordinates is taken in a different direction, the factor $R_{nl}(r)$ in the expression for $\psi_{nl\lambda}$ as given in (186) is the same as before, but the geometrical significance of the directional factors $e^{i\lambda\phi}P_{\lambda l}(\cos\theta)$ is different. It can be shown, however, that the new directional factor can be written as a linear combination of the directional factors relative to the original axis, and, in fact, in terms of those factors, $2l + 1$ in number, which are labeled with the same value of l .

The infinite variety of the quantum states for given n and l corresponds to the infinity of different possible orientations of a Bohr orbit, and it is closely related to the isotropy of space.

(d) *Probability Density and Charge-cloud Density*.—The probability density corresponding to one of the wave functions $\psi_{nl\lambda}$ is easily seen from Eq. (186) to be

$$|\psi_{nl\lambda}|^2 = |C_{nl}|^2 [P_{\lambda l}(\cos\theta)]^2 [R_{nl}(r)]^2. \quad (189)$$

There is little suggestion here of anything like *orbital motion*. The quantity $|\psi_{nl\lambda}|^2 dr$ represents the probability that, if the electron were located experimentally at any instant, it would be found in the element of volume dr (Sec. 111). It would not be possible, however, to follow the electron in an orbital motion around the nucleus by means of a succession of such observations, as the astronomers follow the planets around the sun; for one observation with a γ -ray microscope of sufficient resolving power would suffice, because of the Compton effect, to knock the electron entirely out of the atom.

Often it is convenient to imagine the electronic charge e to be distributed in space as a sort of *charge cloud*, with a density n of magnitude

$$\eta = e|\psi|^2.$$

Many effects of the atom on its surroundings are approximately the same as if the atom actually contained a distribution of charge of density η , acting according to the ordinary laws of electrostatics. In any case, this furnishes a convenient means of visualizing the probability distribution.

Owing to the absence of the coordinate ϕ in the right-hand member of Eq. (189), it is evident that, when the electron is in one of its quantum states, the probability density or charge-cloud density is symmetrical about the axis of polars. The variation with θ , on the other hand, may be large. This variation is of minor interest, however, and for this reason it is more illuminating to consider the *average radial* variation in the density. Let $P_r dr$ denote the total amount of probability, or $n_r dr$ the numerical amount of charge, con-

$$\int_0^\infty r^2 R_{nl}^2 dr = 1$$

$$|C_n|^2 \int \int [P_n(\cos \theta)]^2 \sin \theta d\theta d\phi = 1),$$

$$P_r = r^2 R_{nl}^2,$$

$$\eta_r = eP_r = er^2 R_{nl}^2.$$

$$a_0 = \frac{\hbar^2}{4\pi^2 m e^2} = 5.27 \times 10^{-9} \text{ cm.};$$

$$n = 1, \quad l = 0, \quad \lambda = 0: \quad \psi_{100} = \frac{1}{\sqrt{4\pi}} R_{10}, \quad R_{10} = 2 \left(\frac{Z}{a_0} \right)^{3/2} e^{-Zr/a_0};$$

$$n = 2, \quad l = 0, \quad \lambda = 0: \quad \psi_{200} = \frac{1}{\sqrt{4\pi}} R_{20},$$

$$R_{20} = \frac{1}{\sqrt{2}} \left(\frac{Z}{a_0} \right)^{3/2} \left(1 - \frac{Zr}{2a_0} \right) e^{-\frac{Zr}{2a_0}};$$

$$l = 1, \quad \lambda = 0: \quad \psi_{210} = \sqrt{\frac{3}{4\pi}} \cos \theta \cdot R_{21},$$

$$R_{21} = \frac{1}{2\sqrt{6}} \left(\frac{Z}{a_0} \right)^{3/2} \frac{Zr}{a_0} e^{-\frac{Zr}{2a_0}};$$

$$\lambda = \pm 1: \psi_{21,\pm 1} = \sqrt{\frac{3}{8\pi}} e^{\pm i\varphi} \sin \theta \cdot R_{21};$$

for $n = 3$,

$$R_{30} = \frac{2}{3\sqrt{3}} \left(\frac{Z}{a_0} \right)^{3/2} \left[1 - \frac{2Zr}{3a_0} + \frac{2}{27} \left(\frac{Zr}{a_0} \right)^2 \right] e^{-\frac{Zr}{3a_0}},$$

$$R_{31} = \frac{4}{27} \sqrt{\frac{2}{3}} \left(\frac{Z}{a_0} \right)^{3/2} \left(1 - \frac{1}{6} \frac{Zr}{a_0} \right) \frac{Zr}{a_0} e^{-\frac{Zr}{3a_0}},$$

$$R_{32} = \frac{4}{81} \sqrt{\frac{1}{30}} \left(\frac{Zr}{a_0} \right)^2 e^{-\frac{Zr}{3a_0}}.$$

The constant a_0 is the same number that represented the radius of the smallest orbit for hydrogen [cf. Eq. (135)] in the Bohr theory.

For $Z = 1$, these formulas refer to hydrogen; for $Z = 2$, to ionized

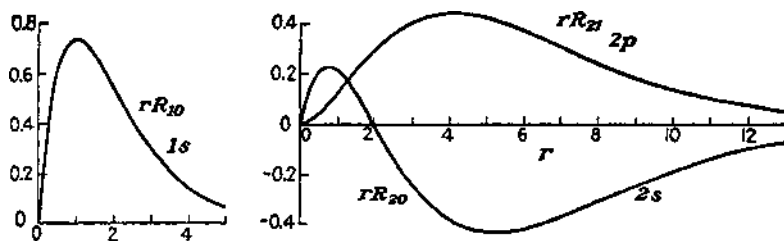


FIG. 82.—The radial function rR_{nl} for several quantum states of the hydrogen atom.

helium; and so on. In Fig. 82 are plotted, for hydrogen, not R_{nl} but rR_{nl} ; and in Fig. 83 are shown the corresponding curves for P_r or n_r , which is proportional to $r^2 R_{nl}^2$. The areas under the latter curves

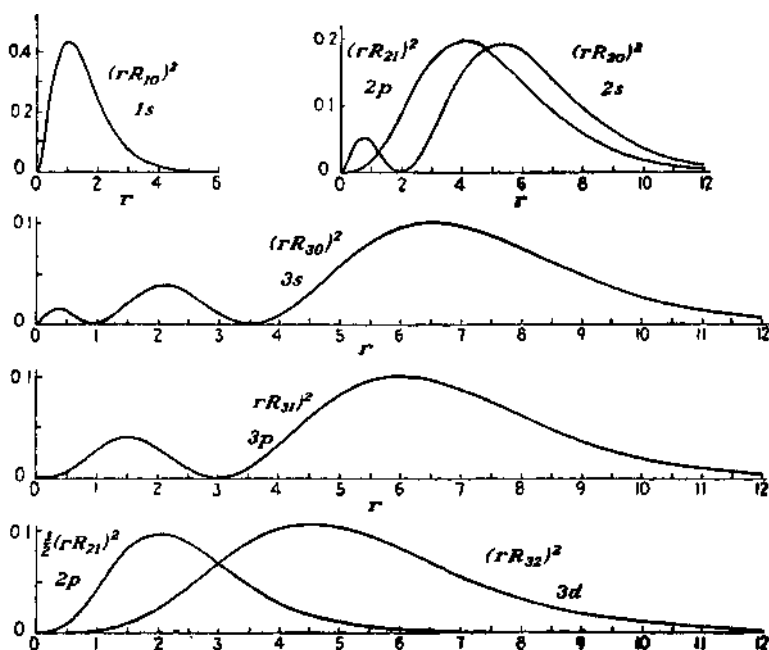


FIG. 83.—The radial probability density $P_r = r^2 R_{nl}^2$ for several quantum states of the hydrogen atom (one curve being drawn twice for comparison).

would all be equal if they were plotted to the same scale. The curves are labeled in the notation of the spectroscopists, the letters $s, p, d, f, \dots$ being used to indicate $l = 0, 1, 2, 3, \dots$, preceded by a

number giving the value of n . The meaning of these letters should be learned:

1-0	1	2	3	4	5	6	7	-	-	-
s	p	d	f	g	h	i	k	-	-	-

The reason for the choice of letters is stated in Sec. 131.

Although there is nothing in the wave functions that suggests the former Bohr orbits, a certain correspondence between them can be traced. The Bohr orbit has one more unit of angular momentum than has the corresponding state, viz., k units of $\hbar/2\pi$ where $k = l + 1$. In the normal or lowest energy state (or "ground state"), there is no angular momentum, according to wave mechanics, as against 1 unit in the Bohr theory. States with $l = n - 1$ correspond to circular orbits, the other states to elliptical orbits of various eccentricities; the s state, $l = 0$, corresponds to the ellipse of maximum eccentricity for given n . The various values of λ correspond to various inclinations of the plane of the orbit to the axis, the plane being parallel to the axis for $\lambda = 0$ and perpendicular to it for $\lambda = \pm l$.

A further connection with the Bohr orbits is found to be that, in states with $l = n - 1$, the maximum value of the radial probability density P_r or of the charge density n_r occurs at a value of r equal to the radius of the corresponding Bohr circle. In these states with $l = n - 1$, as is evident from Fig. 83, the density is crowded into a single large hump; it is most widely distributed in the s states ($l = 0$).

From the formulas, it is evident that for s states ψ does not vanish at the nucleus ($r = 0$). This fact tends to give peculiar properties to s states.

As n increases, the curves for R_{nl} , or for P_r or n_r , spread out more and more, just as did the Bohr orbits; the atom swells, so to speak. If Z increases, on the other hand, it is evident from the formulas that the atom shrinks in inverse ratio to Z .

As a final remark, it may be stated that the motion of the nucleus introduces exactly the same effect as we found it to do on the Bohr theory: the electronic mass m is merely replaced in all formulas by the reduced mass (Sec. 94),

$$m' = \frac{mM}{m + M}$$

(M = mass of nucleus).

116. Emission and Absorption of Radiation.—One of the most important properties of atoms and molecules is their ability to emit or absorb radiant energy. We cannot take up here the mathematical theory of this process; but we may describe the quantities that are

calculated theoretically **with** the **help of** wave mechanics, and a word or two may be added in regard **to** the method of procedure.

The most comprehensive method of treating radiation problems is to resolve the electromagnetic field inside a large box into standing-wave systems or harmonic modes of oscillation, just as we did in Sec. 80 for the purpose of deducing the laws for black-body radiation. Each mode of oscillation of the field is then treated as a harmonic oscillator, the wave-mechanical theory described in Sec. 114 being used for it; and radiation problems are solved by calculating the effect of the interaction between these field oscillators and any atoms or molecules that may be immersed in the field.

(a) *Einstein's A and B Coefficients. Mean Life.*—Problems in the emission of radiation center about the determination of constants known as the *probabilities for spontaneous radiative transition* or *Einstein's A coefficients*. These quantities can be defined as follows. Suppose that an atom is in its quantum state number n and that state number j has lower energy. Then, in any short interval of time dt , there will be a certain probability that the atom will jump from state n into state j , the difference in energy between the two states being emitted as a photon of radiation. This probability is denoted by $A_{nj} dt$. Or, we can say that out of a large number N of atoms in state n , $NA_{nj} dt$ will jump into state j during the interval of time dt . Thus A_{nj} represents the probability per second of a spontaneous jump from state n to state j accompanied by the emission of a photon.

The A coefficients are connected with the *mean life* of an atom in a given quantum state, an idea that is often useful. Suppose that N_0 atoms start in state n at time $t = 0$, and that after a time t all have jumped into other states except N of them. Then, during the next element dt of time,

$$N \sum_j' A_{nj} dt$$

of them will jump, where $\sum_j'$ denotes a summation over all states of lower energy only. Thus, during dt , N changes by dN where

$$dN = -N\gamma dt, \quad \gamma = \sum_j' A_{nj},$$

so that, after integrating,

$$N = N_0 e^{-\gamma t};$$

here we have **chosen the constant of integration so that** $N = N_0$

when $t = 0$. (Cf. Fig. 84). Let us calculate the average time spent by all N_0 atoms in state n , which is called the *mean life* of an atom in that state; denote it by τ_n . Since $-dN$ or $N\gamma dt$ of them jump after a time t spent in state number n , the average time spent in that state is

$$\tau_n = \frac{1}{N_0} \sum t(-dN) = \frac{1}{N_0} \int_0^\infty N\gamma t dt,$$

$$\tau_n = -te^{-\gamma t} \Big|_{t=0}^\infty + \int_0^\infty e^{-\gamma t} dt.$$

But $te^{-\gamma t} \rightarrow 0$ as $t \rightarrow 0$ or as $t \rightarrow \infty$; and $\int_0^\infty e^{-\gamma t} dt = 1/\gamma$. Hence

$$\tau_n = \frac{1}{\gamma} = \left(\sum_j' A_{nj} \right)^{-1}. \quad (190)$$

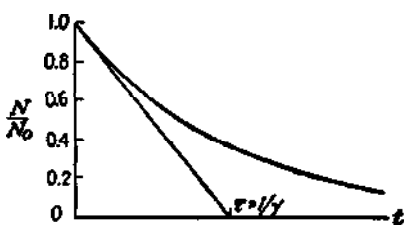


FIG. 84.—Illustrating exponential decrease by radiative transitions, $T = T_n =$ mean life.

tion falling on a vibrating oscillator causes it either to gain or to lose energy depending upon the phase relation between the waves and the vibration of the oscillator.) The quantity B_{ni} often called "Einstein's B coefficient" might be called the '*probability coefficient for an induced radiative transition*'.

Although A_{ni} and B_{ni} seem to be quite distinct from each other, theory indicates that a simple general relationship should exist between them, viz.:

$$B_{ni} = \frac{c^3}{8\pi h \nu_{nj}^3} A_{ni}, \quad (191)$$

where h is Planck's constant and c the speed of light. Thus, the ratio B_{ni}/A_{ni} is the same for all radiating systems. It follows that for different atoms or molecules the ability to absorb and the ability to emit radiation of any given frequency are proportional to each other. This constitutes an extension to gases of Kirchhoff's law (Sec. 70).

(b) *Einstein's Deduction of Planck's Law.*—The coefficients A and B were originally introduced by Einstein in giving a deduction of Planck's law for thermal radiation based upon the processes of emission and absorption.¹ The proof is very simple and is interesting.

In an isothermal enclosure, let there be N_n molecules in state n and N_i molecules in state j , and let W_n , W_j denote the molecular energies for these two states. Then, when thermal equilibrium obtains, according to the Boltzmann formula [Eq. (148a) in Sec. 101],

$$N_n = C e^{-W_n/kT}, \quad N_j = C e^{-W_j/kT},$$

C being a constant of proportionality. The existence of equilibrium requires, now, that as many atoms jump in a second from n to j , with the emission of a photon, as jump from j to n , accompanied by the absorption of a photon. Hence, in terms of the probability coefficients just defined.

$$N_n (A_{ni} + B_{ni} u_\nu) = N_j B_{nj} u_\nu;$$

whence

$$u_\nu = \frac{A_{ni}}{B_{nj} (N_j/N_n) - 1}.$$

But, by the preceding equations,

$$\frac{N_j}{N_n} = e^{h\nu/kT}, \quad \nu = \frac{W_n - W_j}{h}.$$

Hence, using Eq. (191) with ν_{nj} replaced by ν , we

$$u_\nu = \frac{8\pi h \nu^3}{c^3} \frac{1}{e^{h\nu/kT} - 1}.$$

¹ EINSTEIN, *Phys. Zeits.*, vol. 18, p. 121 (1917).

This is readily shown to represent Planck's radiation law expressed in terms of frequency instead of wave length. For the frequency ν and wave length λ are related to the speed of light c by the equation $\nu\lambda = c$; hence, as ν and λ vary, $\nu d\lambda + \lambda d\nu = 0$. If we omit the negative sign, taking $d\lambda$ and $d\nu$ both to be positive, we may write

$$\nu d\lambda = \lambda d\nu.$$

Now the same energy which is represented here by $u_\nu d\nu$ was represented by $\psi_\lambda d\lambda$ in Chap. VI. Hence,

$$\psi_\lambda d\lambda = u_\nu d\nu.$$

From these two equations we have

$$\psi_\lambda = \frac{\nu u_\nu}{\lambda}.$$

Inserting here the value just obtained for u_ν and then $\nu = c/\lambda$,

$$\psi_\lambda = \frac{8\pi ch}{\lambda^5} \frac{1}{e^{ch/\lambda kT} - 1},$$

which agrees with Eq. (118) in Sec. 85.

(c) *Calculation of Transition Probabilities.*—When theoretical values of the transition probabilities are calculated by the method indicated at the beginning of this section, they are found to depend upon certain quantities called *matrix components* between the wave functions of the two states that are involved in the transition.

The most important term in the expression that is obtained for A_{ni} from wave mechanics, for an atom containing 1 electron, is as follows:

$$A_{ni} = \frac{64\pi^2 e^2 \nu_{ni}^3}{3hc^3} (|x_{ni}|^2 + |y_{ni}|^2 + |z_{ni}|^2). \quad (192)$$

Here e = electronic charge, h = Planck's constant, c = speed of light, and ν_{ni} is the frequency of the radiation that is emitted as the electron jumps from state n to state i ; x, y, z are the coordinates of the electron and

$$x_{ni} = \iiint x \psi_n^* \psi_i dx dy dz,$$

where ψ_n and ψ_i are the wave functions for the two states, y_{ni} and z_{ni} being given by similar integrals with x replaced by y or z , respectively. The symbol ψ_n^* denotes the complex conjugate of the value of ψ_n , which is in general a complex number. The numbers x_{ni} , y_{ni} , z_{ni} are called *matrix components* of x , y , or z , respectively, between the functions ψ_n and ψ_i . They vary according to the values of the two indices,

n and j , and the possible values of each of them might be written as a matrix with n numbering the rows and j the columns.

It is of interest to compare the expression just written for A_{ni} with that for the emission of radiation by a classical oscillator. In Eq. (34) in Sec. 35, for the energy radiated per second by a classical harmonic oscillator, consisting of a charge q vibrating in linear harmonic motion with amplitude a , we obtained

$$\frac{16\pi^4\nu^4q^2a^2}{3c^3} \text{ ergs.}$$

If we divide this expression by the amount of energy in a photon, $h\nu$, we have

$$\frac{16\pi^4\nu^3q^2a^2}{3hc^3}$$

as the number of photons of energy emitted, by the oscillator per second. This agrees with the value given in Eq. (192) for A_{nj} , which is the probability per second of the emission of a photon, provided we drop y_{ni} and z_{nj} , so as to have an expression referring to one-dimensional motion, and assume that $q = e$, $\nu = \nu_{nj}$, and

$$a = 2x_{ni}.$$

Because of such correspondences with classical theory the term in A_{nj} that is given in (192) is said to refer to *dipole radiation* by the atom. The average rate of emission of dipole radiation by atoms in jumping from state n to state j is the same as if the atoms contained classical oscillators of the corresponding frequency vibrating with amplitudes equal to twice the matrix components of the electronic coordinates between the two states.

There are also less important terms in the complete expression for A_{ni} . The dipole term, given in Eq. (192), is usually the principal term, unless it actually vanishes. If the dipole terms vanish, the transition from level n to level j cannot occur by dipole emission of radiation. Such transitions are said to be "forbidden" so far as dipole emission is concerned. "Selection rules" stating which transitions are forbidden are of great importance in spectroscopy. A small probability of transition may exist, however, owing to the other terms in A_{ni} . The next most important term is called the "quadrupole" term.

117. Relativistic Effects and Electron Spin.—The theory described so far has been a nonrelativistic one suited to cases of small energy only. Like Newtonian mechanics, it requires modification to bring it into harmony with the principles of relativity. We cannot say much

here about the relativistic form of wave mechanics; but it happens that its principal new feature can readily be described in terms of a physical picture. This feature was, in fact, discovered before the advent of wave mechanics.

In 1925, Uhlenbeck and Goudsmit pointed out¹ that certain features in atomic spectra could be explained if it was assumed that the electron "spins," or rotates about an axis through its center of mass, and that it has both *angular momentum* and a *magnetic moment* associated with this rotation. The angular momentum is constant in magnitude, but because of it the electron possesses two *internal* quantum states. These states can be so chosen, with reference to any axis, that, when the electron is in one of these quantum states, it has internal angular momentum about the axis equal to $\frac{1}{2}$ unit or $h/4\pi m$ in one direction; whereas, when it is in the other state it has an equal amount of angular momentum in the opposite direction. When the electron is in an atom, its angular momentum of spin is to be added to its angular momentum due to orbital motion. The interaction of the associated magnetic moment with electric or magnetic fields, including that of the nucleus, or with the magnetic moments of other electrons, modifies the atomic energy levels. Examples of such effects will be described in the next chapter.

An astonishing new turn was given to the theory in 1928 by Dirac. He showed² that the most natural way to bring wave-mechanical theory into harmony with the theory of relativity is to adopt quite a different wave equation and that, when this is done, the new equation leads automatically to effects equivalent to those deduced from electron spin. It need not be *assumed* that the electron is spinning or turning on its axis. This theory of Dirac's, although not free from objection, is the most satisfactory one that we have today. According to it, the electron does behave *as if* it had an internal angular momentum of the sort just described and an associated magnetic moment. Furthermore, if a wave packet is formed representing an electron, there is in it, in general, something like a closed current or eddy of probability that can be regarded as analogous to an actual motion of spin.

In the Dirac theory there are four wave functions for an electron instead of one. The complete Dirac theory is seldom used, however, in spectroscopic calculations. A simplified form usually suffices in

¹ UHLENBECK and GOUDSMIT, *Naturwiss.*, vol. 13, p. 593 (1925); *Nature*, vol. 117, p. 264 (1926).

DIRAC, *Roy. Soc. Proc.*, vol. 117, p. 610; vol. 118, p. 351 (1928). Cf. also DARWIN, *Roy. Soc. Proc.*, vol. 118, p. 654 (1928).

which only two functions are used, one for each direction of the spin moment. The spatial probability density is $\psi_1^2 + \psi_2^2$, and ψ_1 and ψ_2 stand for the two functions.

For a first approximation, furthermore, it is frequently convenient to ignore the effect of spin upon the energy. Then, for the hydrogen atom, both wave functions have the same form as functions of the space coordinates, and it is convenient to take account of spin merely by adding two more quantum numbers, s and u . The total spin quantum number s is analogous to l but has always the value $\frac{1}{2}$; the total magnitude of the spin angular momentum is $\sqrt{s(s+1)} \hbar/2\pi$ or $\hbar \sqrt{3}/4\pi$. The axial quantum number u has the value $\frac{1}{2}$ for one of the two wave functions and $-\frac{1}{2}$ for the other. Since s never changes, it need not be indicated; a sufficient notation for one of the wave functions is then $\psi_{nl\lambda\mu}$. The total angular momentum about the axis of polar coordinates associated with such a function is $(\lambda + \mu)\hbar/2\pi$.

When the effects of the spin magnetic moment are allowed for, it is found necessary to modify the wave functions, by making suitable linear combinations of the functions $\psi_{nl\lambda\mu}$. This feature is discussed further in Sec. 133; but no actual mathematical details of the theory can be given in this book.

CHAPTER VIII

ATOMIC STRUCTURE AND OPTICAL SPECTRA

In the discussion of quantum theory in Chap. VI, we considered only atoms containing a single electron. We shall now attempt to explain the properties of the numerous kinds of atoms (and molecules) that contain 2 or more electrons.

The Rutherford-Bohr model of the atom was built out of raw material which grew in the field of physics. Originally, however, the atom was a child of chemistry. The ultimate theory of its structure must explain and agree with the facts of chemistry, which are just as cogent as are the facts of physics. In part of the present chapter, we shall deal with the explanation of the periodic table of the elements. In another part, we shall return to the study of spectra. The two topics are closely related; for wave mechanics furnishes the key to the theoretical understanding of both subjects. The mathematical theory encounters such great difficulties, however, that empirical evidence from the facts of spectroscopy has to be invoked in order to interpret some of the finer details of atomic structure which are important to the chemist.

Before taking up either of these topics we shall first describe briefly the manner in which the theory has been extended to cover systems of 2 or more electrons.

COMPLEX ATOMS

118. Theory of the Many-electron Atom.—A completely correct wave-mechanical theory of the many-electron atom can scarcely be given, but an approximate theory has been developed that should be adequate to cover all known experimental facts if only the mathematical conclusions to be derived from it could all be worked out. Even this approximate theory is so elaborate that only a few details concerning it can be given in this book.

The approximate theory indicates the existence of *stationary states* or quantum states, for the atom as a whole, in close analogy to those described in Sec. 115 for a single electron. To a large extent the experimental facts can be described in terms of these quantum states without mentioning the mathematical theory at all. The states are commonly labeled with appropriate quantum numbers; and

it happens that they are characterized by certain features as regards angular momentum that serve greatly to assist the memory. As a matter of fact, considerable progress was made toward understanding complex spectra in terms of atomic quantum states and their characterization in terms of angular momenta, even before the advent of wave mechanics. The discussion will be phrased consistently in terms of these atomic quantum states. From time to time a few remarks will be added concerning the mathematics of the wave-mechanical theory, but a reading of such remarks will not be necessary for an understanding of the remainder of the text.

A description of the atomic quantum states is most easily developed by imagining first that *the electrons in the atom do not interact at all with each other*. On this assumption, each electron moves in the field of the nucleus as if the other electrons were absent; and, if the effect of spin upon the energy is also ignored, each electron may be assumed to occupy one of the set of possible quantum states that were described in Secs. 115 and 117. These states were labeled with four quantum numbers n, l, λ, u . Three of these, l, λ , and u , possess physical significance in terms of angular momentum, u representing momentum due to spin; but the *energy* of an electron depends only on the number n . There is also a fifth quantum number s , but this is always equal to $\frac{1}{2}$ and hence can be ignored.

Now in reality this picture is unwarranted because the electrostatic repulsion between the electrons can never be neglected. Since, however, according to classical theory the electrons would be moving about, or, according to wave mechanics, they do not occupy definite positions but have only a certain probability of being found here or there, it would be expected that the collection of electrons as a whole would be more or less equivalent in their electrical effects to a continuously distributed cloud of negative electric charge surrounding the nucleus. Each electron will then be subject to the electrical action of that part of this cloud charge which arises from the remaining electrons.

We are thus led to imagine, as a first step toward an exact theory, that each electron moves, not in the electric field due to the nucleus alone, but in a modified field which includes a part due to the charge cloud representing all the other electrons, and that, aside from their participation in the production of this field, the electrons exert no further effects on each other. The resultant electric field in which each electron moves may be assumed to possess spherical symmetry about the nucleus. Now classical theory leads to the conclusion that an electron moving in such a field will have a fixed angular momentum

about the nucleus. We are thus led to expect that the quantum states for the electrons in the modified field will be characterized by the same properties in terms of angular momentum as are the quantum states in the nuclear field alone, so that they may be labeled with the same set of quantum numbers $n \ l \ \lambda \ \mu$. The same conclusion follows from the detailed wave-mechanical theory [Sec. 120(6)].

The *energy* associated with each electronic quantum state depends again chiefly upon the quantum number n , increasing (algebraically) as n increases; it is not, however, the same function of n as in the 1-electron atom, since the potential-energy term that occurs in Schrodinger's equation [Eq. (182)] is now decreased because the field of the nucleus is partly screened off by the negative charge cloud, so that the potential-energy term V in Schrodinger's equation is diminished. Furthermore, the energy in the modified field is found to depend slightly upon I , increasing as I is increased. The energy of an electronic quantum state is independent of I only in the special case of a Coulomb field such as that due to a point charge.

The possible values of the spin quantum number u , as before, are $\frac{1}{2}$ and $-\frac{1}{2}$; while n , l , and X may be any integers (or zero) such that

$$n > l \geq |X|.$$

Sometimes ml is written for X , and m_l for u . Furthermore, in place of I , the notation developed by the spectroscopists, as described in Sec. 115(e), is often more convenient; in this notation I values of 0, 1, 2, 3, 4, 5, 6, 7, etc., are denoted by the letters s, p, d, f, g, h, j, k , etc. The number of electrons that have the same n and I in an atom is often indicated by a superscript. For example,

$$2s^23d$$

denotes 3 electrons of which 2 are in quantum states with $n = 2$, $I = 0$, while the third is in a state with $n = 3$, $I = 2$. The states of lowest energy for an electron are the $1s$ states ($n = 1$, $I = 0$, $X = 0$, $u = \pm\frac{1}{2}$); the next lowest are the $2s$ states ($n = 2$, $I = 0$, $X = 0$, $M = \pm\frac{1}{2}$) with slightly lower energy than the six $2p$ states ($n = 2$, $I = 1$, $X = -1, 0$ or 1 , $u = \pm\frac{1}{2}$); and so on.

When two or more electrons coexist in the same atom, however, it is also necessary to take account of an important principle known as the *exclusion principle*. In the first place, *no two electrons in an atom can be in the same quantum state, corresponding to the same set of the quantum numbers $n \ I \ u$* . This part of the principle was postulated by Pauli before the advent of wave mechanics (in 1925). In the second place, *it makes no difference which electron is in which quantum*

state; just one atomic state corresponds to an assignment of a set of quantum states, all different, sufficient in number to contain all of the electrons in an atom. The interpretation of the exclusion principle in terms of wave mechanics will be discussed later [Sec. 120 (6)]. All deductions from it have been found to be confirmed by experiment.

Finally, it should be remarked that, as in the hydrogen atom, the set of electronic quantum states is not unique, since the axis that is used in defining them can be drawn in any direction through the nucleus. This freedom as to the direction of the axis is lost only when the atom is subjected to an external field of force.

In dealing with many-electron atoms it is sufficiently accurate to suppose that the nucleus remains at rest.

119. Shells and Subshells.—On the basis of the principles just stated, let us consider how the electrons should normally be distributed among their quantum states in a neutral atom. Let Z electrons be added in succession to a nucleus carrying Z units of positive charge, in such manner as to form a neutral atom in its normal state. Since this is the atomic state of lowest energy, each electron will be assumed to go into the lowest electronic state that is available.

The first two electrons will go into $1s$ quantum states with $n = 1$, $l = 0$, $\lambda = 0$, $\mu = \frac{1}{2}$ or $-\frac{1}{2}$. No other states with $n = 1$ exist; hence the next 2 electrons will go into $2s$ states ($n = 2$, $l = 0$, $\lambda = 0$, $\mu = \pm \frac{1}{2}$), with considerably higher energy; and the next six will go into $2p$ states, with slightly higher energy ($l = 1$, $\lambda = 1, 0$ or 1 and in each case $\mu = \pm \frac{1}{2}$). No other states with $n = 2$ exist. Hence the next 2 electrons must go into $3s$ states (provided Z is large enough), with much higher energy, and so on.

Thus, so far as energy is concerned, the electrons will form distinct groups differing in the value of the quantum number n . Electrons having the same n are said to belong to the same *shell*. Each shell is then regarded as divided into *subshells* according to the values of l . Any subshell may contain up to

$$2(2l + 1)$$

electrons, but no more. For X can range from $X = l$ down through zero to $X = -l$, making $2l + 1$ different values for X ; and, in each case, u may be either $\frac{1}{2}$ or $-\frac{1}{2}$, whence the factor 2. The numbers of electrons in the various types of closed subshells are, therefore, as follows:

s	p	d	f	g
2	6	10	14	18

The maximum possible number of electrons in a whole *shell*, on the

other hand, is

$$2n^2.$$

For, since $0 \leq l < n$, it is

$$\sum_{l=0}^{n-1} 2(2l+1) = 4 \sum_{l=0}^{n-1} l + 2 \sum_{l=0}^{n-1} 1.$$

Here in the series $\sum_{l=0}^{n-1} l = 0 + 1 + 2 + \cdots + (n-1)$, there are n terms, and their average value is $[0 + (n-1)]/2 = (n-1)/2$; hence, the sum of this series is $n(n-1)/2$. Furthermore, $\sum_{l=0}^{n-1} 1 = n$.

Shell.....	K		L		M	
n	1		2		3	
Subshell, l	0	0	1	0	1	2
Letter designation	s	s	p	s	p	d
Number of electrons in subshell or shell. {	2	2	6	2	6	10
	2		8		18	

A more complete list is contained in Table II in Sec. 128.

The spatial volume occupied by the successive electronic shells may be supposed to increase with increasing n . For, in the semi-classical theory, the diameters of the electronic orbits increase rapidly as n increases, and in wave mechanics the probability distribution for the electrons becomes correspondingly more widely extended. This is illustrated for the hydrogen atom in Sees. 94 and 115. A shell that contains as many electrons as can be put into it, viz., $2n^2$ electrons, is said to be completed, or full, or closed. In a neutral atom it can only happen exceptionally that the outermost shell is complete, since the total number of electrons in the atom must equal the atomic number.

-The ideas thus described suffice for the explanation of the periodic table of the elements. Later it will be found necessary to take account

more accurately of the interaction between the various electrons and of the interaction between their spins and their orbital motions. It will then be necessary to modify considerably our conception of the electronic quantum states. In the ultimate analysis, in fact, it becomes impossible to imagine that each electron occupies a quantum state of its own; only the atom as a whole has distinct quantum states. But these changes in the picture will not affect materially the explanation that is presently to be given of the periodic table.

Before proceeding to a discussion of the periodic table, a brief description will be interposed of the mathematical procedure that is used in developing a wave-mechanical theory of atoms containing more than 1 electron, for the benefit of those students who may wish to form some idea of this procedure, but without going into actual mathematical details.

120. Wave Mechanics of Many-electron Atoms, (a) *Fundamentals*.—In Sees. 110 to 115 and 117 we have described the wave-mechanical theory of a single electron in a field of force. When more than 1 electron is present in an atom, there is still a single *probability amplitude* Ψ and a single *wave equation* (at least in the most compact formulation of the approximately correct theory); but both Ψ and the wave equation now contain the coordinates and symbols denoting the spins of *all* the electrons. Since our purpose is not to develop the mathematical theory in detail, we shall not actually write out the wave equation but shall only describe some of its terms.

Only the *stationary or quantum states* of the atom will be discussed. For any one of these states we can write

$$\Psi = e^{-\frac{2\pi i}{h} W t} \psi,$$

in which W is the energy of the atom, t , the time, A , Planck's constant, and ψ , a function which does not contain t (cf. Sec. 113).

The function ψ will be a solution, satisfying certain boundary conditions, of a wave equation not containing the time which is a generalization of Eq. (185) as written for 1 electron. This equation contains a term of the form $V\psi$, but the potential-energy function V now includes not only the potential energy of all the electrons due to the field of the nucleus but also their mutual potential energy due to the repulsion of each one by all of the others. With this preliminary remark, the principal groups of terms in the wave equation (or amplitude equation) for ψ may be listed as follows:

(1) terms containing second derivatives, like those in the equation for 1 electron;

(2) terms due to the nuclear field, which can be written as the sum

$$-e^2 Z \left(\sum_r \frac{1}{r_r} \right) \psi,$$

where e = electronic charge, Z = atomic number, $1/r_r$ = distance of the r th electron from the nucleus, and the sum extends over all electrons in the atom;

(3) terms representing the electrostatic interaction of the electrons, which can be written

where r_{pT} is the distance between electrons number p and r expressed in terms of their coordinates, so that e^2/r_{pT} is the mutual potential energy of this pair, and the sum extends over all pairs that can be formed out of the whole group of electrons;

(4) spin-orbit terms representing an interaction between the spin magnetic moment of each electron and its own orbital motion.

In addition, there are a few terms of minor importance representing mixed spin effects and certain other relativistic effects, but with these we need not concern ourselves.

In the terms of type (3), the coordinates of different electrons occur combined together in such fashion as to make the solution of the wave equation very difficult. For this reason, little progress has been made except by means of a method of approximation known as *perturbation theory*. In using this method, the interaction of the electrons with each other is at first omitted entirely; i.e., each electron is assumed to move in the field of the nucleus alone, just as if the other electrons were not there. Or, still better, the field of the nucleus is replaced by a *modified central field* in which an allowance is made for the average effect of the electrons on one another.

The introduction of the modified field is justified in wave mechanics as follows. In the case of 1 electron, as explained in Sec. I 11, $|\psi|^2 dr$ is interpreted as the probability of finding the electron, and also its total electrical charge, in the element of volume dr , ψ being given the value that it has at some point in dr . This rule can be visualized by imagining the electronic charge e distributed continuously in space with a density $e|\psi|^2$; the fraction of the charge thus assigned to dr is the same as the probability of finding the electron in dr . In a similar way, the charges on all the electrons in a many-electron atom may be imagined spread out in a continuous "charge cloud" whose

density is determined in a certain manner by $|\psi|$. The conception is an artificial one, but it is useful.

The modified central field is now formed by superposing the field of potential due to the electronic charge cloud upon the Coulomb field due to the nucleus. The result is a field the potential of which becomes indistinguishable from the nuclear potential Ze/r (e = numerical electronic charge, Z = atomic number) near the nucleus but falls off more rapidly at greater distances. The electronic charge cloud may be thought of as "screening off" part of the nuclear field. Outside of the atom, the potential in the central field drops rapidly to zero if the atom is electrically neutral, otherwise it passes into the Coulomb field of a charge equal to the net charge on the atom.

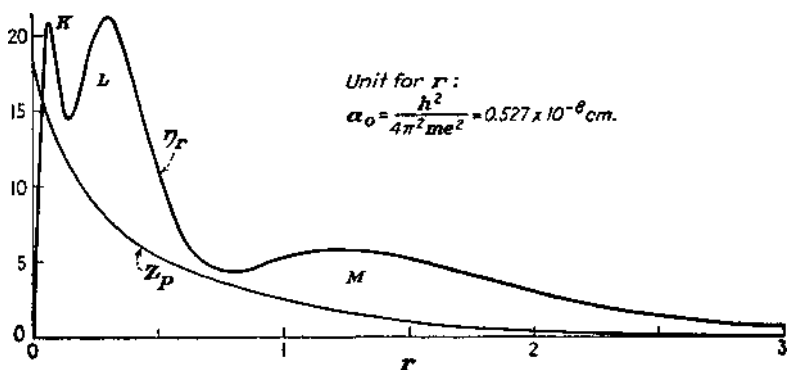


FIG. 85.—The Hartree field and charge-cloud density for argon. The potential energy of an additional electron inside the argon atom would be $-eZ_p/r$ (e = numerical electronic charge, r = distance from the nucleus); the amount of electronic charge between r and $r + dr$ in terms of e as a unit is $n r dr$. K , L , M indicate the approximate locations of the K , L , and M shells.

Strictly speaking, however, the central field formed exactly as described is suitable only for treating the deflection of an external electron by the atom, as in calculating the scattering of cathode rays. In dealing with the motion of one of the electrons in the atom, only the potential due to the charge cloud of the others, or the fraction $(N - 1)/N$ of the potential due to all N electrons in the atom, should be superposed upon the nuclear field. Several good mathematical methods have been invented for obtaining appropriate central fields without a prior knowledge of ψ .¹

In Fig. 85 is shown such a central field for the atom of argon, and also the charge-cloud density, as obtained using the Hartree con-

¹ Cf. CONDON, E. U., and G. H. SHORTLEY, "Theory of Atomic Spectra," 1935.

sistent-field method with allowance for the effects of exchange.¹ The curve labeled Z_p shows, in a certain manner, the potential energy V that an electron would have due to this field at various distances r from the nucleus. Here the reference is to an extraneous electron not belonging to the argon atom. The potential energy itself is not plotted, however, because it varies so rapidly with r , but, rather, the number of units of charge Z_p on an equivalent nucleus; i.e., the ordinate of the curve at any given value of r shows the atomic number of a nucleus which, at that particular distance, would give to an electron a potential energy equal to U , so that $U = -eZ_p/r$. Thus, Z_p varies from $Z_p = Z = 18$ at $r = 0$ to $Z_p = 0$ at the periphery of the atom. The other curve in Fig. 85 is a plot of n_r , $nr\,dr$ being the amount of electronic charge between distances r and $r + dr$ from the nucleus, in terms of e as a unit of charge. The maxima in n_r correspond to the various electronic shells, as is indicated in the figure.

(b) *The Zero-order Approximation: Spin Energy Omitted.*—Having selected a suitable central field, the next step is to ascertain the wave functions corresponding to the various possible *quantum states for a single electron* in this field. For this purpose, it is often convenient at first to omit the spin terms from the wave equation. When this is done, the 1-electron wave functions can be so chosen that each one contains the same sort of angle function as occurs in the wave functions for the 1-electron atom as written above (Sec. 115). The radial factor in the wave function, or the function of the distance r from the nucleus, is different here, however, and can be found only by a process of numerical integration. Furthermore, although the spin terms have been omitted from the wave equation, spin must still be taken into account in drawing up the list of 1-electron quantum states. Each quantum state or wave function for 1 electron is found to be characterized by four quantum numbers, $n\,l\,m\,s$ as already indicated in Sec. 118. The radial factor $R_{nl}(r)$ in the 1-electron wave functions varies both with n and with l ; the energy varies chiefly with n but to some extent also with l . Thus, part of the degeneracy found with a field of Coulomb type was of the nature of a mathematical accident and does not occur with more general types of field.

The first step in obtaining a wave function ψ corresponding to a certain quantum state for the atom as a whole is then to assign a 1-electron wave function for each electron. All of these electronic wave functions must be different, in accord with the exclusion principle. As stated above, this principle can be derived from a certain sort of symmetry that the wave function for the entire atom must

¹ HAKTBEE and HARTKBE, *Ray. Soc., Proc.*, vol. 166, p. 450 (1938).

possess,¹ but no reason can be given (as yet) for requiring this type of symmetry in ψ , except the fact that, if it is required, the results of the theory are thereby brought into agreement with experiment. The restriction thus imposed upon the admissible solutions of the wave equation plays an important role in atomic theory.

A first or "zero-order" approximation to a wave function for the atom as a whole is then constructed out of these 1-electron functions in a manner that need not be described here. This atomic function, or the corresponding atomic quantum state, can be regarded as characterized by the sets of quantum numbers specifying the chosen 1-electron wave functions, *e.g.*, for N electrons, by the following N set of numbers, all sets differing in at least one number:

$$n_1 \ l_1 \ \lambda_1 \ \mu_1, \quad n_2 \ l_2 \ \lambda_2 \ \mu_2, \quad \cdot \cdot \cdot, \quad n_N \ l_N \ \lambda_N \ \mu_N. \quad (A)$$

Even in this first stage of perturbation theory, however, it makes no sense to ask *which* electron is represented by which of the 1-electron wave functions. Here is another point of contrast with classical ideas. The atomic wave function is constructed out of the 1-electron functions in such a way that the coordinates of each electron occur once in each of the N 1-electron functions (just as in the function, $\sin ax \sin by - \sin ay \sin bx$, x and y both occur once in each of the two sine functions characterized by a and b). Thus, according to wave mechanics, the electrons do not seem to possess the same degree of individuality that we attach, say, to separate grains of sand. Sometimes the electrons are imagined to take turns moving in orbits corresponding to the various 1-electron functions; such an assumption is harmless, but there is nothing in the theory actually to suggest such a periodic exchange of positions.

For the same reason, the total charge-cloud density at a given point is not just the sum of the charge densities corresponding to the separate electronic functions, as classical analogies would lead us to expect. Its integrated value, however, must come out equal to the total charge on all the electrons; and, for some purposes, the finer details concerning the charge cloud can be ignored. As a first approximation, therefore, we may, if we wish, imagine the charge on each electron to be distributed in space in the characteristic manner indicated by its 1-electron wave function.

When an approximate wave function has been constructed for the

¹ The atomic wave function ψ must be antisymmetric in the electrons, *i.e.*, $\wedge$ must be such a function that it reverses its sign, without other change, when the coordinates and spins of any 2 electrons are interchanged irf it. For example, $\sin (x_1 - x_2)$ is antisymmetric in x_1 and x_2 , but $\sin x_1 \cos x_2$ or $\sin (x_1 + x_2)$ are not.

atom in this manner, the energy of the atom can be calculated, by a method that will not be described. This energy is not just the sum of the energies of the electrons in the central field, because, when these energies are added up, part of the total represents mutual energy arising from the mutual repulsion of the electrons and is counted twice in making the addition. The simplest physical significance that can be attached to the energy of an individual electronic quantum state is that it represents, except for a change of sign, approximately the work that must be done in removing an electron out of that quantum state in the neutral atom and carrying it off to infinity into a state of rest, thereby ionizing the atom.

(c) *Spatial Degeneracy of the Wave Functions.*—Since the axis that is used in defining the electronic quantum states can be drawn in any direction, the set of atomic wave functions that is obtained in the manner just described is not unique. The functions defined in terms of different axes are not, however, independent of each other. Any function ψ' that is obtained with the use of one axis can be expressed as a linear combination of certain ones of the wave functions obtained with use of any other axis, thus,

$$\psi' = a_1\psi_1 + a_2\psi_2 + a_3\psi_3 + \cdots + a_k\psi_k, \quad (193)$$

where $a_1 a_2 \cdots a_k$ are certain constants depending on the relative orientations of the two axes and $\psi_1, \psi_2, \cdots \psi_k$ are wave functions referred to the second axis. The latter wave functions are all constructed out of electronic functions having the same set of the numbers $n_1 l_2 n_2 l_2, \cdots n_N l_N$ that occur in the electronic functions out of which ψ' is constructed; only the numbers $\lambda_1 \mu_1, \lambda_2 \mu_2, \cdots \lambda_N \mu_N$ are allowed to vary.

Furthermore, it is also possible to make linear combinations among the *electronic* wave functions having the same $n l$ and hence the same electronic energies. In this way a new set of electronic wave functions is obtained, which can be regarded as representing a new set of electronic quantum states. This procedure will be employed in Sec. 133 in order to deal with the spin-orbit effect.

Further discussion of the perturbation theory will be postponed to Sec. 137. At this point we shall return to the discussion of the periodic table.

THE PERIODIC TABLE OF THE ELEMENTS

121. Atomic Numbers of the Elements.—Mention was made, in discussing the Rutherford atom model, of the concept of atomic number. This concept was firmly established by the work of Moseley on

X-rays, to be described later. As a result of Moseley's work, the total number of elements up to and including uranium was fixed at 92, and the atomic number of each element was unambiguously assigned. The *periodic table* of the elements, drawn up by the chemists on the basis of chemical facts primarily, was thus established with the assurance that *no unknown* gaps were left in it. The atomic numbers of the various chemical elements are shown in Appendixes I and II, the latter giving the periodic table in its standard form as used by the chemists.

The atomic number of an atom appears, in Rutherford's theory of the scattering of α particles and in Moseley's work, as the numerical ratio of the charge on the nucleus to the charge on the electron. This ratio is also numerically equal to the number of electrons surrounding the nucleus when the atom is in its normal, neutral condition. When, as a result of the action of some physical agency, an atom becomes *ionized* with the loss of 1 or more electrons, the number of electrons in it is less than its atomic number. The *chemical properties* of an element, however, are understood to refer to the *neutral atom*. Hence the *chemical properties* of an element can be regarded as depending *either* upon its *atomic number* or upon the *number of electrons* in it when it is electrically neutral.

The problem of the periodic table is to be solved, therefore, by determining somehow the *arrangement of the electrons* in the atoms of the various chemical elements and the physical and chemical properties that follow from this arrangement.

122. Some Features of the Periodic Table.—A bird's-eye view of the chief features to be explained is best gained from an arrangement of the periodic table given by Bohr, which is reproduced in Fig. 86. The most striking feature is undoubtedly the repeated recurrence of a noble or *inert* gas, forming a series of turning points in the progression of the elements. These gases are the elements helium ($Z = 2$), neon (10), argon (18), krypton (36), xenon (54), and radon or radium emanation (86). It was pointed out by Rydberg that the values of the atomic number Z for the inert gases can be expressed by simple numerical series:

$$Z = 2(1^2 + 2^2 + 3^2 + 4^2 + 5^2 + \dots);$$

i.e., $2 \times 1^2 = 2$ (helium), $2(1^2 + 2^2) = 10$ (neon); and so on up to 86. Such a fact must obviously find its explanation somehow in the arrangement of the electrons in the atom. The numerical series suggests that the electrons in these inert gases may be arranged in "layers" of some sort; the third inert gas, argon, for instance, would consist of three layers, the first containing 2×1^2 electrons, the second, 2×2^2 ,

and the third, 2×2^2 . The factor 2 suggests symmetry of some kind, or some general reason for the occurrence of electrons in closely associated pairs.

The "inert" gases, as the name indicates, show comparatively little tendency for their atoms to associate themselves with other atoms

Period I
1H
2He

Period II
3Li
4Be
5B
6C
7N
8O
9F
10Ne

Period III
11Na
12Mg
13Al
14Si
15P
16S
17Cl
18Ar

Period IV
19K
20Ca
21Sc
22Ti
23V
24Cr
25Mn
26Fe
27Co
28Ni
29Cu
30Zn
31Ga
32Ge
33As
34Se
35Br
36Kr

Period V
37Rb
38Sr
39Y
40Zr
41Nb
42Mo
43Tc
44Ru
45Rh
46Pd
47Ag
48Cd
49In
50Sn
51Sb
52Te
53I
54Xe

Period VI
55Cs
56Ba
57La
58Ce
59Pr
60Nd
61Pm
62Sm
63Eu
64Gd
65Tb
66Dy
67Ho
68Er
69Tm
70Yb
71Lu
72Hf
73Ta
74W
75Re
76Os
77Ir
78Pt
79Au
80Hg
81Tl
82Pb
83Bi
84Po
85At
86Rn

Period VII
87Fr
88Ra
89Ac
90Th
91Pa
92U
93Np
94Pu
95Am
96Cm
97Bk
98Cf
99Es
100Fm
101Md
102No
103Lr

(1s) (2s) (3s) (4s) (5s) (6s) (7s)

(1p) (2p) (3p) (4p) (5p) (6p) (7p)

FIG. 86.—Bohr's periodic table of the elements. Completed subshells contain two *s*, six *p*, ten *d*, or fourteen *f* electrons. Irregularities in the progression occur at Cr, Cb, Mo, Ma, Ru, Rh (see Appendix III).

in chemical combination. The elements standing in the table on either side of an inert gas, on the other hand, are strongly active chemically, **and** have contrasting properties. Those closely *following* an inert gas, like lithium, beryllium, potassium, calcium, and so on, are metallic and strongly electropositive; they readily form *positive* electrolytic **ions**. **Furthermore, their maximum** valence in chemical com-

pounds is equal to the number of steps by which they lie beyond the inert gas (*e.g.*, 1 for sodium, 2 for magnesium, 3 for aluminum; sodium forms univalent ions in solution, magnesium, bivalent, aluminum, trivalent). The elements closely *preceding* an inert gas, on the contrary, are electrically nonconducting, perhaps even gaseous, and strongly electronegative; they tend to form *negative* electrolytic ions, alone or in combination with other atoms, and they tend to exhibit chemical valence equal to the number of steps by which they precede the inert gas in the table. An element of either of these two kinds scarcely combines at all with another element of the same kind; but an element closely preceding an inert gas in the table combines readily with any other element closely following an inert gas.

These facts invite the conclusion that there is something very peculiar about the arrangement of the electrons in an atom of an inert gas. If we make the reasonable assumption that the chemical activity of an atom is conditioned somehow upon the magnitude of its external electric (or perhaps magnetic) field, we may conclude that the atoms of the inert gases must be surrounded by very weak fields. If so, there should be little tendency for the atoms of these gases to combine into molecules or to condense into liquid or solid form; and it is a fact that the inert gases are composed of monatomic molecules and also have very low boiling and freezing points.

The properties of the elements adjacent to the inert gases are then accounted for if we suppose, further, that the arrangement of the outer electrons in an atom of an inert gas is an especially *stable* one, *i.e.*, an arrangement of especially low energy. An atom of an element following an inert gas in the table will then contain 1 or 2 extra electrons outside of an inert-gas core (*e.g.*, sodium has 1 and magnesium 2 electrons outside of a neon core); and these extra electrons may well be comparatively easy to detach. The tendency of such atoms to form positive ions would thus be explained. Furthermore, in the solid state these extra "valence" electrons may easily come loose under the attraction of neighboring atoms, functioning, therefore, as free electrons, so that the elements in question ought to be good conductors of electricity, which they are.

An element such as chlorine, on the other hand, could arrange its electrons as they are arranged in argon if it had 1 more electron. Such elements, closely preceding an inert gas in the periodic table, might, therefore, be expected to exhibit a tendency to pick up an extra electron, thereby forming a negative ion. Some of these elements do, in fact, form negative ions that are more stable, *i.e.*, have lower energy, than the neutral atom; this is true of the halogens and of

oxygen and sulfur. In the solid state, such atoms would probably contain no electrons with a tendency to become free; thus, the absence of electrical conductivity in these elements would be accounted for.

These ideas furnish a ready explanation for the formation of a compound such as sodium chloride. In combining, the sodium atom loses 1 electron, its remaining 10 electrons then forming the stable configuration that is characteristic of neon (but, doubtless, somewhat more compressed because the nuclear charge of sodium is higher by 1 unit than that of neon). The chlorine atom adds the electron lost by the sodium to its own 17, making 18 electrons arranged in the stable argon configuration (but slightly expanded because of the weaker nuclear charge of chlorine). The electrostatic attraction of the 2 ions thus formed then binds them tightly together into a molecule. When the molecule of sodium chloride thus formed is put into water, the attraction of the ions is weakened and the molecules fall apart into the constituent ions, each with its outer group of electrons in the arrangement characteristic of an inert gas. In a crystal of sodium chloride, also, the grouping into molecules disappears; a crystal of this type is composed of ions but not of molecules. Thus many chemical and physical facts can be correlated if we make the assumption that the arrangement of the electrons occurring in an inert gas is a peculiarly stable one of low energy.

One of the principal problems of an explanatory theory of the atom is thus to account for the high stability of certain configurations of the electrons. Then an explanation must be found for the occurrence of the long sequences of elements that are observed to be inserted *between* the groups of highly active elements surrounding the inert gases. These intermediate elements do not differ strikingly from each other in chemical properties. Many of the "rare earths" are even difficult to separate chemically. The theory should be such that it leads automatically to all of these relations between chemical properties and atomic number on the basis of as few assumptions as possible.

123. The Static Atom.—As was remarked in Sec. 102, while Bohr was endeavoring to develop a theory of atoms on the basis of the older quantum theory, G. N. Lewis and Langmuir suggested a *static* theory of the atom which had some success in correlating chemical and physical facts. It was natural that the physicist should picture an atom as full of *moving* mechanisms, for his information has been acquired from watching atoms *in action* or, at least, from watching the results of their activity, such as the emission of radiation, or the scattering of *a* particles. The chemist, on the other hand, is more concerned with

the atom at rest. He weighs it in finding what combinations it makep with other atoms; he observes it in crystals apparently at rest except for possible thermal agitation; he thinks of it as occupying a definite position as one of the constituent atoms of various complex molecules, being held in place by certain interatomic forces, perhaps of electrostatic origin. It was difficult to see how a thing so full of whirling mechanisms as the physicist pictured the atom to be could keep as quiet as the chemist found it when he looked at it.

Now there is in the periodic table a remarkable tendency, not yet mentioned, for chemical properties to recur in *groups of eight*. Each inert gas except helium, in fact, forms the terminus of such an octet of elements. The first two octets, lithium to neon and sodium to argon (*cf.* Fig. 86), follow each other immediately; the other three are separated by intermediate sequences of elements. It is reasonable to suppose that each of these groups of eight represents the building up of a particularly stable octet or group of 8 electrons in the outermost shell of the atom. Since we know of no reason why the 8 electrons should not be symmetrically disposed about the nucleus, it was assumed by Lewis and by Langmuir that they were located at the corners of a cube.

The static theory of the atom never got beyond the qualitative stage, however, and it was soon abandoned.

124. Wave Mechanics and the Periodic Table: the First Two Periods.—The key to the periodic table was finally furnished by wave mechanics, with the help of two new principles that we have seen to be associated with it: the Pauli exclusion principle and the principle of electron spin. These two additions to the theory were not made until 1925. They have been given due weight in the theoretical discussion in preceding sections. It will now be shown, taking the elements in the order of their atomic numbers, that the ideas and assumptions developed in Secs. 118 to 120 enable us to understand the sequence of the elements. The *neutral atom* will always be assumed to be under discussion unless the contrary is stated.

(a) $Z = 1$. *Hydrogen*.—A single electron in an atom will exist normally in a 1s electronic quantum state, with $n = 1$, $l = 0$, $m = 0$, $s = \pm \frac{1}{2}$; for the energy associated with this state is less than that for any state having $n > 1$. Degeneracy exists because of the two possible values of s . This doubling of the 1-electron quantum states due to spin was not considered in our previous treatment of hydrogen (Sec. 115). It is of importance, however, when a magnetic field is present. (Actually, the state of the electron will be one characterized by other quantum numbers than n and l , as explained in Sec. 133,

but the number of states and their approximate relations as to energy are the same as are obtained with use of the more convenient numbers X and u .)

(6) $Z = 2$. *Helium*.—If 2 electrons are combined with a nucleus for which $Z = 2$, they can both go into $1s$ states, with quantum numbers

$$1\ 0\ 0\ \frac{1}{2}, \quad 1\ 0\ 0\ -\frac{1}{2}.$$

It can be shown from wave mechanics that the electronic charge cloud in helium is spherically symmetrical about the nucleus. It follows easily, by Gauss' theorem, that there should be no electric field at all outside the atom. Atoms of helium ought, therefore, to exhibit comparatively little tendency to associate themselves into a liquid or solid phase, in agreement with observation. The *chemical* inertness of helium, on the other hand, results from other causes (see Sec. 125).

In helium the K shell is complete. The electronic formula of helium is $1s^2$. All heavier atoms will be expected to contain, next to the nucleus, a complete K shell of this sort. Any additional electrons must then go into outer shells; for there are no other electronic states with $n = 1$.

(c) $Z = 3$. *Lithium*.—The third electron in lithium, outside of the heliumlike core, must occupy a state with $n = 2$. An approximate calculation shows that the electronic energy associated with a $2s$ wave function is somewhat less than that associated with a $2p$ function for an electron in the atomic central field; hence, we should expect the third electron to be in a $2s$ state when the atom as a whole is in its normal state. Spectroscopic evidence confirming this conclusion will be described later. The electronic constitution of normal lithium should be, therefore,

$$1s^2 2s,$$

$1s^2$ denoting 2 electrons in $1s$ states. (Since s means $l = 0$, we have $X = 0$, but $M = \pm \frac{1}{2}$, so again there is a twofold degeneracy, or a "statistical weight" of 2, in the normal state of lithium.)

The energy for a $2s$ wave function is found upon calculation to lie much higher than that for a $1s$ function. If the field were a Coulomb field, as for a 1-electron atom, we could use Eq. (187) in Sec. 115, which shows that in such a field the $2s$ state ($n = 2$) lies only a quarter as far below the ionization level as does the $1s$ state ($n = 1$). Hence, the $2s$ electron should be comparatively easy to remove, certainly much easier than either electron in helium. Lithium ought, therefore, to form positive ions easily, and it ought to conduct electricity when in the solid state, the $2s$ electrons of the atoms easily coming loose and

functioning as "free" electrons. Both conclusions are in agreement with fact. Since the ion formed by removal of the 2s electron is univalent, lithium might be expected to combine chemically with a valence of 1. Such is the fact, as shown by such compounds as Li_2O , LiOH , LiCl .

To remove also a 1s electron should require much more energy than to remove the 2s or "valence" electron. In harmony with this conclusion, the first ionization potential of lithium is observed to be only 5.38 volts, as against 75.5 volts for the second. Furthermore, since there is little symmetry to be expected from a single electron in the second shell, the lithium atoms should be surrounded by stray electric fields and should readily group themselves into a condensed phase. The melting point of lithium is 186°C .

The *arc* spectrum from lithium is predominantly that to be expected from a single electron, as described below (Sees. 131, 132 and 134). Evidently the emission of radiation is chiefly due to the outer or valence electron. The so-called *spark* spectrum, on the other hand, which is ascribed to emission by singly ionized atoms of lithium, contains singlets and triplets of lines like the *arc* spectrum of *neutral helium*. This is what we should expect on the basis of the theory, for a singly ionized lithium atom contains the same number of electrons as neutral helium. The frequencies of corresponding lines are much higher than in helium, however, because the stronger nuclear charge causes all energy levels to lie much lower.

(d) $Z = 4$. *Beryllium*.—Two electrons outside of the *K* shell can both be in 2s states but with opposite spins ($u = \pm \frac{1}{2}$)—The resulting element should be a metal rather like lithium but bivalent, since both 2s electrons should come off relatively easily. Such is the case. The oxide, hydroxide, and chloride have the respective formulas BeO , Be(OH)_2 , BeCl_2 .

The second electron is harder to remove than the first, however, since the second helps push the first off by repelling it; the first two ionization potentials of beryllium are 9.3 and 18.1 volts. To remove a third electron, out of the 1s shell, requires 153 volts, and to remove the last one, 217 volts.

Again the spectral evidence confirms the theory. The *arc* spectrum of beryllium is a 2-electron spectrum of singlets and triplets like that of neutral helium. In this case both valence electrons play a role in the emission. In the *spark* spectrum, on the other hand, doublet lines like those from neutral lithium are found; these are emitted by singly ionized beryllium atoms. A few singlet lines are also known in the spark spectrum; it is presumed that they are part of a 2-electron

spectrum, emitted by doubly ionized beryllium atoms. Two lines have even been found which are ascribed to triply ionized atoms.

So far it is obvious that the theory succeeds admirably. But perhaps the student will have wondered why beryllium is not, like helium, a gas with a low boiling point. For the 2s wave functions are just as symmetrical as the 1s functions. As a matter of fact, even the single 2s function for the valence electron in lithium is symmetrical about the nucleus! Only a detailed study of the mathematics of the perturbation theory can really throw light on this question. A satisfactory theoretical answer to such questions can be given, but it is hard to describe in concrete terms.

(e) $Z = 5$. *Boron*.—With 5 electrons in the atom, one of them must go into a 2p state. Boron is a trivalent element, as witness B_2O_3 , $B(OH)_3$, BCl_3 . It is not metallic, however; crystals of boron are good insulators. Evidently, with 3 electrons present in the L shell, conditions are not favorable to the formation of free electrons in the solid state. No simple reason for this is apparent, however.

To remove the 2p electron from a boron atom requires only 8.3 volts; to remove the two 2s electrons in succession requires 25 and 38 volts, whereas to remove one of the 1s electrons as well requires an additional 258 volts.

The next elements in order are most easily understood if we pass them by for the moment and consider neon next.

(f) $Z = 10$. *Neon*.—It is possible to put into an atom two 2s electrons (with $n = \pm\frac{1}{2}$) and six 2p electrons ($l = 1$, $\therefore X = 1$ or 0 or -1 , and with each value of X , $u = \pm\frac{1}{2}$), or 8 in all, but no more electrons with $n = 2$. The L shell is then filled. With every possible value of X and u represented, it can be shown that the electronic charge cloud is symmetrical about the nucleus, as it is in helium; and now the symmetry is sufficiently complete so that a gas having a low boiling point results. Neutral neon in its normal state has the electronic constitution $1s^2 2s^2 2p^6$.

A possible origin thus seems to emerge for all of the octets or groups of eight elements, pointed out above. The first octet, from lithium to neon, results from the progressive filling of the 2s and 2p subshells which constitute the L shell. Any outer shell, if filled, will likewise contain an s and a p subshell, with 8 electrons in the two combined. May it be that each octet of elements, ending in a noble gas, represents the filling up of a new pair of s and p subshells?

Laying this question aside for a moment, let us proceed backward from neon to the remaining elements in the first octet.

(g) $Z = 9$. *Fluorine*: $1s^2 2s^2 2p^5$ — li $Z = 9$, the neutral atom

contains only 7 electrons in the L shell, or 1 less than enough to fill it. If 1 more electron were added, we should have a negative ion the exterior of which would be a closed shell, as in neutral neon. It cannot be expected that the same loss of energy would occur when an electron is added to a *neutral* atom as when it is added to a *positive ion*; but, on the other hand, the electron is added in the shell for $n = 2$ in an atom having a fairly strong nuclear field ($Z = 9$). We can thus understand the fact that fluorine forms univalent negative ions which are stable, *i.e.*, have lower energy than the neutral atom, and that it exhibits a negative valence of 1 in chemical combination. There should be little tendency for a *second* extra electron to be bound by the fluorine atom, for it would have to occupy a state with $n = 3$, the energy of which would lie considerably above that for $n = 2$.

Similarly, *oxygen* ($Z = 8$) commonly exhibits a negative valence of 2 in chemical combination, as in lithium oxide, Li_2O ; and *nitrogen* ($Z = 7$) is commonly *trivalent*, as in lithium nitride, Li_3N . Nitrides are not actually common, however; more often nitrogen is united with oxygen into a compound negative radical, as in lithium nitrate, LiNO_3 .

The examples of Li_2O and Li_3N illustrate the tendency of elements immediately following a rare gas in the order of atomic number to combine with those just preceding some rare gas. This can be interpreted as arising from a lowering of the energy when closed subshells are formed in both atoms. In forming the molecule Li_2O , for example, an electron passes over from each lithium atom to the oxygen atom, leaving the lithium atoms with closed $1s^2$ shells like helium, while the oxygen acquires a closed $2s^2 2p^6$ shell, as in neon. The tendency, suggested above in general terms, toward the formation of the electron configuration that is characteristic of an inert gas, is thus interpreted in wave mechanics as a tendency to form closed subshells.

(*h*) $Z = 6$. *Carbon*: $1s^2 2s^2 2p^2$ —With 4 electrons in the L shell, carbon would seem to have an equal chance either to lose electrons or to add more in an effort to form the closed group of 8. Actually, it probably never does either! The four equal valences of carbon are famous, but its compounds are not formed by a transfer of electrons. Carbon is a typical transition element. In some forms (amorphous carbon, graphite), it exhibits fair metallic conductivity; but in diamond it is an excellent insulator.

At this point, we may with advantage interrupt the discussion of the elements in order to describe the wave-mechanical theory of valence. This forms part of the general theory of the force-action between neutral atoms.

126. Valence Bonds.—Chemists have long regarded the binding of atoms into molecules as being of two distinct types. In some compounds, called "homopolar," the various atoms seem to play similar roles; examples are the compounds in which carbon atoms are joined to each other, and also the diatomic molecules of certain elements, such as N_2 or Cl_2 . Other compounds, called "heteropolar" or simply "polar," are distinguished by the fact that the atoms easily fall apart into charged atoms; examples are salts like $NaCl$. It was suspected even before the advent of wave mechanics that the nonpolar bonds might arise from some sort of cooperative action between a superficial electron in 1 atom and a similar electron in the other, whereas in a polar compound it was assumed that an electron became transferred from 1 atom to the other.

This interpretation of the two kinds of bonds was later derived, with characteristic modifications, from the wave-mechanical theory. In order to explain how the binding comes about, consider the interaction of 2 neutral atoms as they approach each other from a distance. To a first approximation, the electrons in each atom may be treated as being equivalent to a spherically symmetrical charge cloud surrounding the nucleus. In this approximation neither atom is found to be surrounded by an electric field, since both are electrically neutral; hence there is no force-action between them. The picture of the electrons as revolving in orbits, on the other hand, suggests that, although on the average the field due to the electrons might just cancel the nuclear field, there should also be a rapidly fluctuating residual field. This field would polarize the other atom, by repelling its electrons, and would consequently exert a fluctuating attractive effect upon it.

The analog of this classical effect is encountered when the wave-mechanical perturbation theory is developed further. It is found that atoms should exert an attractive force upon each other. This attraction is responsible for the phenomenon of *cohesion*, but it has not a great deal to do with the binding of atoms into chemical compounds.

Chemical forces come into play only when 2 atoms approach so closely that their *electronic charge clouds begin to overlap*. New effects then occur. The most important one is the *electron-exchange effect*. *This effect has no analog in classical physics*. In the mathematical theory it is connected with what happens to the wave function when the coordinates of 2 electrons are interchanged in it; no further description can be given in this book. The physical consequences of the exchange effect are of opposite character according as the 2 electrons involved have the same or opposite spins.

Where charge clouds due to 2 electrons of the *same spin* overlap,

the exchange effect results in a *repulsion* between the atoms. For this effect tends to thin out the charge density, piling it up outside the region of overlap. (The exclusion principle itself, described in Sec. 118, can be regarded as an extreme case of this effect, 1 electron crowding the other out of the atom completely.) The region in which the overlap of the charge clouds first occurs as the atoms approach each other will be the region lying between the approaching atoms and hence will be relatively close to both nuclei; this region is, therefore, a region of especially low energy for an electron, owing to its attraction by the nuclei. Hence, the exchange effect, by shifting some of the electronic charge cloud away from a region of low energy to a region of higher energy, tends to raise the mutual energy of the atoms. This change in their energy causes the atoms to tend to fly apart, since an isolated system tends to move so as to diminish its energy.

If, however, the overlapping charge clouds belong to electrons of *opposite spin*, the exchange effect draws the charge density into the region of overlap at the expense of other regions. To explain how this occurs, we should have to go too far into the mathematical theory. Cloud charge is thereby drawn into a region of comparatively low energy, so that the mutual energy of the 2 atoms is lowered. The atoms tend, therefore, to come closer together, as if they *attracted* each other.

When all the electrons in an atom are considered, the sign of the net exchange effect depends on the relative arrangement of the electrons in the 2 atoms. As between an electron and a closed subshell, which contains equal numbers of electrons of both spins, the net exchange effect turns out to be equivalent to repulsion. This is the cause of the observed general *impenetrability* of atoms and molecules; this impenetrability is due chiefly to repulsion between complete subshells as wholes.

If, on the other hand, each atom has an electron sticking out on its surface, so to speak, beyond any closed subshells that may also be present, this pair of valence electrons, one in each atom, are able to adjust their spins so that they are opposite. The exchange effect between these two electrons can then draw the atoms together into a molecule. Atoms thus combined by means of a pair of electrons acting by virtue of exchange are said to be held together in the molecule by a *covalent bond*; and binding of this type is believed to account for the compounds called *homopolar* by the chemist.

The amount of overlapping of the charge clouds varies greatly according to circumstances. During a collision caused by thermal agitation, a very slight overlapping suffices to develop sufficient repul-

sion to cause the molecules to rebound from each other. Hence in such collisions the centers of the molecules remain relatively far apart. In the case of argon, for example, the "radius" of the atom as calculated by means of the kinetic theory of gases, which represents about half the distance between 2 atoms in an ordinary collision, is at the value 3.45 of the abscissa in Fig. 85, and is thus entirely off the plot. Only a minute part of the charge cloud lies as far as this from the nucleus. In chemical combination, on the other hand, the overlapping of the two atomic charge clouds may be considerable.

Exactly as many covalent bonds can be formed for a given atom as there are extra electrons in it outside of closed subshells, so long as these electrons do not more than half fill the next possible subshell. The atom will thus exhibit chemical valence equal to the number of these extra electrons.

If, however, the extra electrons more than half fill a new subshell, the number available to form covalent bonds tends not to exceed the number of electrons that stand alone in the subshell unmatched by other electrons having the same X but opposite spin. This number is also equal to the number of electrons that would be required to complete the subshell, or to the number of "holes" in it. Thus, when a subshell is more than half full, we can imagine the covalent bonds to be formed by the *holes* in connection with superficial electrons, or perhaps with similar holes, in other atoms. Electrons in an incomplete outer subshell are called *valence* electrons. Perhaps this term might also be extended to the holes in those subshells that are more than half full.

The presence of holes leads also to the existence of a second distinct type of chemical bonding. If the charge cloud of an extra electron lying outside of a closed subshell on one atom comes opposite a hole in an incomplete subshell in another atom, the charge cloud may be drawn into the hole, so to speak, thereby getting closer to the nucleus of the second atom than it can to its own nucleus and therefore moving into a location of lower energy. In this way 2 ions are formed, one charged positively and the other negatively. The 2 ions may then be drawn together into a molecule by electrostatic attraction. In this way are explained the *heteropolar* compounds of the chemist, and particularly the *ionic compounds* between strongly electropositive and strongly electronegative elements, examples of which have been mentioned above.

Both modes of formation result in the same number of valence bonds per atom. Furthermore, in reality all grades of transition occur between the two types; or, rather, the mathematical treatment is

best described by saying that all actual chemical bonds are partly of one type and partly of the other. Bonds between strongly electro-positive and strongly electronegative elements (i.e., between those closely following or closely preceding an inert gas) are predominantly ionic (except in the case of hydrogen!). In general, however, most chemical bonds are predominantly covalent. Molecules such as O_2 and N_2 are also held together by predominantly covalent bonds.

A special case is presented by those atoms whose electrons are all included within completed shells or subshells. Such atoms should exhibit no tendency whatever to enter into chemical combination, and when in the gaseous state they should have monatomic molecules. In this way is explained the occurrence of the noble gases such as helium and neon.

The subject of molecular combination and the related topic of the formation of crystals are complicated subjects and the student is referred for further information in regard to them to other books.¹

126. The Third Period of the Periodic Table.—The second octet of elements, from $Z = 11$ to $Z = 18$, parallels so closely the first octet that brief comments will suffice. (Corresponding elements are exhibited clearly in Fig. 86.) *Sodium* ($Z = 11$, $1s^2 2s^2 2p^6 3s$) contains a single valence electron outside the neon core, but this electron has a $3s$ wave function, as against the $2s$ function for the valence electron of lithium. Sodium is univalent, as in $NaOH$ and the familiar $NaCl$. Its arc spectrum is a typical 1-electron spectrum. In general, it resembles lithium closely. *Magnesium* ($Z = 12$) is a bivalent element similar to beryllium. It burns with a brilliant white flame to form the oxide MgO . It has an arc spectrum of singlets and triplets resembling that of helium. *Aluminum* ($Z = 13$) is trivalent, like boron, but it is metallic and an excellent conductor of electricity. The "sesquioxide," Al_2O_3 , occurs in crystalline form as corundum and sapphire and the ruby. *Silicon* ($Z = 14$) is a good deal like carbon. The dioxide, SiO_2 , however, which occurs in quartz, is a substance of extremely high melting point, whereas the analogous compound, CO_2 , is a gas! Only an elaborately refined application of wave mechanics can explain contrasts such as these. *Phosphorus* ($Z = 15$), although chemically much more active, forms compounds analogous to those of nitrogen. In the poisonous and bad-smelling gas, phosphine or PH_3 , phosphorus is trivalent, just as nitrogen is in gaseous ammonia, NH_3 . *Sulfur* ($Z = 16$) corresponds to oxygen. In H_2S , it is bivalent, just as oxygen is in H_2O . *Chlorine* ($Z = 17$)

¹ Cf. PAULING, L., "The Nature of the Chemical Bond," 1940; SLATER, J. C., "Introduction to Chemical Physics," 1939.

is univalent and easily forms negative ions in solution; in general, it resembles fluorine closely but is less active. Finally, in *argon* ($Z = 18$), we reach again an atom composed of completed subshells, with electronic formula $1s^2 2s^2 2p^6 3s^2 3p^6$. Thus, argon has completed *K* and *L* shells and, outside of these, two more completed subshells. The next or *M* shell is not yet full, since there are no $3d$ electrons. The symmetry of the $3p^e$ configuration, however, is evidently enough to make the external field around argon very weak, so that it behaves as an inert gas with a very low boiling point (-186°C).

127. The Fourth and Fifth Periods.—At $Z = 19$, we should naturally expect the addition of an electron in a $3d$ state {i.e., $n = 3$, $l = 2$). For, in the 1-electron atom, we found the energy to increase regularly with increasing values of n ; hence, we might expect that even in our modified field any wave function with $n = 4$ would correspond to a higher electronic energy than a $3d$ function. But the next element in order, potassium, closely resembles sodium and lithium, not only in chemical properties but also in its spectra even to the finest details. There is abundant reason to believe that atomic spectra in the visible and ultraviolet regions are emitted by electrons in the periphery of the atom. These spectra, therefore, furnish valuable information concerning the state of the outermost electrons. Can it be, then, that the valence electron of potassium is not a $3d$ but a $4s$ electron?

The conclusion that it is a $4s$ electron was drawn from the empirical facts before wave mechanics came on the scene. Subsequently this conclusion was confirmed theoretically by several approximate calculations. Wave functions for $l = 0$ and $l = 2$ are very dissimilar. The radial wave functions for electrons in the modified central field resemble qualitatively those for the 1-electron atom, a few of which are given in Sec. 115 and are plotted (partially) in Fig. 82. The $3d$ radial function, like any function for $Z = n - 1$, shows a single maximum. The $3s$ function, on the other hand, is much more drawn out, corresponding to the Bohr-Sommerfeld elliptical orbit of maximum eccentricity; it has 3 (i.e., $n - l$) maxima and, like all the s functions, is not zero at the nucleus. The $4s$ function is similar to the $3s$ function but has 4 maxima. One can well believe that the energies associated with such different functions may be quite different, although it is not easy to see which energy should be higher. Approximate calculations show that in *potassium* ($Z = 19$) the theoretical energy of a $4s$ electron actually is below that of a $3d$ electron; hence, in the normal state of the potassium atom, the last electron should be in a $4s$ state. The formula for potassium is thus $1s^2 2s^2 2p^6 3s^2 3p^6 4s$, and the facts are fully accounted for.

The next element, *calcium* ($Z = 20$), has two $4s$ electrons and corresponds rather well to magnesium in the preceding octet. Thus, we appear here to be witnessing the building up of another octet. The next four elements, *scandium* (21), *titanium* (22), *vanadium* (23), and *chromium* (24) show some degree of resemblance to the corresponding elements in the preceding period (Al, Si, P, S). The next element after these, however, *manganese* (25), is not at all like chlorine [*cf.* $\text{Mn}(\text{OH})_2$!]; and, in the group that follows manganese, consisting of *iron* (26), *cobalt* (27), and *nickel* (28), we encounter something definitely new.

The accepted theoretical explanation of this failure to finish out an octet at this point in the periodic table is the following. As far back as scandium ($Z = 21$), there is theoretical evidence for the belief that the electron added last goes into the $3d$ subshell; otherwise, it would necessarily be in a $4p$ state, the $4s$ subshell being full, whereas calculation shows that the energy of a $4p$ electron should lie above that of a $3d$. Spectroscopic evidence confirms this theoretical conclusion. The next step in atom-building after calcium ($Z = 20$) is, therefore, the addition of ten $3d$ electrons in succession. Only in chromium and copper are there irregularities; in these atoms one of the $4s$ electrons is replaced by an extra $3d$.

Copper ($Z = 29$, $1s^2 2s^2 2p^6 3s^2 3p^6 3d^{10} 4s$) is the first element that contains, in its normal state, a completed $3d$ subshell. It thus has the K , L , and M shells all full, and 1 valence electron outside of these. The last $3d$ electron, however, added out of turn as if in haste to complete the $3d$ subshell, is rather lightly bound and comes off rather easily, so that copper frequently exhibits a valence of 2, arising from binding by both the $4s$ and one $3d$ electron, as an alternative to its expected valence of 1 due to the $4s$ electron (*cf.* cuprous oxide, Cu_2O , and cupric oxide, CuO).

In *zinc* ($Z = 30$, $1s^2 2s^2 2p^6 3s^2 3p^6 3d^{10} 4s^2$), the $4s$ subshell is completed. Thus zinc, like calcium and magnesium, has 2 valence electrons outside of closed subshells. These three elements are much alike both chemically and spectroscopically. In magnesium and zinc, however, the core of the atom consists of entire closed *shells* (K and L in magnesium, K , L , and M in zinc); perhaps this is the reason why the chemical resemblance is, in general, closest between these two of the three elements. However, whereas the oxides MgO and CaO react with water to form the hydroxide, $\text{Mg}(\text{OH})_2$ or $\text{Ca}(\text{OH})_2$, respectively, ZnO occurs naturally in a zinc ore. In this respect, magnesium resembles calcium more closely. The carbonates all occur naturally as minerals, MgCO_3 as magnesite, CaCO_3 as limestone, ZnCO_3 as smithsonite.

Beyond zinc, the formation of an octet sequence is definitely resumed, the 4p subshell being progressively filled. *Gallium, germanium, arsenic, selenium, and bromine* are good chemical analogs of aluminum, silicon, phosphorus, sulfur, and chlorine. Finally, at $Z = 36$, we have the next inert gas, *krypton*, with electronic formula $1s^2 2s^2 2p^6 3s^2 3p^6 3d^{10} 4s^2 4p^6$. Thus for the third time an inert gas occurs just when a new p subshell has been completed, in accord with the rule that we guessed in connection with the study of neon.

The *fifth period* of 18 elements, from $Z = 37$ to $Z = 54$, corresponds so closely to the fourth period that detailed comment is unnecessary. Two 5s, ten 4d, and six 5p electrons are added, the final element in the period being the inert gas xenon. The analogs in the fifth period of iron, cobalt, and nickel in the fourth period are the three elements ruthenium, rhodium, and palladium. Although the latter elements are not ferromagnetic, they exhibit many chemical similarities to the ferromagnetic elements Fe, Co, Ni. The explanation of ferromagnetism on the basis of wave mechanics is intricate; that it occurs at all is rather of the nature of an accident. Silver, the analog of copper in the preceding period, is univalent, not exhibiting the abnormal alternative bivalency.

128. The Periodic Table Concluded.—*Cesium* (55) and *barium* (56) start the sixth period with one and two 6s electrons, respectively, outside of an argon core; they constitute the fifth pair, consisting of an alkali and an alkaline earth, to be formed by the addition of two s electrons in succession (the first four pairs being lithium and beryllium, sodium and magnesium, potassium and calcium, rubidium and strontium).

Lanthanum (57) then follows the precedent set by scandium and yttrium in the preceding periods and adds a 5d electron. But beyond this point something new occurs again. With $n = 4$, 4/ states with $l = 3$ are possible. From *cerium* (58) to *lutecium* (71), the 4f electronic energy probably lies below the 5d; hence, over this range in Z , fourteen 4f electrons are added in succession

$$[14 = 2 \times (2l + 1) = 2 \times (6 + 1)].$$

These electrons lie rather deep in the atom, however; their wave functions have a single maximum, whereas the 6s and 5d functions are considerably drawn out and thus project more on the exterior. Hence the addition of the 4f electrons causes very little change in chemical properties, and we have the group of the chemically similar, and in part almost indistinguishable, *rare earths*.

Only when the 4f subshell has been completed is the filling of the

bd subshell resumed. From *hafnium* (72) to *mercury* (80), the remaining nine *5d* electrons are added, with a few irregularities. In *gold* ($Z = 79$), the *5d* subshell is completed out of turn, in analogy with copper and silver. Gold, however, manages to combine as a trivalent element as well as a univalent one {e.g., AuCl_3 as well as AuCl). The next element, *mercury* ($Z = 80$), is analogous to beryllium, magnesium,

TABLE II.—COMPLETED ELECTRON SHELLS AND SUBSHELLS

n, <i>I</i>	X-ray symbol	Electron symbol for sub-group	Electrons in subgroup [2(2 <i>l</i> +1)]	Element at which subgroup is completed	Electrons in completed <i>n</i> group, $2n^2$	Element at which <i>n</i> group is completed
1, 0	K	1s	2	He(2)	2	He(2)
2, 0	<i>L_I</i>	2s	2	Be(4)	8	Ne(10)
2, 1	-LIIIII	2p	6	Ne(10)		
3, 0	MI	3s	2	Mg(12)	18	Cu(29)
3, 1	MI,III	3p	6	Ar(18)		
3, 2	Mrv.v	3 <i>d</i>	10	Cu(29)		
4, 0	NI	4s	2	Zn(30)	32	, Lu(71)
4, 1	NI,III	4p	6	Kr(36)		
4, 2	Niv.v	4 <i>d</i>	10	Pd(46)		
4, 3	NVI,VII	4 <i>f</i>	14	Lu(71)		
5, 0	OI	5s	2	Cd(48)	50	
5, 1	OII,III	5 <i>p</i>	6	Xe(54)		
5, 2	Oiv,v	5 <i>d</i>	10	Au(79)		
5, 3	Ovi.vn		14			
5, 4	OVIII,IX		18			
6, 0	PI	6s	2	Ba(56)		
6, 1	PII,III	6 <i>p</i>	6	Rn(86)		
6, 2	PIV.V		10			
7, 0	QI	7s	2	Ra(88)		

zinc, and cadmium. Then the *6p* electrons are added, completing an octet of elements (except for the missing 85) which corresponds weakly to previous octets and ends with the last known rare gas, *radon* or radium emanation ($Z = 86$, $1s^2 2s^2 2p^6 3s^2 3p^6 3d^{10} 4s^2 4p^6 4d^{10} 4f^{14} 5s^2 5p^6 5d^{10} 6s^2 6p^6$).

Beyond radon, we find a fragment of a *seventh* period, including radium (88), thorium (90), and uranium (92). All of the known elements in this period, however, are radioactive and therefore unsta-

ble. It would seem that wave mechanics, as applied to the electrons, imposes no obstacle to the occurrence of elements beyond uranium; the electrons would "know how" to group themselves about a nucleus with $Z > 92$, for example, and actually do it on occasion for $Z = 93$ and $Z = 94$ (see Sec. 227).

From this discussion it is evident that wave mechanics, as applied to the electrons in atoms of varying atomic number, leads without the addition of special assumptions to a sequence of chemical properties such as is actually observed.

The probable electronic constitution of all the elements, when in their normal state, is shown for reference in Appendix III. An outline showing the completed shells and subshells is given in Table II. For convenience in future reference, the table includes also the designations of the subshells that are used in the discussion of X-rays.

OPTICAL SPECTRA

In previous chapters we have become acquainted with a few features of atomic spectra. We saw in Chap. VI that a general characteristic of such spectra is the occurrence of series of lines converging toward short-wave-length limits. In Chap. VII, we have seen how wave mechanics was able to supply a quantitative theory of the general features of spectra emitted by 1 electron in the field of a nucleus, yielding the same formulas as did the original quantum theory of Bohr. A brief but more systematic account will now be given of some representative phenomena of the spectra emitted by atoms or molecules containing more than 1 electron, and of their theoretical interpretation by means of wave mechanics. In connection with this discussion, we shall also return briefly to the 1-electron atom in order to discuss the "fine structure" of its lines, which is due to spin and other relativistic effects. In the present chapter, however, the discussion will be limited to spectra whose emission is connected with the motion of the exterior electrons of atoms or molecules or with nuclear motions; these spectra occur in the infrared, visible, or ultraviolet regions. The high-frequency radiation known as X-rays will be reserved for special treatment in a later chapter.

129. Atomic and Molecular Spectra.—The variety of spectral lines that can be emitted by an atom containing several or many electrons is very large, but it is less than might be expected for several reasons.

For one thing, in discussing atomic spectra, the various types can all be covered by discussing the spectra emitted by *neutral* atoms. For it is a universal rule that *the same types of spectra that are emitted by neutral*

atoms of given atomic number Z are emitted also by singly ionized atoms of atomic number $Z + 1$, by doubly ionized atoms of number $Z + 2$, by triply ionized atoms of number $Z + 3$, and so on. The principal difference is that the frequencies of corresponding lines increase with each stage of ionization, owing to the stronger field in which the active electrons move. The physical basis for this rule lies in the fact that the atoms specified all have the same number of electrons, and these electrons arrange themselves in the same types of configurations. Such atoms are called *isoelectronic*. The similarity of their spectra greatly simplifies the *theoretical discussion*; but it also leads to great complications in the *observed* spectrum as emitted by a given element when excited in a high-potential discharge, since in such a case the different spectra emitted by atoms in various stages of ionization are all jumbled together.

The theoretical basis for the treatment of many-electron atoms was sketched in Sees. 118 to 120. As is there explained, the first step is to construct a simplified approximate theory. In this theory, the electrical field due to the nucleus is replaced by another central field which includes the average effect of the fields of the remaining electrons, the latter being supposed for this purpose to be replaced by an equivalent charge cloud. In the first stage of the treatment, each electron is assumed to move in the central field as if the others were not there. When the atom is in a stationary state, therefore, each electron can be supposed to occupy a certain electronic quantum state, represented in wave mechanics by a certain 1-electron wave function. In order to explain the details of atomic spectra it is then necessary to introduce corrections for the finer details of the electrostatic interaction of the electrons with each other and for the interaction between their magnetic moments due to spin and to their orbital motions.

In treating the radiation from atoms, however, it is not necessary to keep track of *all* of the electrons. The wave-mechanical theory of radiation leads to the following simplifying principle: *in so far as it is possible to regard the electrons as occupying 1-electron quantum states, jumps usually occur only between those stationary states of the atom which differ in just one of the electronic quantum states.* Or, stated more briefly, as a rule, only 1 electron jumps at a time.

The complexity of atomic spectra is greatly reduced by this principle. If the atom has 1 or more valence electrons outside of closed subshells, the electrons in the subshells play no direct part in most of the radiation phenomena with which we shall be concerned; usually only the valence electrons are active. Furthermore, only 1 of these valence electrons jumps at a time (as a rule). The valence

electrons that do not jump, if there are such, cannot be ignored entirely, however; as we shall see, their presence has a great deal of influence upon the characteristics of the atomic energy levels between which the jumps occur. If no valence electrons are present, as in an inert gas, then for the emission of spectral lines it is necessary that an electron be first transferred by some process of excitation out of a subshell—from the outermost subshell, in the cases considered here—into a 1-electron quantum state of higher energy.

The complexity of many-electron spectra is further reduced by the existence of various *selection rules*. Wave mechanics not only tells us what frequencies can occur in the spectrum but it also furnishes a certain amount of information concerning the relative intensities of the various spectral lines. For it gives us not only the energies of the various atomic states but also information concerning the relative probability of jumps between them. Sometimes the probability of a given jump is excessively small, or even zero; in that case, the corresponding line is not seen in the spectrum. A rule stating that jumps between a certain pair of levels do not occur or are "forbidden" is called a selection rule.

The principal selection rules are closely associated with the angular momentum of the electrons in the atom; and considerations of angular momenta, which may be treated vectorially, possess great pictorial value as an aid to the memory. For this reason the discussion of atomic spectra will be centered so far as possible about vector-momentum considerations. In doing this we are also following the historical line of approach.

130. Total Angular Momentum and Its Selection Rules.—We should not expect that the angular momentum of any individual electron would be constant, because of the repulsive forces exerted on this electron by the others. There is also an interaction between the spin magnetic moments and the orbital motion (see Sec. 133). Classical analogy does, however, lead us to expect that the *resultant* angular momentum of *all* the electrons about the nucleus, including their angular momentum of spin, will have a constant value; for, in classical mechanics, the vector angular momentum of any isolated system about its center of mass is constant both in magnitude and in direction, and for nearly all purposes the center of mass of a many-electron atom can be supposed to lie in the nucleus itself.

An analogous conclusion follows from wave mechanics, with modifications similar to those that were found to hold in the case of 1 electron [Sec. 115 (6)]. Wave mechanics leads to the following rule for the quantization of the angular momentum of an atom about its

center of mass: *As long as no external forces act on an atom, its quantum states can be so chosen that, when the atom is in one of them, its component of total angular momentum, including that due to spin, about any chosen axis through the nucleus is*

$$M \frac{h}{2\pi},$$

and the magnitude of its total angular momentum is at the same time

$$\sqrt{J(J+1)} \frac{h}{2\pi},$$

where J and M are two quantum numbers characteristic of the quantum state. The number J is a positive integer or half-integer or zero, and M has one of the $2J + 1$ integrally spaced values between $M = -J$ and $M = J$, inclusively. Thus, M is integral or half-integral according as J is. For example, if $J = 0$, $M = 0$; if $J = \frac{1}{2}$, $M = \frac{1}{2} - \frac{1}{2}$; if $J = 1$, $M = 1$ or 0 or -1 ; etc. Often M_j is written for M .

The rule for determining whether J and M are integral or half-integral can be discovered by taking as a basis the electronic quantum states described in Secs. 118 and 120. In those states there is, for each electron, an integral value of the quantum number X , and also a value of u which is either $\frac{1}{2}$ or $-\frac{1}{2}$; and the total angular momentum of the electrons about the chosen axis is $[\Sigma(\lambda + \mu)]h/2\pi$, where the sum 2 is to be taken for all electrons in the atom. Thus,

$$M = \Sigma(\lambda + \mu). \quad (194)$$

From this equality and the character of the values of X and u it is at once obvious that M and hence also J must be integral when the atom contains an even number of electrons, and half-integral when it contains an odd number. The same rule follows from the exact wave-mechanical theory. It can also be seen that the maximum possible value of M , and hence also of J , is $(\Sigma l) + N/2$ where N is the total number of electrons in the atom; for the maximum value of any individual λ is l , and u cannot exceed $\frac{1}{2}$.

The relationship of the quantum numbers J and M can be visualized with the help of a vector diagram, in the same way in which the relation of l and X was visualized for a single electron [Sec. 115 (6)]. Let a vector of length J be drawn at such an angle that its projection on the chosen axis is of length M . If $M = \pm J$, the vector will lie along the axis. Two cases are illustrated in Fig. 87. If $J = 2$, five positions of the l -vector are possible, corresponding to $M = -2, -1,$

0, 1, 2; if $J = \frac{3}{2}$, four positions can occur, corresponding to $M = 0, \pm\frac{1}{2}, \pm\frac{3}{2}$.

The $\vec{L}$ -vector so drawn can be regarded as representing in a certain way the total vector angular momentum of the electrons, whose component in the direction of the axis is $Mh/2\pi$. The magnitude of the total angular momentum, however, is not $Jh/2\pi$ but $\sqrt{J(J+1)} h/2\pi$, according to wave mechanics. Furthermore, as in the case of 1 electron, the vector momentum has no definite azimuth around the axis, as it necessarily has in classical theory. A particular quantum state corresponds to an average of classical motions for all possible azimuths, as was described for 1 electron in Sec. 115 (6).

Selection Rules.—Wave mechanics leads to the following selection

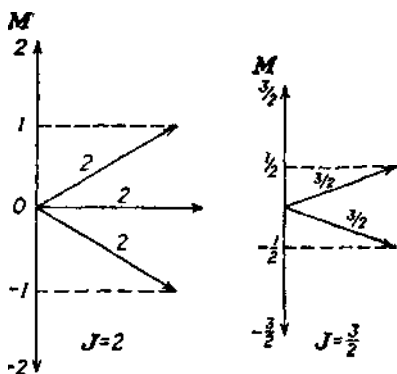


FIG. 87.

rules for J and M , which are of great importance in spectroscopy. In any atomic jump associated with the emission or absorption of a photon, M may remain unchanged or it may increase or decrease by unity, but not by any larger amount. That is, if J_1 and J_2 are the values of J for the two states between which the jump occurs, then either $J_1 = J_2$ or $J_1 = J_2 \pm 1$. The same rule holds also for M .

No jump can occur, moreover, from a state with $J = 0$ to another state with $J = 0$. In symbols:

$$\Delta J = 0 \text{ or } \pm 1 \text{ (and not } 0 \rightarrow 0\text{);} \quad (195)$$

$$\Delta M = 0 \text{ or } \pm 1. \quad (196)$$

These selection rules are not always strictly obeyed, however; sometimes jumps occur for which $\Delta J = \pm 2$ or $\Delta M = \pm 2$. The lines due to such jumps are mostly very weak.

Even these selection rules find an analogy in classical theory (and were first proposed on the basis of this analogy), at least if spin is omitted. Orbital angular momentum, in classical mechanics, is associated with rotation. Suppose that a body capable of rotation contains an electrical vibrator of some sort; and let this vibrator emit electromagnetic radiation of frequency ν when the body is at rest. Then, if the body is set rotating with angular velocity ω , the radiation emitted will contain just the three frequencies ν , $\nu + \omega$, $\nu - \omega$. The reason is similar to that for the analogous effect of a magnetic

field upon a classical vibrating electron, which was described in Sec. 40 in connection with the classical explanation of the Zeeman effect. The three classical frequencies $\nu + w$, ν , $\nu - w$ correspond (in emission) to the three possible changes in M . This is an example of Bohr's correspondence principle.

Angular Momenta of Closed Subshells.—When the approximate theory in terms of electronic quantum states is employed, every electron in a subshell is matched by another whose quantum state differs only in that the sign of X is reversed; and the values of u are similarly paired off. Hence, $\sum X = 0$ and $\sum u = 0$. This fact suggests that, also, $\sum M = 0$. It can be shown, in fact, from the exact wave-mechanical theory that for any closed subshell $M = 0$.

In a similar way it can be shown that the total orbital momentum taken by itself, or the vector sum of the orbital angular momenta of all the electrons, likewise vanishes; and so does the vector sum of their spin momenta. In terms of electronic quantum states, this conclusion results from the fact that $\sum X = 0$ and $\sum S_x = 0$ in a closed subshell.

Thus, in considering the total angular momentum of an atom, all closed subshells can be ignored. An atom of a noble gas, in its normal state, in which the electrons are all grouped into closed subshells, is necessarily in a state with $J = 0$. An atom containing one or more peripheral electrons outside of the inner closed subshells will have angular momenta, orbital, spin, or total, that are determined entirely by the peripheral electrons, with no contributions from the closed subshells. The consideration of atomic angular momenta is thereby greatly simplified.

131. Alkali-type Spectra.—At this point we may to advantage describe the gross features of the simplest type of spectra, which is emitted by atoms containing a single valence electron outside of one or more closed subshells. Among neutral atoms of this sort may be mentioned those of the alkali metals, lithium, sodium, potassium, rubidium, and cesium. Other examples are singly ionized atoms of beryllium, magnesium, calcium, strontium, or barium; doubly ionized atoms of boron or aluminum; and so on.

In an atom of this type only the valence electron should be optically active. The only electronic quantum numbers that vary are, therefore, those of the valence electron as it changes from one to another of its various electronic quantum states, and the atomic energy will be a function of the electronic quantum numbers n and l associated with this electron. The quantum number l , however, serves to specify in a certain way the orbital angular momentum of the electron, and, as we have seen, this is the same as the orbital angular momentum of the

entire atom, since subshells make no contribution on the whole to the momentum. We shall accordingly employ, in place of l , the capital letter L , which is commonly used for the total *orbital* momentum. In any atom the possible values of L are 0 or any positive integer; and levels are commonly labeled with a capital letter according to the following scheme, analogous to that employed for 1 electron:

$$L = 0, \quad 1, \quad 2, \quad 3, \quad 4, \quad 5, \quad 6, \quad 7.$$

$$S \quad P \quad D \quad F \quad G \quad H \quad I \quad K \quad .$$

The letter indicating the value of L may then be preceded, in the case of a single active electron, by a number giving the value of n .

The array of possible levels, arising from changes in the electronic state of the valence electron, is thus as shown (in part) in Fig. 88. The S terms ($L = I - 0$) begin with $n = 1$, the P terms ($L = I - 1$)

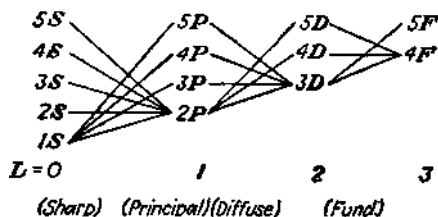


FIG. 88.—Partial LS term system of a typical alkali metal (in actual cases some of the lowest terms shown are missing). Lines indicate allowed transitions.

with $n - 2$, and so on; for always $n > I$. The allowed transitions between these levels are indicated in the figure by oblique lines. The quantum number L is usually subject to the selection rule that ΔL cannot exceed unity; hence, no jumps can occur, for example, between S and D levels. In the case of a single active electron, furthermore, the restriction is actually a little more stringent; for, in any case, it is a special rule that the I of the jumping electron *must* change by unity. Hence, when $L = I$, we must have

$$\Delta L = \pm 1,$$

jumps for which $\Delta L = 0$ being excluded. Thus, jumps cannot ordinarily occur in such spectra between two S levels or between two P levels, and so on.

The scheme just described should obviously be applicable also to the energy levels of the hydrogen atom, which likewise contains a single valence electron. In our previous discussion of hydrogen, we ignored the very slight variation of the energy with l , or L , which is due to spin and relativistic effects. Hence all $S, P, D, \dots$ levels corresponding to a given value of n fell on top of each other, and no

distinction between them needed to be made. We shall return to this point later (Sees. 135 and 136).

The theoretical array of energy levels just inferred from the theory is strongly reminiscent of the spectral terms employed by Rydberg to represent the chief series in the spectra of the alkalis. According to Eqs. (I29a,b,c) in Sec. 89, the wave numbers of the lines of these series are given by the following formulas:

$$\text{Sharp:} \quad \tilde{\nu} = \frac{R}{(1+P)^2} - \frac{R}{(m+S)^2}, \quad m \geq 2;$$

$$\text{Principal:} \quad \tilde{\nu} = \frac{R}{(1+S)^2} - \frac{R}{(m+P)^2}, \quad m \geq 1;$$

$$\text{Diffuse:} \quad \tilde{\nu} = \frac{R}{(1+P)^2} - \frac{R}{(m+D)^2}, \quad m \geq 2;$$

the wave numbers are the differences between expressions or "terms" of just three different types:

$$\frac{R}{(m+S)^2}, \quad \frac{R}{(m+P)^2}, \quad \frac{R}{(m+D)^2}$$

where R is the Rydberg constant (Sec. 88). By varying m in these expressions we obtain three sets of terms similar to the three sets of energy values as derived from wave mechanics. It is reasonable to assume that Rydberg's terms correspond to these latter sets of energy values. The Rydberg terms converge to 0 as $m \rightarrow \infty$, just as the theoretical energy levels converge to zero as $n \rightarrow \infty$, provided we agree to call the energy 0 at the ionization limit. The terms, however, are positive, whereas the energies are all negative. Hence, the terms must represent the *numerical values* of the energy without the negative sign. As has been said before, positive numbers obtained in this way are commonly used by spectroscopists, under the name of "spectral" terms, to represent energy levels. In the formulas for $\tilde{\nu}$, then, the *second* Rydberg term refers to the *initial* atomic state for the transition in question; the second term together with its minus sign equals the energy of the initial state. The *first* term in the formula refers to the *final* state; its (negative) energy is subtracted from the energy of the initial state to obtain $h\nu$ for the transition.

The sequence of terms represented by the first of the three expressions written above, $R/(m+S)^2$, represents, therefore, the sequence of initial states for the various lines of the sharp series. These terms are observed to combine only with terms in the second sequence, $R/(m+P)^2$. Hence, the first of these two sequences of terms must be those derived theoretically with $L = 0$, or the "S" terms; for the

latter is the only theoretical sequence of terms that combines with only one sequence. The second of the Rydberg sequences must be the P terms, with $L = 1$; for S terms combine only with P terms, by the selection rule for L . The *sharp* series of lines thus arises from transitions between an S level (exclusive of the lowest) and the lowest of the P levels. The *principal* series, on the other hand, arises from transitions between various P levels and the lowest S level. The lowest S level is the lowest possible value for the energy and represents the normal state of the atom.

The third set of levels, represented by $R/(m + D)^2$, which combines with P but not with S levels, must, then, be the D levels ($L = 2$). The *diffuse* series of lines thus arises from transitions between the various D levels and the lowest P level. A fourth observed series of lines, named long ago the "fundamental" series, has a convergence limit equal to the difference between the convergence limit of the diffuse series and the frequency of the first line of this latter series, i.e., to the difference in energy between the state of ionization and the lowest D level. The final state for the transitions that give rise to this fourth series must, therefore, be the lowest D level. But the initial energies for the fundamental series are not those of the P terms. They must be, then, the F levels, with $L = 3$. The fundamental series is thus due to transitions from an F level to the lowest D level.

In this way, we arrive at the following interpretation of the chief series of the alkali metals:

Sharp,	$nS \rightarrow n_0P,$	$n > n_0;$
Principal,	$nP \rightarrow n_0S;$	
Diffuse,	$nD \rightarrow n_0P;$	
Fundamental,	$nF \rightarrow n_0D.$	

Here the lines of each series are obtained by assigning various values to n ; n_0 before any letter denotes the lowest value of n that can occur with that letter in the kind of atom under consideration.

The identification thus made of the terms in the alkali spectra constitutes the historical reason for the use by spectroscopists of the mysterious letters S, P, D, F (from the words "sharp, principal, diffuse, fundamental") to represent various values of L (or the corresponding small letters for 1 electron). For higher values of L it was agreed later to continue down the alphabet, skipping J .

As an example, a diagram of the known energy levels for neutral sodium is shown in Fig. 89. Each level is represented in this diagram by a short horizontal line, the levels being grouped into sequences according to the value of L , in contrast to the arrangement in Fig.

62(6). The diagram shows also the splitting of the P levels, which is discussed in Sec. 134 below; the subscripts on the letters, denoting J values, refer to this splitting. As usual, the energies are measured downward from zero at the state of bare ionization (with no kinetic energy in the removed electron) and are expressed in wave-number

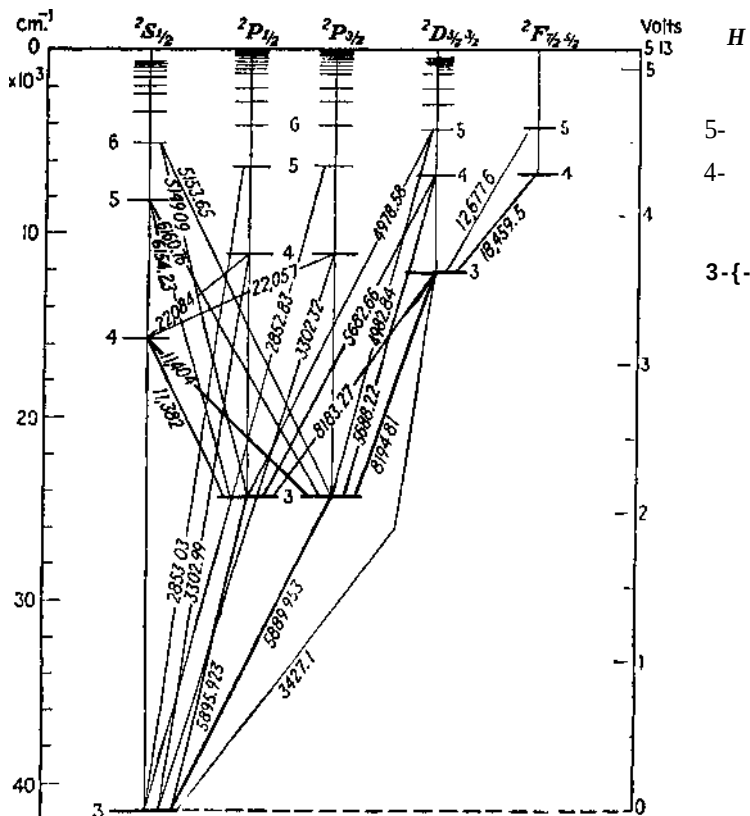


FIG. 89.—Energy-level diagram for neutral sodium, showing also the origin of the principal spectral lines. Some of the levels of hydrogen are shown for comparison. (Wave lengths are in angstroms. The numbers opposite the levels are values of n —see Sec. 132.)

units. The numbers written opposite the levels are the accepted values of n , which differ somewhat, for reasons explained in the next section, from Rydberg's original number m . At the right are shown for comparison some of the levels of hydrogen, the lowest being far below the sodium levels and therefore not shown.

The most important spectral lines are also indicated on the diagram by oblique lines drawn between the appropriate energy levels, with wave lengths attached; the widths of these lines in the diagram give a

very rough idea of the relative intensities of the spectral lines as commonly observed. One line is shown that violates the ordinary selection rule for L ($3D \rightarrow 3S$, $\Delta L = 2$).

The values of some of the terms, expressed in cm.^{-1} , are also given in Table IV (page 337).

132. The Term Energies of the Alkali Metals.—If the mathematical difficulties could be overcome, it should be possible to calculate the energy levels from the wave equation. Unfortunately, however, (or fortunately), this cannot be done with any degree of precision, so that experiment is still necessary. Considerable interest attaches nevertheless to approximate theoretical values of the term energies. Such values can be obtained, for a neutral alkali atom, by disregarding the spin-orbit effect and solving the wave equation for the valence electron moving in a suitable central field.

As described in Secs. 118 and 120, the field in which the valence electron is assumed to move is chosen so as to allow for the average-effect of the remaining electrons in the atom. At points well outside the atom, this field must approximate closely to the field of a proton. For, by ordinary electrostatic reasoning, based upon Gauss' law, it can be shown that a spherically symmetrical distribution of charge produces the same field at external points as if the charge were concentrated at its center; and the total negative charge on the remaining electrons in the atom is less by 1 electronic unit than the positive charge on the nucleus. If the field were everywhere of this simple Coulomb type, the energy levels for the valence electron would be exactly the same as those for hydrogen. Now, in the wave-mechanical theory, the wave functions for large n and not too small l will have their large values chiefly in regions outside the atom, as is illustrated by the graphs for *hydrogen* in Fig. 82 in Sec. 115. Hence, the alkali levels for large n and not too small l should approximate closely to the hydrogen levels for the same n .

In Fig. 90 are shown the known energy levels below $2,000 \text{ cm.}^{-1}$ for all of the alkali metals, represented by dots. The hydrogen levels in the same spectral region (except the lowest) are shown by the horizontal lines labeled with the appropriate value of n . We note that the D and F levels of the alkalis actually agree closely in all cases with the corresponding hydrogen levels.

Such an approximate agreement with the hydrogen levels serves to fix the value of n unambiguously for the D and F levels of the alkalis. The quantum number n does not have any direct spectroscopic significance, hence the particular set of values that is used can only be chosen for some theoretical reason. The P and S levels, on the other

hand, depart widely from the corresponding hydrogen levels. The lowest values of n for S and P terms are taken to be 2 for lithium, 3 for sodium, and so on, partly on the basis of the way in which the periodic table is built up, as was described previously in this chapter, and partly on the basis of various theoretical calculations. When an approximate theory is developed with the electrons moving independently in a central field, the quantum number n is introduced in such a way as to number the 1-electron states in the order of increasing energy. With the values of n assigned as in the diagram, levels having the same n and L have the property that they sink as the atomic

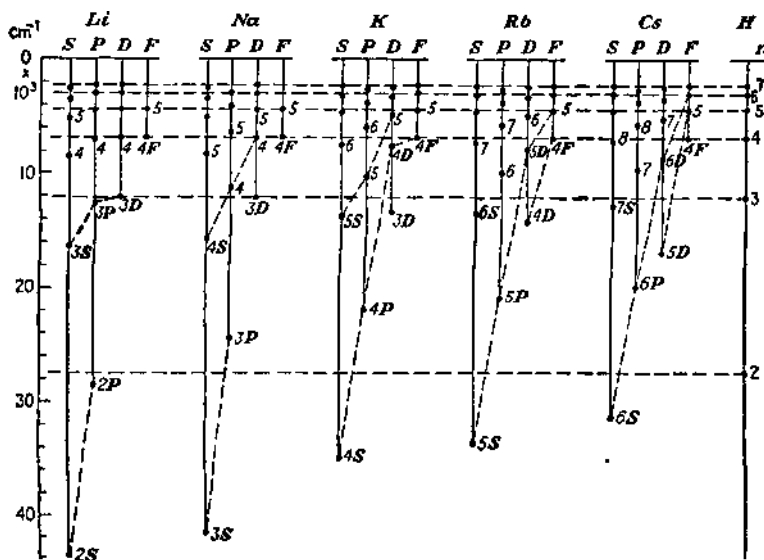


FIG. 90.—The known energy levels below $2,000 \text{ cm.}^{-1}$ of the five alkali metals and of hydrogen (except that for $n = 1$). All doublet structure is ignored.

number increases. Theoretical estimates indicate that they should do this if n has the theoretical significance just stated. In the wave-mechanical theory, 1-electron functions having the same n and l penetrate to about the same relative extent into the cores of all atoms, and, as the strength of the nuclear field in the interior of the atom is increased with increasing Z , calculation shows that the corresponding electronic energy is lowered. The observed energies of the lowest S and P levels of the alkalis, on the other hand, rise with increasing Z . This fact is brought into consistency with the rule just stated if we suppose that n increases by unity from one alkali to the next.

Since the lowest S level is always below the lowest P , the normal state for all of the alkali metals is an S state.

The relation with the hydrogen levels just described suggests a certain mode of writing the alkali energies that is of interest. Expressed in wave-number units, the energy of the n th level for a hydrogen atom is

$$W_n^{(H)} = -\frac{R}{n^2},$$

where R is the Rydberg constant [cf. Eqs. (136), (140c) in Sees. 94 and 95]. For any atomic level with energy W , an *effective principal quantum number* n_e can be defined by the similar equation

$$W = -\frac{R}{n_e^2}. \quad (197)$$

The difference $n - n_e$ for corresponding levels is sometimes called the *quantum defect*. In the alkalis, it can be regarded as a measure of the effect of the departure of the central field from a Coulomb field. The quantum defect is often denoted by u , so that

$$n_e = n - \mu;$$

then

$$W = -\frac{R}{(n - \mu)^2}. \quad (198)$$

A similar formula is obtained from Rydberg's expression for the spectral terms, given in Sec. 89:

$$W = -\frac{R}{(m + \mu)^2}.$$

The values of u are not the same in the two formulas, nor did Rydberg's choice of m for a given term agree with the modern value of n ; but,

TABLE III.—SOME VALUES OF THE QUANTUM DEFECT FOR NEUTRAL SODIUM

Term	$n = 3$	$n = 4$	$n = 5$	$n = 6$	$n = 7$	$n = 8$
<i>S</i>	1.373	1.357	1.352	1.349	1.348	1.351
<i>P</i>	.883	.867	.862	.859	.858	.857
<i>D</i>	.010	.011	.013	.011	.009	.013
<i>F</i>		.000	— .001	— .008	— .012	— .015

after choosing the lowest values of the integers n and of m in any convenient manner, the two values of n can obviously be chosen so as to bring the two formulas into agreement. Since Rydberg's formula is found to fit the observed terms fairly well, when his u is treated as a constant for given L , it follows that the quantum defect must be almost a constant for the sequence of terms belonging to a given value of L .

The values of the quantum defect, $n - n_e$, or u in Eq. (198), for some of the levels of sodium (disregarding the fine-structure splitting) are shown in Table III.¹ Their approximate independence of n is evident. The corresponding value of the effective quantum number n_e may be obtained in each case by subtracting the number shown here from n ; thus for a 4P term, $n_e = 4 - 0.867 = 3.133$. A small value of the quantum defect means close approximation of the energy to the corresponding level for hydrogen.

Formulas of the Rydberg type, as in Eq. (198), represent only a first approximation to the actual term values. For greater accuracy more complicated functions of n are sometimes employed in the

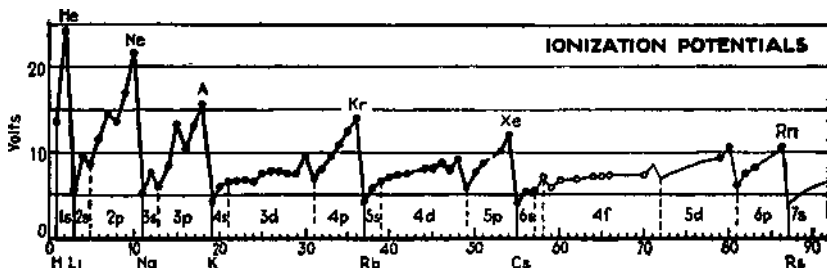


FIG. 91.—Ionization potentials as a function of atomic number. (Reprinted from H. E. White, "Introduction to Atomic Spectra.")

denominator.² Such formulas are indispensable in determining series limits, and a knowledge of these limits is essential in order to fix the zero level of the energy, which is the limit of that series of lines which arises from transitions into the normal state. Once the energies or term values have been determined for a single set of terms, the others can be calculated from the observed frequencies of lines due to transitions between the various terms.

The term value, or negative of the energy, for the lowest state represents the *ionization energy* of the neutral atom. In Fig. 91 are plotted values of the ionization energy calculated in this way and expressed in electron-volts for most of the elements. The electronic subshell that contains the most easily removed electron is shown below the curve. A systematic relation is evident between the quantum numbers of this electron and the ionization potentials.

133. Spin-orbit Effect for an Electron in a Central Field.—Besides accounting for the number and location of the terms, an atomic theory, to be successful, must lead also to the correct fine structure for the

¹ They are calculated as $n - n_e$ from values of n_e or n_{eff} given in H. E. White, "Introduction to Atomic Spectra," 1934.

² Cf. WHITE, "Introduction to Atomic Spectra," 1034; FOWLER, A., "Report on Series in Line Spectra," Fleetway Press, London, 1922.

spectral lines. In order to account for the fine structure it is necessary to consider the effects that arise from the magnetic moment of the electron that is associated with its spin.

The idea of an electron spinning on its axis can only be considered as a picture which suggests more or less correctly the effects of certain terms in the relativistic equations of Dirac, which represent the nearest known approach to the correct wave equations (Sec. 117). The Dirac equations do not rest on a concrete picture of the electron. It seems as if theoretical physics were coming to be based upon certain mathematical assumptions rather than upon concrete pictures of reality. There is, however, a certain similarity between the wave-mechanical mathematics and the mathematical treatment of certain classical structures, and this analogy may help by illustrating, and perhaps making plausible, the more abstract treatment by means of wave mechanics.

Suppose the electron really were, as Lorentz thought it might be, a little hollow sphere of electricity, held together in some unknown manner but otherwise obeying the laws of classical theory, relativity included. Suppose the sphere were set into rotation. Then the moving charges on its surface would give rise to circular currents, and the sphere would be surrounded by a magnetic field resembling that due to a short bar magnet; it would, therefore, have a certain magnetic moment. Because of the coexistence of the electric and magnetic fields in the space surrounding the electron, electromagnetic momentum would also exist there. The resulting vector angular momentum about the center of the sphere is found to be just half as great in proportion to the magnetic moment as in the analogous case of revolution of an electron in an orbit. Because of this fact, it was early surmised that the quantized value of the spin momentum would be half of the ordinary quantum unit and thus of magnitude $\hbar/4\pi$.

Suppose, now, that such a rotating sphere is set in motion across an electric field. Then ordinary electromagnetic theory would not lead us to expect any special forces to act on the sphere because of its magnetic moment. But suppose we view the situation from a moving frame of reference, in which the sphere is momentarily at rest and apply the theory of relativity. In terms of this moving frame, the electric field is accompanied by a magnetic field perpendicular both to the electric field and to the direction of motion, as is shown by the equations of transformation in Sec. 66. If the electric field is that

¹ In those equations, we put $H_x = H_y = H_z = 0$, so that in S there is only an electric field. Then $H'_x = 0$, showing that the magnetic field in S' is perpendicular to the x -axis, which is the direction of the motion. Each of the components E_y and E_z then gives rise to a component of H' equal to itself multiplied by u/c , but in a perpendicular direction.

due to a bare nucleus, the nucleus itself will be moving relatively to the new frame, of reference and may be thought of as generating the magnetic field by the usual rule for moving charges. This magnetic field will cause the magnetic axis of the sphere to tend to line up with the field.

Returning to the original frame, therefore, we may conclude that any rotating charged sphere moving transversely across an electric field tends to turn so as to make its axis of rotation perpendicular both to its direction of motion and to the direction of the field. If the motion occurs in the direction of the field, or if there is no field at all, the axis of rotation retains its original direction. As viewed in this frame, the immediate cause of the change in the direction of the axis is to be found in a relativistic change in the forces that hold the charged sphere together against electrostatic repulsion.

We do not need to investigate these forces further, but the source of the energy associated with them is of interest and is easily discovered. When the sphere is aligned in the manner just described (Fig. 92), one side of it is moving faster than the other side because of the combined rotational and translational motion. Now in a frame in which the sphere is at rest, the rotational velocity of all parts of its surface must be the same, by symmetry; but then an application of the Lorentz transformation for velocity, Eqs. (56a,b,c) in Sec. G2, shows that, in a frame in which the sphere has translational motion, the surface will rotate more slowly on the side on which the resultant velocity is increased by the translational motion than on the opposite side. This difference in velocities must result in a piling up of charge on the side of higher resultant velocity at the expense of charge on the opposite side.¹ If the sphere is moving across an electric field, the distortion shifts the charge into a region of slightly different potential and so alters the electrostatic energy. A simple consideration of signs shows that, in the case of orbital motion of the sphere under the influence of an attracting center, the energy is least when the direction of spin rotation is opposite to the direction of orbital revolution and is a maximum when these two directions of rotation are the same, regardless of the sign of the charge on the sphere.



Denser
dagger i (Lorentz)
spinning electron
J uTof rotation/ *

The magnitude of the change in energy for a Lorentz electron

¹ The distortion of the distributed charge can be regarded as equivalent to the creation of an electric moment relative to the center of charge. Thus what is a pure magnetic moment in one frame of reference becomes in part an electric moment in another.

can readily be calculated, but we shall only cite the result.' in dealing with an electron in a central field, it is convenient to introduce the angular momentum G_i of the electron about the center of the field. Let G_s be its angular momentum of spin, m its mass, c the speed of light, and $V(r)$ its potential energy when at a distance r from the center of the field. Then the change in energy due to interaction of the spin and the field turns out to be

$$\frac{1}{2m^2c^2} \frac{1}{r} \frac{dV}{dr} G_i G_s \cos(G_i, G_s), \quad (199)$$

where G_i and G_s denote magnitudes of these vectors and (G_i, G_s) stands for the angle between their directions.

The change in energy as thus calculated from classical theory gives a good idea of the analogous wave-mechanical effect, although the physical picture of the rotating sphere finds no place in the new theory, and the mathematical formulation is naturally quite different. It would take us too far afield to describe the wave-mechanical mathematics in more detail. We shall merely cite the results and endeavor to make them plausible by means of semiclassical analogies.

The quantum states for an electron moving in a central field under the influence of the spin-orbit forces are found to differ considerably from the $n l \lambda u$ states described in Sec. 118. The new states correspond in a way to the two simplest cases of the classical motion. These are the cases in which the vector angular momentum of spin either has the same direction as that of the vector angular momentum due to the orbital motion, or has exactly the opposite direction. The new states may be characterized as follows. The total orbital momentum has a fixed value, represented in the usual way by the quantum number l , and the total spin momentum has likewise a fixed value, represented by the number s where $s = \frac{1}{2}$; the magnitudes of these momenta are $\sqrt{l(l+1)} h/2\pi$ and $\sqrt{s(s+1)} h/2\pi$, respectively. Furthermore, the resultant momentum, or the vector sum of orbital and spin momenta, is also quantized, and may be represented by a quantum number j . For some states, $j = l + s = l + \frac{1}{2}$, for others,

$$j = l - s = l - \frac{1}{2}$$

except that, if $l = 0, j = l + s = \frac{1}{2}$. The magnitude of the resultant momentum is $\sqrt{j(j+1)} h/2\pi$. Finally, the component of the resultant momentum in the direction of some chosen axis of reference has a definite value, $mh/2\pi$, where m may have any one of the integrally

¹ FRBNKKL, J., "Lehrbuch der Elektrodynamik," vol. II, Kap. 7, §§ 9, 1926.

Thus, if $l = 0$, the only possible values for j and m are $j = \frac{1}{2}$ and $m = \frac{1}{2}$ or $-\frac{1}{2}$. If $l = 1$, either $j = \frac{3}{2}$ and $m = \frac{3}{2}, \frac{1}{2}, -\frac{1}{2},$ or $-\frac{3}{2}$, or else $j = \frac{1}{2}$ and $m = \frac{1}{2}$ or $-\frac{1}{2}$. If $l = 2$, $j = \frac{5}{2}$ or $\frac{3}{2}$, and so on.

Clearly, both j and m necessarily have half-integral values,, in harmony with the general rule stated in Sec. 130. The relations between l, j , and m may be exhibited in a vector diagram of the same general type as that employed for 1 electron in Sec. 115(b). Two

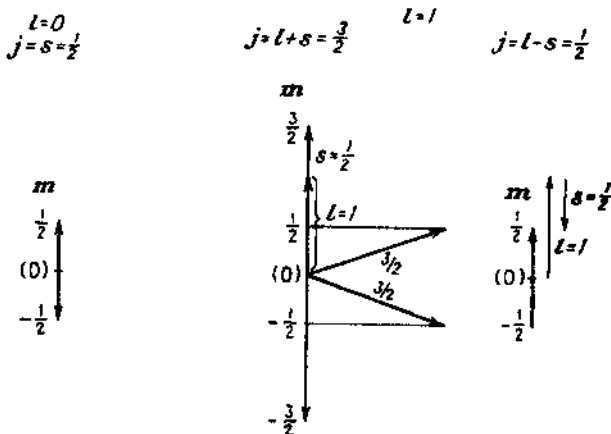


FIG. 93.—Diagrams illustrating addition of l and s -vectors into a j -vector.

cases are shown in Fig. 93. The s -vector is drawn in either the same or the opposite direction to the l -vector, to form a resultant j -vector; and this vector is then drawn at such an angle to the chosen axis that its projection on the axis is of length m .

The same selection rules are found to hold for j and m that were stated in Sec. 130 for the corresponding atomic numbers J and AT , viz.,

$$\Delta j = 0 \text{ or } \pm 1 \text{ (and not } j = 0 \text{ to } j = 0);$$

$$\Delta m = 0 \text{ or } \pm 1.$$

For a single electron there also exists the special rule, already mentioned in Sec. 131, that l must (in general), change by unity, or

$$\Delta l = \pm 1.$$

The energy is found, as before, to depend chiefly upon the principal quantum number n but also to some extent upon l (Sec. 118); and because of the spin-orbit interaction it also varies a little with j .

Thus the single energy level characterized by values of n and l which is found to occur when the spin effect is ignored, is split by the spin-orbit effect into two levels corresponding to $j = I \pm \frac{1}{2}$, except in the case $l = 0$. Usually the energy is (algebraically) higher for the larger value of j , viz., $j = I + s$, just as it is in the analogous classical motion when the two vector angular momenta have the same direction.

A set of quantum states is thus obtained which is characterized by the quantum numbers $n I j m$ instead of $n I X u$. In most of these new states the orbital momentum taken by itself does not have a fixed component in the direction of the axis, nor does the spin momentum; hence the quantum numbers X and u do not exist. The only exceptions are the states, two in number for each l , which have

$$j = l + s = l + \frac{1}{2}$$

and $m = \pm j$; for these states either $X = I$ and $u = M$ or $X = -\frac{1}{2}$, $u = -\frac{1}{2}$ respectively.

The general absence of definite X and u values is understandable if it is recalled that a quantum state corresponds not to some particular classical motion, which may possess special properties, but in a certain averaging fashion to a whole range of classical motions. The general case in classical mechanics would be that the spin angular momentum is inclined at some angle to the vector orbital momentum; then it is easily seen that many different orientations in space are possible, all of which allow the same value of the resultant angular momentum and of its component along the axis, but in which the components of the spin and orbital momentum taken separately vary.

Even if the spin-orbit interaction itself is omitted entirely, the set of $n I j m$ states may be employed in place of the $n l \backslash \backslash$ states; the two sets are equally valid alternatives. With the spin-orbit interaction omitted, the energy is the same for $j = I - \frac{1}{2}$ as for $j = I + \frac{1}{2}$ and thus varies only with n and l .

The number of the $j m$ states belonging to given values of n and l is the same as the number of the $X J U$ states. It is easily seen that there are $2(j + \frac{1}{2})$ or $2l + 2$ of the $j m$ states with different values of n , for $j = I + \frac{1}{2}$, and $2(j - \frac{1}{2})$ or $2l$ states for $j = I - \frac{1}{2}$, or a total of $2(2l + 1)$ states. This is the same number as that of the original $X J U$ states, where $2l + 1$ different values of X were associated each with two alternative values of u . It is a general rule that, in cases of degeneracy, the number of the independent quantum states is always the same, however these states may be chosen.

Wave Functions for the $j m$ States.—A few words may be added concerning the wave functions. For an electron in a central field, the

wave function representing one of the jm states contains a radial factor analogous to that in Eq. (186) for the hydrogen atom, but this factor varies somewhat with j as well as with n and L . The directional factor, on the other hand, is in general a linear combination of the directional factors for two of the XM states. These two states are associated with opposite spins; for one $\lambda = m - \frac{1}{2}$ and $\mu = \frac{1}{2}$, for the other $\lambda = m + \frac{1}{2}$ and $\mu = -\frac{1}{2}$. However, if $j = l \pm \frac{1}{2}$, only one $\lambda\mu$ factor is used, that with $X = I$ and $j = \frac{1}{2}$ for $m = j$, and that with $\lambda = -l$ and $\mu = -\frac{1}{2}$ for $m = -j$.

If the spin-orbit interaction is ignored, the radial factor is found not to differ at all for the two values of j which are associated with a given value of Z , and we merely have an alternative set of wave functions which are individually equal to certain linear combinations of the $n l J L$ wave functions.

134. Fine Structure in Alkali-type Spectra.—The theory developed in the last section is immediately applicable to the energy levels of a single valence electron outside of one or more closed subshells, as described in Sec. 131. Since the closed subshells do not contribute to the angular momenta, the quantum numbers $l j m$ may be replaced by the numbers $L J M$, which are understood to have reference to the whole atom.

The effect of spin-orbit interaction is then to split each of the energy levels or terms as previously described into two, one with $J = L + \frac{1}{2}$ and the other with $J = L - \frac{1}{2}$. Only the S terms ($L = 0$) are single, with $J = \frac{1}{2}$. Almost always the level with the larger J is found experimentally to have the higher energy and, therefore, the smaller term value, in agreement with the conclusion derived from wave mechanics. A J level is commonly indicated by adding the value of J as a subscript to the letter that is employed to denote the value of L ; and a superscript 2 may be written in front of the letter to indicate that the level belongs to a doublet system.

Transitions between two given terms will then give rise to several spectral lines, forming a spectral multiplet. The number of the lines is limited by the selection rule, $\Delta J = 0$ or ± 1 (Sec. 130). Jumps are possible between the single level of any S term, $^2S_{\frac{1}{2}}$ and either of the two levels in a P term, $^2P_{\frac{1}{2}}$ and $^2P_{\frac{3}{2}}$, for which $\Delta J = 0$ or 1, respectively. Between a P and a D , with levels $^2D_{\frac{1}{2}}$ $^2D_{\frac{3}{2}}$, three jumps are possible, viz.,

$$D3_{\frac{1}{2}} - ^2P_{\frac{1}{2}}, \quad ^2D_{\frac{1}{2}} - 2P_{\frac{3}{2}}, \quad 2D_{\frac{3}{2}} - ^2P_{\frac{3}{2}};$$

the jump $^2D_{\frac{5}{2}} - ^2P_{\frac{1}{2}}$ is forbidden, since for it $\Delta J = 2$. A triplet of lines results in a similar way from the allowed transitions between

any other pair of terms. Thus the chief spectrum of the alkali metals should consist of spectral doublets and triplets. The origin of the lines just mentioned is illustrated in Fig. 94. Numbers are also added in the figure showing the relative intensities that the lines of each multiplet should have according to wave mechanics, under certain simple conditions.

For comparison with the theory, the wave numbers and separations of the first four lines of each of the chief series in the spectrum of neutral sodium are listed in Table IV.¹

The theoretical conclusions as to wave length are seen to be confirmed by the observations. The sharp series, $nS \rightarrow 3P$, consists of

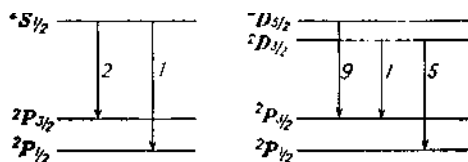


FIG. 94. - Transitions between doublet levels.

spectral doublets, and, as shown in column 5 of Table IV, the frequency difference is constant within the experimental errors, representing the difference between the $3^2P_{1/2}$ and $3^2P_{3/2}$ levels. This fact helps to confirm the identification of the S levels as made above.

The principal series of lines, $nP \rightarrow 3S$, also consists of doublets but with frequency differences that represent the spacings between the two levels in various P terms and hence decrease rapidly with increasing n . This decrease is in accord with theoretical expectations. For an electron in a Coulomb field, due to a nucleus of charge Ze , the separation of the two levels in a term with quantum numbers n and l , due to the spin-orbit effect, is readily calculated to be

$$\Delta\tilde{\nu} = \frac{\alpha^2 R Z^4}{n^3 l(l+1)} = 5.87 \frac{Z^4}{n^3 l(l+1)} \text{ cm.}^{-1} \quad (200)$$

where α is the "fine-structure constant," or $\alpha = 2\pi e^2 / ch = 1/136.7$, and R is the Rydberg constant (in cm.^{-1}). This expression is obtained by subtracting the value of W_j as given in Eq. (204) for $j = l - \frac{1}{2}$ from its value for $j = l + \frac{1}{2}$. Thus in a hydrogenlike atom the spin-orbit separation decreases as $1/n^8$. The observed decrease in the spectrum of sodium is even more rapid.

¹ Except for $\lambda 11,404$ and $\lambda 11,382$, the wave lengths of the lines were taken from the "Massachusetts Institute of Technology Wavelength Tables" and converted into wave numbers with the help of Kayser's "Tabelle der Schwingungszahlen"; the term values were then calculated from the wave numbers, the value of the SS term (41,499.0) being taken from Bacher and Goudsmit's "Atomic Energy States."

TABLE IV.—SOME LINES IN THE SPECTRUM OF NEUTRAL SODIUM

Initial term, cm. ⁻¹		$\bar{\nu}$, cm. ⁻¹ *	λ , angstroms	Doublet $\Delta\lambda$
Principal series, $n^1P \rightarrow 3^1S$. $3S = 41,449.0$ cm. ⁻¹				
$3P_{1/2}$	24,492.8	16,956.19	5,895.92	17.19
$3P_{3/2}$	24,475.6	16,973.38	5,889.95	
$4P_{1/2}$	11,182.1	30,266.91	3,302.99	6.15
$4P_{3/2}$	11,175.9	30,273.06	3,303.32	
$5P_{1/2}$	6,406.4	35,042.64	2,852.83	2.46
$5P_{3/2}$	6,408.8	35,040.18	2,853.03	
$6P_{1/2}$	4,151.4	37,297.64	2,680.34	1.40
$6P_{3/2}$	4,152.8	37,296.24	2,680.44	
Sharp series, $n^1S \rightarrow 3^1P$.				
$4S$	15,709.4	8,766.5	11,404.2	16.6
		8,783.1	11,382.4	
$5S$	8,248.3	16,227.28	6,160.76	17.22
		16,244.50	6,154.23	
$6S$	5,077.3	19,398.38	5,153.64	17.13
		19,415.51	5,149.09	
$7S$	3,437.2	21,038.40	4,751.89	17.14
		21,055.54	4,748.02	
Diffuse series, * $n^1D \rightarrow 3^1P$.				
$3D$	12,276.12	12,199.50	8,194.81	17.20
		12,216.70	8,183.27	
$4D$	6,900.29	17,575.33	5,688.22	17.19
		17,592.52	5,682.66	
$5D$	4,412.3	20,063.30	4,982.84	17.16
		20,080.46	4,978.58	
$6D$	3,061.9	21,413.72	4,668.60	17.18
		21,430.90	4,664.86	

* Only the mean value of the $D_{5/2}$ and $D_{3/2}$ levels is shown and only the two strong lines,

$$nD_{5/2} \rightarrow 3P_{3/2}, \quad nD_{3/2} \rightarrow 3P_{1/2}.$$

The lines of the *diffuse series*, $nD \rightarrow 3P$, should be *triplets*, according to the theory. One line should be very weak, however. The two brightest lines result from the jumps

$$n^2D_{5/2} \rightarrow 3^2P_{3/2}, \quad n^2D_{3/2} \rightarrow 3^2P_{1/2}.$$

If Eq. (200) holds roughly for sodium, then, because of the factor

$1(1 + 1)$ in the denominator, the separation $nD_{3/2} - nD_{5/2}$ stands to the separation $nP_{3/2} - nP_{1/2}$ in the ratio $1/(2 \times 3): 1/(1 \times 2)$ or as only 1:3; and, as n increases, the D separation should rapidly diminish further. Thus with ordinary resolving power the lines of the diffuse series can easily be observed with a frequency difference that is nearly constant and equal to the difference $3P_{3/2} - 3P_{1/2}$. Doublets of exactly this separation are formed of course by the faint line $D_{3/2} \rightarrow 3P_{3/2}$ and the brighter line $D_{3/2} \rightarrow 3P_{1/2}$. The faint line is not usually seen; when it was discovered, it was called a "satellite" of the apparent doublet formed by the two brighter lines.

Thus, the theory accounts very well for the principal features of the spectrum emitted by neutral sodium atoms. It is equally success-

TABLE V.—FIRST LINES OF THE PRINCIPAL SERIES FOR THE ALKALI METALS AND ONE LINE OF HYDROGEN
X is in angstroms, ν in cm^{-1}

	H	Li	Na	K	Rb	
Z	1	3	11	19	37	
n	$3 \rightarrow 2$	2	3	4	5	
$\lambda_{3/2}$	6,562.8	6,707.8	5,895.9	7,699.0	7,947.6	8
$\lambda_{5/2}$			5,890.0	7,664.9	7,800.2	8
$\tilde{\nu}_{3/2}$	15,233	14,904	16,956	12,985	12,579	11
$\tilde{\nu}_{5/2}$			16,973	13,043	12,817	11
$\Delta\tilde{\nu}$	0.367	0.34	17	58	238	

ful with the other alkali metals, the spectra of which are qualitatively very similar to that of sodium.

The separation of the doublet levels, however, increases rapidly with increasing Z . In Table V are shown the wave lengths and the wave numbers of the D lines or their analogs (*i.e.*, the first lines of the principal series) for all of the alkali metals, and also the doublet differences for these lines. The theoretical value of $\Delta\nu$ for the first Balmer line of hydrogen, obtained by setting $Z = 1, n = 2, l = 1$ in Eq. (200), is also given for comparison. The data in the table show that the D lines, which are 6 angstroms apart in the sodium spectrum, are 422 angstroms apart in the spectrum of cesium.

The enormous departure of $\Delta\nu$ from the hydrogen value for all alkali metals except lithium, in spite of the progressive increase in n , is in strong contrast with the more moderate departure of the energy levels themselves, as shown by Fig. 90. This may be regarded as

resulting from great sensitiveness of the spin-orbit effect to the character of the central field near the nucleus, which is clearly evident from the expression for the spin-orbit energy as given in Eq. (199). Near the nucleus the field of the nucleus itself must predominate, so that approximately $V = -Ze^2/r$, and

$$\frac{1}{r} \frac{dV}{dr} = \frac{Ze^2}{r^3},$$

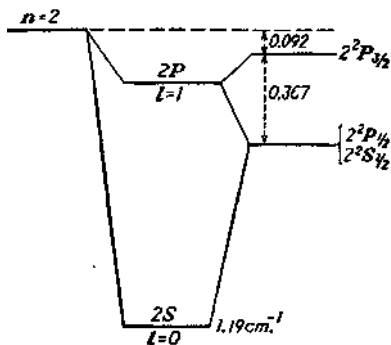
which increases much more rapidly than does the numerical value of V itself as the nucleus is approached. The variation of the spin-orbit effect among the alkalis furnishes an excellent illustration of the tendency of this effect to be small in atoms of low atomic number but to increase to an enormous magnitude as the atomic number becomes large.

135. Multiplet Levels for One-electron Atoms.—In the last section, a theory of the fine structure of spectral lines due to spin was developed for atoms containing a valence electron outside of closed subshells. The same theoretical treatment should be applicable also to atoms containing no subshells at all, *i.e.*, to atoms such as those of hydrogen which contain *only a single electron*. In this latter case, it is customary to use small letters, Ijm (or m_i) instead of LJM ; then $j = \frac{1}{2}$ if $I = 0$, otherwise, $j = I + \frac{1}{2}$ or $I - \frac{1}{2}$.

In the 1-electron atom, however, a curious accident occurs. In the nonrelativistic theory, as we have seen in Sec. 115, all states for a given n have the same energy, regardless of the value of I . In the relativistic theory this is not quite true. The wave equation contains certain other small terms of relativistic origin, besides those giving rise to the spin-orbit effect, and these other terms cause the energy to vary somewhat with I . Now, in a 1-electron atom this relativistic variation happens to be of the same order of magnitude as the spin-orbit effect itself; in fact, the net result of both effects is that levels having different values of I but the same value of j fall exactly on top of each other. The energy diagram for such atoms is thus very peculiar. Somewhat similar relativistic effects occur also in atoms containing 2 electrons (and in them the interaction between the spins of the 2 electrons is important); but in such cases the relativistic displacement of the various L terms is much smaller than is their displacement due to the electrostatic repulsion between the electrons, so that the relativistic correction is unimportant.

The situation in the 1-electron atoms is illustrated in Fig. 95, which refers to the hydrogen levels with $n = 2$. The single line at the left represents the energy as given by Bohr's theory, Eq. (136) in

Sec. 94. The two solid lines in the center of the diagram show this level as split into two, a P and an S term, in consequence of what is called the "ordinary" relativistic correction. The addition of the spin-orbit effect then splits the $2P$ term into the usual doublet with $j = \frac{1}{2}$ or $3/2$; but an additional "special" spin correction raises the $2S$



level so that it coincides with the $2P_{1/2}$ level. The net result, as shown at the right in the figure, is just *two separate levels*, a single one with $j = 3/2$ and a composite level with $j = \frac{1}{2}$. In a similar way, the $3S_{1/2}$ level comes to coincide with the $3P_{1/2}$ level, and $3P_{3/2}$ also coincides with $3D_{3/2}$ whereas $3D_{5/2}$ stands alone; and so on. Thus for each value of n there are only n different energy levels, with $j = \frac{1}{2}, 3/2, 5/2, \dots$ $n - \frac{1}{2}$; but all levels except that for the largest $j, j = n - \frac{1}{2}$ are

double. The lowest level of all, with $n = 1$, is still single.

Exact formulas for the energy levels can be obtained by solving the Dirac relativistic wave equation (Sec. 117). The result for the energy of a level specified by quantum numbers n and j is¹

$$W_{nj} = mc^2 \left\{ \left[1 + \frac{\alpha^2 Z^2}{(n - j - \frac{1}{2} + \sqrt{(j + \frac{1}{2})^2 - \alpha^2 Z^2})^2} \right]^{-\frac{1}{2}} - 1 \right\}.$$

$$\alpha = \frac{2\pi e^2}{ch} = 0.007314 = \frac{1}{136.7}.$$

expansion. By patiently applying the binomial theorem to one radical or fraction after another, one finds eventually that

$$W_{ni} = -\frac{mc^2\alpha^2Z^2}{2n^2} + \frac{mc^2\alpha^4Z^4}{2n^4}\left(\frac{3}{4n} - \frac{1}{j + \frac{1}{2}}\right) \dots$$

ergs. Let us divide this by hc to convert it into wave-number units and then introduce the Rydberg constant or [cf. Eq. (1466) in Sec. 97]

$$R = \frac{2\pi^2me^4}{ch^3} = \frac{mc\alpha^2}{2h} \quad (202)$$

by (201). Then, in wave-number units, as far as terms in α^2 ,

$$W_{ni} = -\frac{RZ^2}{n^2} + \frac{\alpha^2RZ^4}{n^4}\left(\frac{3}{4n} - \frac{1}{j + \frac{1}{2}}\right). \quad (203)$$

The first term in this expression for W_{ni} is the ordinary nonrelativistic value of the energy [Eq. (187) in Sec. 115]. The second term represents the combined relativity-spin correction; it cannot exceed in magnitude a fraction α^2 or 5.35×10^{-5} of the first term. The part of the second term that varies with j can be written

$$\Delta_i W = -\frac{\alpha^2RZ^4}{n^3(j + \frac{1}{2})} = -5.87 \frac{Z^4}{n^3(j + \frac{1}{2})} \text{ cm.}^{-1} \quad (204)$$

We note that this term in the energy, which determines the fine structure of the energy levels, decreases rapidly with increasing n .

The *approximate* formula (203) is easily obtained by making in succession the two corrections mentioned above, with the help of the perturbation theory, to be described in Sec. 137. It may be of interest to write down the results of the successive steps. The *ordinary relativistic* correction to the energy is found to be

$$\Delta_i W' = \frac{3}{4} \frac{\alpha^2RZ^4}{n^4} - \frac{\alpha^2RZ^4}{n^3(l + \frac{1}{2})}. \quad (205)$$

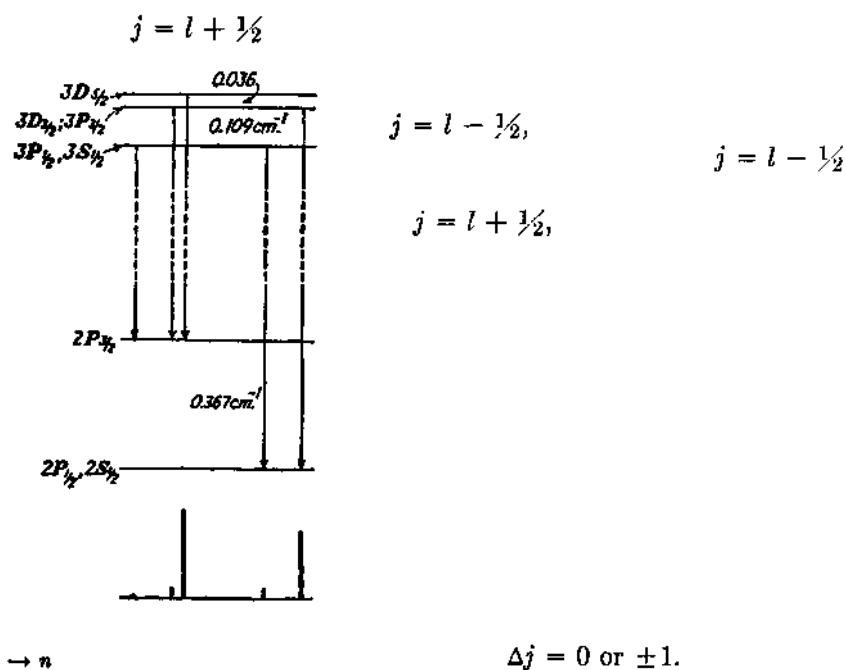
this is the same as Sommerfeld's correction for variable mass. It causes the splitting of each level for given n into n sublevels with $l = 0, 1, \dots, n - 1$. The *spin-orbit* correction is then found to be, or $l > 0$:

$$j = l + \frac{1}{2}: \quad \Delta_i W = \frac{1}{l + 1} \frac{\alpha^2RZ^4}{n^3(2l + 1)}; \quad (206a)$$

$$j = l - \frac{1}{2}: \quad \Delta_i W = -\frac{1}{l} \frac{\alpha^2RZ^4}{n^3(2l + 1)}. \quad (206b)$$

For $I = 0$, the spin-orbit effect vanishes. The *special spin* correction is likewise found to vanish for $I > 0$, but for $I = 0$ it is

$$\frac{\alpha^2 R Z^4}{n^8}.$$



$$3D_{3/2} \rightarrow 2P_{3/2}, 3D_{3/2} \rightarrow 2P_{3/2}, 3D_{3/2} \rightarrow 2P_{1/2},$$

$3P_H \rightarrow 2S_{1/2}$ $3P_{1/2} \rightarrow 2S_{1/2}$ $3S_{1/2} \rightarrow 2P_{1/2}$, and $3S_{1/2} \rightarrow 2P_{3/2}$. The relative spacing of the five distinct lines is shown in the lower part of Fig. 96, the heights of the lines as drawn representing the theoretical relative intensities of the spectral lines on the assumption that all five quantum states for $n = 3$ are equally excited (i.e., that as many atoms are excited into one state as into another).

(6) *Comparison with Experiment for Hydrogen.*—The "lines" of the Balmer series of hydrogen were early observed as close doublets. In 1887, Michelson and Morley measured the doublet separation for the line of longest wave length, $H\alpha$ or $n = 3 \rightarrow n = 2$, and found for it 0.253 cm^{-1} . Subsequent measurements by other investigators gave values ranging from 0.293 to 0.357. The modern interpretation of the doublet appearance is that the five component lines are smeared

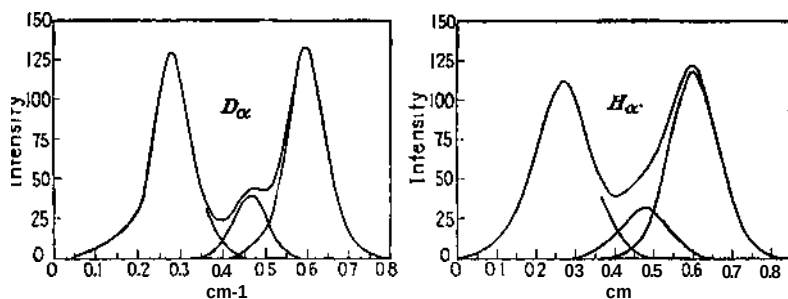


FIG. 97.—Typical intensity curves obtained for the $H\alpha$ line from hydrogen ($H\alpha$) and deuterium (Da). Intensity is on an arbitrary scale, and the abscissa is numbered from an arbitrary point. {After Williams.}

together because of the Doppler broadening of all lines due to the thermal motion of the molecules, but two of them are much more intense than the others. The only means of comparing the observations with the theoretical predictions is, therefore, to calculate the contour of the total line from the theory, using the theoretical separations of the fine structure and the theoretical estimates of the relative intensities of the component lines, and then making an approximate allowance for the Doppler effect.

The line $H\alpha$ has been studied very carefully in this manner. In Fig. 97 are shown typical curves obtained by R. C. Williams¹ for the spectral distribution of intensity in the $H\alpha$ line from ordinary hydrogen and from deuterium, or heavy hydrogen, the atoms of which are about twice as heavy as those of ordinary hydrogen and ought, therefore, to be influenced less by the Doppler effect [Sec. 152(a)]. The discharge tube was cooled by liquid air to **minimize the broad-**

¹ WILLIAMS, *Phys. Rev.*, vol. 54, p. 558 (1938).

ening. The light from the tube was dispersed by a triple-prism spectrograph containing a quartz Fabry-Perot etalon placed in the parallel beam of the collimator, and photographs were taken of the spectral region containing the Ha line. The blackening on the film was measured with a Moll microphotometer using a thermocouple and galvanometer, and from these measurements the relative distribution of intensity in the line could be determined.

At most three of the theoretical component lines are evident in Fig. 97; one of the others is too close to the left-hand main line to be resolved, and the remaining one is evidently too weak to be seen. Theoretical line shapes, as influenced by the Doppler effect, are drawn in the figure for the three lines, the assumed strengths of the lines being so adjusted as best to reproduce the observed curve of resultant intensity. The strengths of the lines as thus inferred from the observations agree only roughly with the theoretical predictions. The left-hand line ought to be somewhat stronger and the middle one only about half as strong. Such deviations might be due, however, to unequal excitation of the initial levels ($S\frac{1}{2}$, $P\frac{3}{2}$, $D\frac{5}{2}$). A much more serious discrepancy is that the *spacing* of the lines does not quite agree with the theoretical predictions. The distance between the two main peaks was found by Williams to be consistently 0.319 to 0.321 cm^{-1} for Da, 0.315 to 0.319 cm^{-1} for Ha, whereas wave-mechanical theory predicts, from the level separations as shown in Fig. 96, $0.367 - 0.036 = 0.331 \text{ cm}^{-1}$. The small central peak seems also to occur at about 0.134 cm^{-1} from the right-hand one, whereas the theory gives for this separation 0.109 cm^{-1} . The cause of these discrepancies between theory and observation—whether they are due to some unrecognized experimental complication or to some further slight correction required in the theory—is not yet known.¹

The remark may be added that, because of the fine structure and the relativity correction, the simple Balmer formula for the frequencies of the Balmer series is not quite adequate. The following empirical formula of the more general Rydberg type gives very closely the wave numbers of the centers of the lines of the Balmer series of hydrogen when seen unresolved:²

$$\tilde{\nu} = 109,678.28 \left[\frac{1}{(2 - .00000383)^2} - \frac{1}{(m + .00000210)^2} \right]$$

$m = 3, 4, 5, \dots$

¹ Cf. 'DRINKWATER, RICHARDSON, and W. E. WILLIAMS, *Roy. Soc. Proc.*, vol. 174, p. 164 (1940).

² FOWLER, "Report on Series in Line Spectra," Fleetway Press, London, 1922.

(c) *Ionized Helium*.—According to Eq. (204), the fine-structure separations should be 2^4 or 16 times as great in lines emitted by an ionized helium atom as in the corresponding lines from hydrogen. A careful study was made by Paschen¹ of the line $\lambda = 4,686 \text{ \AA}$ from ionized helium. At the time, his observations were important because they confirmed the wave-mechanical theory as against Sommerfeld's extension of Bohr's theory.

The line in question results from the jump $n = 4$ to $n = 3$ and so corresponds to the longest line of Paschen's infrared series for hydrogen. From the energy-level diagram, shown in Fig. 98, and the selection rules, it is evident that there should be eight fine-structure lines, here labeled *abcdefgh*. According to Sommerfeld's theory the lines *c* and *g*, arising from jumps in which $\Delta j = 0$, should not appear. In place of $j + \frac{1}{2}$ in the formulas, Sommerfeld wrote a quantum number referring to the orbital angular momentum and this quantum number was subject to the requirement that in a transition it must change by unity. This rule rested on grounds equivalent to those that lead to the modern rule that $\Delta Z = \pm 1$, and it seemed to be very hard to eliminate it from the theory.

Paschen's observations are compared with the theoretical predictions in Fig. 99. Components *a* and *b*, as predicted by both theories, are undoubtedly unresolved in Paschen's *ab*. The weak components, *c* and *e*, were probably unresolved from *d* and *f*. But Paschen found a fairly strong component *g*, which is predicted by the newer theory but is absent in the older one. This fact is decisive in favor of wave mechanics.

137. LS or Russell-Saunders Coupling.—In dealing with the spectra emitted by atoms containing a single valence electron it was possible largely to ignore all the other electrons because they were grouped into closed subshells. Effective allowance for the electrostatic repulsive action of these electrons on the valence electron could be made by assuming a suitable modified central field in which the valence electron was assumed to move. This procedure is no longer

PASCHEN, *Ann. d. Physik*, vol. 82, p. 689 (1927)..

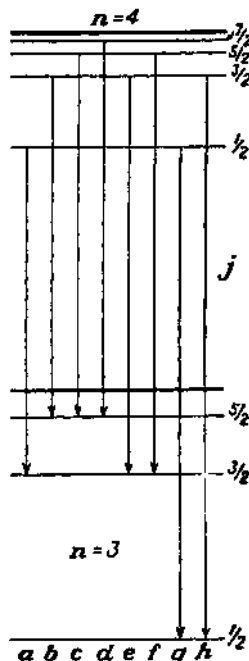


FIG. 98.—Diagram for the fine structure of the line $\lambda 4,686$ ($n = 4 \rightarrow n = 3$) of ionized helium. The two heavy lines indicate the levels according to the original Bohr theory.

adequate when there are 2 or more electrons outside the last closed subshell. Then, in addition to spin-orbit effects, the mutual repulsion of these outer electrons must be dealt with more explicitly.

The principal mathematical method that is available in extending the wave-mechanical theory to such atoms is what is called the "perturbation" theory already mentioned in Sec. 120. The most important results that follow from the use of this method will be described here in terms of quantum states and the associated angular momenta. In the next section, a few remarks will be added concerning the mathematical method itself.

In applying the perturbation method, both the residual effect of electrostatic repulsion, over and above the average part of this effect

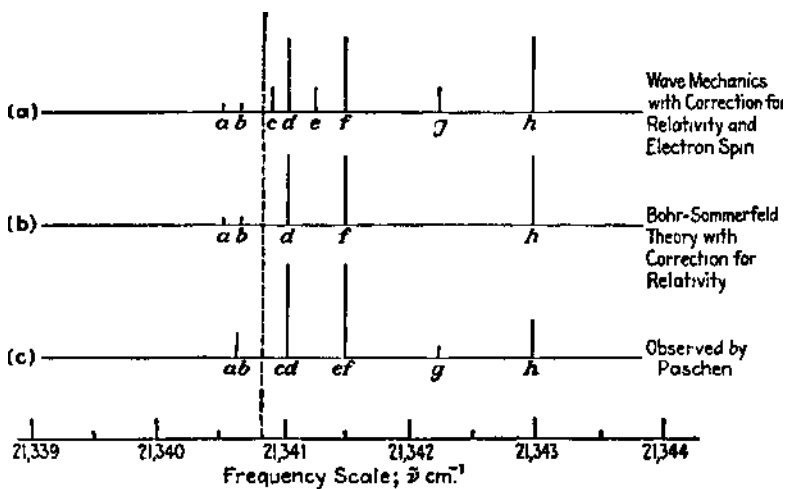


FIG. 99.—Fine structure of the helium line X4,686 (cf. Fig. 98). The dotted line gives the position of the single line predicted by Bohi's theory. Lengths of lines give relative intensities (qualitatively). *

that is included in the central field, and all spin-orbit interactions are at first ignored. Mathematical expressions representing the quantum states for the atom are then easily found. Afterward, a correction is introduced for the omitted effects. If these effects are small, the correction for them possesses high accuracy. The method is named after the analogous method used by astronomers in calculating the motions of the solar system; in the astronomical calculation, each planet is assumed to move in an elliptical orbit around the sun, and then corrections are made for the perturbations due to the attractions of the other planets.

When the outer electrons are assumed to move in a spherically symmetrical field, without mutual repulsion and without any spin-

orbit interactions, each electron can be assumed to occupy an electronic quantum state labeled with quantum numbers $n\ l\ \lambda\ \mu$, as described in Sec. 120. If N electrons are simultaneously present in the field, they will occupy n such quantum states, characterized by quantum numbers $n_1\ l_1\ \lambda_1\ \mu_1, n_2\ l_2\ \lambda_2\ \mu_2, \dots, n_N\ l_N\ \lambda_N\ \mu_N$. Any such set of n electronic quantum states determines a quantum state of the atom, which is called in perturbation theory a state of "zero order." Alternatively, however, we can also start with electronic quantum states of the type employed in Sec. 133, where the states are labeled with the numbers $n\ I\ j\ m$. Which type of electronic state it is best to use depends upon circumstances. In the present section the $n\ l$ n states will be assumed.

With either type of electronic state, so long as spin-orbit effects are ignored, the energies of the electronic states themselves, and also the energy of the whole atom, depend only upon the values of n and I . A set of values of n and I , $n_1\ l_1, n_2\ l_2, \dots, n_N\ l_N$, is said to define a *configuration* of the N electrons. A compact form of notation is in use among the spectroscopists for configurations. Thus, $2s3d$ denotes a configuration of 2 electrons in which, for 1 electron $n = 2$ and $I = 0$, while for the other $n = 3$ and $I = 2$. Again, lp^22p means that there are 2 electrons with $n = 1$ and $I = 1$, and another with $n = 2$ and $I = 1$.

An important selection rule holds for the configurations. It was partly stated in Sec. 129. Radiative transitions occur only between levels belonging to two configurations which differ in just one electronic state; furthermore, the values of I in these two differing states must differ exactly by unity. Thus transitions may occur between a level belonging to a $2s3p$ configuration and one belonging to a $2s3d$ configuration, but not between $2s3p$ and $2p3d$, because here both electronic states are different, and not between $2s3p$ and $2s4f$, because here $\Delta I = 2$ for the second electronic state. The existence of this selection rule for configurations results in a great simplification, because it diminishes the number of possible transitions that have to be considered.

Let it be assumed for the present that no 2 electrons have the same values of both n and I ; for example, the combination 2,1 might occur together with 1,2 or 2,2, but not with another 2,1. Then all values of the X 's and the u 's may be assumed freely in defining quantum states, without fear of conflict with the Pauli exclusion principle. The consequences of dropping this restriction will be described briefly in Sec. 141.

Now, taking a step forward in the perturbation argument, con-

sider the effect of introducing the residual electrostatic repulsion of the N electrons. If this effect is small, it will modify the atomic quantum states and energies only slightly, *provided* we start with the right set of initial zero-order states. The necessity for starting with the right initial states corresponds closely to the following classical analogy.

A pendulum bob hung on a stout, symmetrical wire can swing with a certain definite frequency ν_0 in any plane. It can swing east and west, for example, or north and south; and vibration in any other plane can be regarded as a combination of these two motions, each with a suitable amplitude and phase. As an alternative, any other pair of perpendicular planes might be chosen for the component vibrations. This is a case of degeneracy. But now squeeze the wire with pliers so as to flatten it in a NE - SW direction. Thereafter, steady oscillations in a fixed plane are possible in only two directions. The bob can swing either in a NE - SW plane with a frequency ν_1 or in a NW - SE plane with a frequency ν_2 , ν_1 and ν_2 being slightly different from each other and from ν_0 . If the bob is started swinging in any other direction, a combination of these two vibrations occurs and the bob executes a Lissajous figure. The two fundamental vibrations must now be chosen in two fixed directions; thus flattening the wire has removed the degeneracy.

In the atomic case a clue to the correct choice of the initial quantum states is furnished by considerations of angular momentum. If the electrostatic repulsion is fully active, according to classical mechanics the individual orbital vector angular momenta of the electrons will be subject to large fluctuations; but the vector sum of these angular momenta, or the total orbital momentum of the atom about the nucleus, will remain fixed, since the forces representing the mutual repulsion of any two electrons, although equal in magnitude, act in opposite directions along the same line. This fact suggests, in agreement with conclusions drawn from wave mechanics, that, in order to accommodate the residual electrostatic repulsion as a perturbation, the zero-order atomic states should be chosen so as to quantize the total orbital momentum. At the same time it is possible and advantageous to quantize the total spin momentum.

In zero order of the perturbation theory, with the electrostatic repulsion still missing, it is found possible to construct atomic states having the desired characteristics without disturbing the electronic quantum numbers $h, l_1, l_2 \dots l_N$, or their significance in terms of electronic momenta. Thus it is possible to speak of each atomic state as belonging to a definite electronic configuration. The electronic quantum numbers $\lambda_1 \mu_1, \lambda_2 \mu_2, \dots \lambda_N \mu_N$, however, disappear

and are replaced by four atomic numbers. Two of these atomic numbers may be defined as follows: The total orbital momentum has a magnitude $\sqrt{L(L+1)} h/2\pi$, and the total spin momentum has a magnitude $\sqrt{S(S+1)} h/2\pi$, where L and S are positive numbers (or zero). The possible values of L that can occur for atomic states belonging to a given configuration always form a series differing from each other by unity, and the same is true for the possible values of S .

The possible values of L that can occur in a given configuration may be inferred from a vector diagram that represents in a limited sense the vector addition of the electronic orbital momenta into a resultant. This is illustrated for two different configurations of 2 electrons in Fig. 100; the diagram for any number of electrons is analogous. From the diagram it is obvious that the maximum possible value of L for a given configuration is $2J$, the sum of the values of I for all the electrons; and all values of L will be, like I for any electron, integral

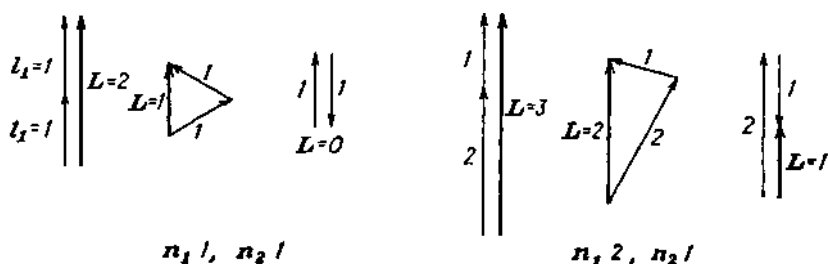


FIG. 100.—Diagrams illustrating addition of two l -vectors into an L -vector.

or 0. States occur with all values of L that differ from $2J$ by an integer and are allowed by the diagram. For example, if $l_1 = l_2 = l_3 = 1$, then alternative values of L are $L = 3$ or 2 or 1 or 0 . If $l_1 = 4$, $l_2 = 1$, $l_3 = 1$, then $L = 6, 5, 4, 3$, or 2 . If any $I = 0$, it may be disregarded.

As was noted in Sec. 131, letters are very frequently written in place of numbers to denote values of L : S, P, D, F, G, H, J, K , etc., stand, respectively, for $L = 0, 1, 2, 3, 4, 5, 6, 7 \dots$ (The letter S representing a value of L must not be confused with the quantum number S . This feature of the established notation is unfortunate.)

An exactly analogous method holds for S , as is illustrated for $N = 3$ and $N = 4$ in Fig. 101. Since all electronic spin vectors are of length $s = \frac{1}{2}$, the maximum value of S is obviously $S_s = N/2$, and all values of S derived from a given configuration must be integral (or zero) or half-integral according as the number N of electrons in the configuration is even or odd.

Simple selection rules are found to hold for L and S . This is one

of the most important properties of these numbers. For S the rule is that, in any radiative transition from one atomic state to another,

$$\Delta S = 0. \quad (207)$$

This restriction corresponds in classical mechanics to the fact that the existence of spin and of its associated magnetic moment has no appreciable effect upon the emission of radiation. Because of this selection rule, in the absence of all spin-orbit interaction, the atomic energy levels may be divided into broad, noncombining classes, each distinguished by a different value of S ; levels belonging to one class combine in the emission or absorption of spectral lines only with other levels belonging to the same class.

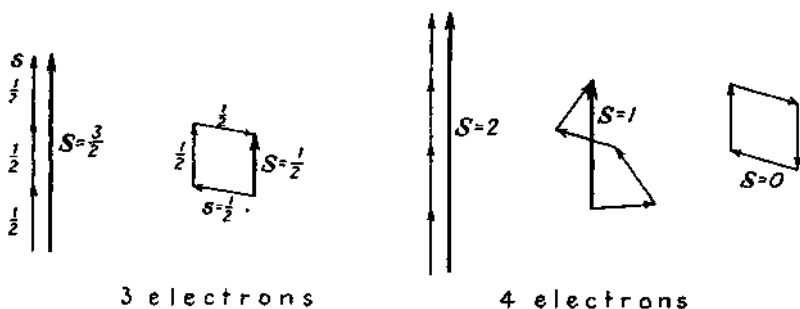


FIG. 101.—Diagrams illustrating addition of s -vectors into an S -vector.

The analogous selection rule for L is

$$\Delta L = 0 \text{ or } \pm 1 \text{ (and not } 0 \rightarrow 0 \text{)}.$$
 (208)

Thus S levels ($L = 0$) can combine only with P levels ($L = 1$); P levels can combine either with S or with D levels ($L = 2$); D levels can combine either with P or with F levels ($L = 3$); and so on. A special case of this selection rule has already been encountered in dealing with atoms that contain a single valence electron (Sec. 131). So long as the closed subshells in such an atom remain undisturbed, all momenta are determined by the momenta of the valence electron alone (Sec. 130); thus $S = s = \frac{1}{2}$ and only one class of atomic states occurs. The occurrence of atomic states characterized by quantum numbers L and S obeying the selection rules that have been stated is called LS or Russell-Saunders coupling (of the electronic momenta).¹

The individual atomic states are not completely determined, however, by their values of L and S . The specification of a state can be completed in many different ways, of which the most important yield, respectively, $L S A 2J$ or $L S J M$ states.

¹ RUSSELL and SAUNDERS, *Astrophys. J.*, vol. 61, p. 38 (1925).

In one of the $L S A \Sigma$ states, the total orbital momentum has a definite component of magnitude $\Delta h/2\pi$, and the total spin momentum has a component $\Sigma h/2\pi$, in the direction of any chosen axis, where A is a number that may assume any one of the integrally spaced values from $-L$ to L , and 2 may assume independently any one of the integrally spaced values from $-S$ to S . Thus A is always integral, whereas 2 is integral or half-integral according as S is. The selection rules for A and 2 , in the absence of spin-orbit interaction, are

$$\Delta A = 0 \text{ or } \pm 1, \quad \Delta \Sigma = 0. \quad (209a,b)$$

Alternatively, atomic states may be chosen in which the vector sum of the orbital and spin momentum has a fixed magnitude of

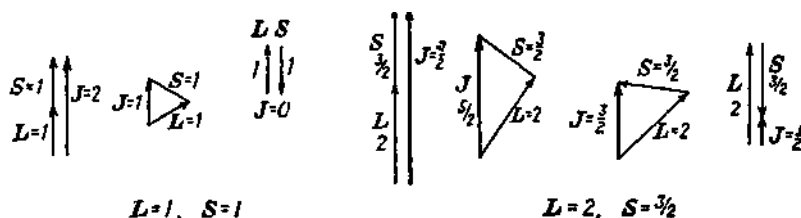


FIG. 102.—Diagrams illustrating addition of an L - and an S -vector into a J -vector.

, and its component in the direction of the chosen

axis has simultaneously the fixed value $Mh/2\pi$, where M may be any one of the integrally spaced numbers from $-J$ to J . This is an example of the $J M$ states described in Sec. 130, which are always possible states for an atom free from external force.

The values of J that can occur, in association with given values of L and S , may easily be found from another vector diagram. This one represents the vector addition of the L - and S -vectors, in correspondence with the vector addition of the classical orbital and spin momenta into a resultant momentum. Two cases are illustrated in Fig. 102. The maximum possible value of J is always $J = L + S$; then all integrally spaced values may occur down to $|(L - S)|$. Thus, if $S = 0$, $J = L$, only; if $L = 0$, $J = S$. In all other cases at least two alternative values of J are possible. If $S = 1/2$, $J = 1/2$ for $L = 0$, but $J = L + 1/2$ or $J = L - 1/2$ for any other value of L , as in the atoms with 1 valence electron (Sec. 134). If $S = 1$ and $L = 3$, then $J = 4$ or 3 or 2 . The selection rules for J and M are stated in Sec. 130.

The number of states in an $L S A \Sigma$ set is the same as the number in the corresponding $L S J M$ set; and in either case these states possess no further degeneracy (so long as nuclear spin is neglected).

In some configurations, however, two or more independent sets of such states occur having the same values of L and S .

As already stated, in the zero-order approximation the atomic energy is the same for all of the states belonging to a given configuration, whatever may be their values of L and S . This degeneracy of the atomic states becomes reduced, as we shall see in the next two sections, when allowance is made for the residual electrostatic repulsion of the electrons and for the spin-orbit effect.

Furthermore, in Sec. 142 an alternative form of electronic coupling known as *jj* coupling will be described. This type of coupling is especially useful in dealing with heavy atoms.

The LS Terms.—The electrostatic repulsion of the electrons, as already stated, has no effect upon the orbital and spin angular momenta of the atom as a whole. Hence, regardless of the strength of this repulsion, the atom can occupy a quantum state of either $L S A S$ or $L S J M$ type, and the appropriate selection rules will hold. The repulsion has, however, the effect of separating the various groups of states, each characterized by a value of L and of S , with respect to the atomic energy. The separated energy levels thus produced may be called *LS* levels, or, when the negative of the atomic energy is used, *LS* terms.

It seems obviously reasonable that the residual electrostatic repulsion should separate terms that differ only in the value of L ; but it may seem strange that electrostatic repulsion should separate even terms having the same L but different values of S , since the repulsion has nothing to do with the magnetic moments associated with electronic spin. In the mathematical theory the repulsion is effective in such cases by means of the feature called the "electronic exchange effect," which has no analog in classical mechanics. This same feature was cited in Sec. 125 as the cause of covalent bonding in chemical compounds. A difference in the arrangement of the electronic spins requires that the atomic wave function shall be a different function of the coordinates of the electrons, and these differences in the wave function are associated with differences of energy.

The electrostatic repulsion of the electrons also disturbs the separate angular momenta of the individual electrons. If the repulsion is weak, however, its effect in classical mechanics is chiefly to cause a precession of the normals to the orbital planes of the electrons, similar to the precession of a gyroscope under applied forces; there is little change in the actual magnitudes of the electronic momenta themselves. Analogously, in an atom in which the electrostatic repulsion is pretty effectively represented by the electronic contribution to the central

field, so that only a small residual repulsion remains for further consideration, each atomic state can be assigned definitely to a certain electronic configuration, as defined by a certain set of the electronic quantum numbers n and l . The selection rule for configurations will then hold well for the emission of spectral lines. At the other extreme, however, there are cases in which the residual repulsive effect is so strong that it is difficult or impossible to assign certain LS terms to any particular configuration. In such cases the quantum numbers n and l lose their significance, and the selection rule for configurations fails more or less completely. Many intermediate cases also occur.

Note on the Wave-mechanical Theory.—The discussion of the mathematical theory that was given in Sec. 120 may be resumed here for a moment. The construction of wave functions representing the zero-order $LSA\ \Sigma$ or $LSJ\ M$ states belonging to a given configuration of the electrons is effected by making suitable linear combinations of the atomic wave functions as these are first obtained out of the electronic $n\ l\ m$ wave functions. The requisite linear combinations are analogous in mathematical form to those mentioned in Secs. 115 and 120. The appropriate selection rules are then discovered by calculating matrix components of the electronic coordinates; this procedure constitutes an extension of the calculations mentioned in Sec. 116 (c).

When the wave equation is then corrected by adding terms to represent the residual electrostatic repulsion and the spin-orbit interaction for each electron, the zero-order wave functions are no longer exact solutions of the equation. If, however, the effect of these added terms in the wave equation is small, the zero-order wave functions may constitute very good approximations to the exact solutions, provided the zero-order functions are properly chosen from among the alternative sets that are possible because of the degeneracy. A general mathematical method exists for selecting the proper linear combinations of the zero-order functions, and it leads to the selection of functions having the properties in terms of angular momenta that have been stated in the text.

A first-order correction to the atomic energy that is associated with any approximate wave function ψ can then be found by evaluating the integral

$$\int \psi^* H_1 \psi dq$$

where ψ^* denotes the complex conjugate of ψ , H_1 represents the added term or terms in the wave equation, and $\int \dots dq$ indicates an integration over the entire range of possible values for all coordinates of all

electrons in the atom. A first-order correction for ψ itself can also be formed; and then a second-order correction to the energy, and so on; but the mathematical difficulties in the way of doing this are so great that the procedure has hardly ever been carried beyond the first order.

Further details of the mathematical theory must be sought in other books or in the literature.

138. *LS Multiplets*.—As has been stated, each *LS* level corresponds, in general, to a group of distinct quantum states for the atom, which can be so chosen that they are appropriately labeled by the quantum numbers L and M . Degeneracy exists, so long as spin-orbit effects are ignored, in that in all the states forming a given *LS* set the atomic energy has the same value.

When, finally, the spin-orbit interactions of the electrons are taken into account, this degeneracy is partly removed. The atomic states continue to form groups each characterized by a certain value of J , but the groups themselves are separated from each other in energy. Thus each of the original *LS* levels is, in general, split into two or more levels characterized by different energies and by different values of J .

If the spin-orbit effect is relatively *strong*, the orbital and spin momenta of the atom are not separately quantized, and significant quantum numbers L and S do not exist. A case of this sort will be discussed later, under the heading of *jj* coupling (Sec. 142). If, on the other hand, the *spin-orbit effect is much weaker than the residual repulsive effect*, the orbital and spin momenta are little affected by the spin-orbit interaction; the quantum numbers L and S then retain approximate validity, and the associated selection rules hold well. In this case we have, out of each of the original *LS* levels, a close group of *LSJ* levels called an *LS multiplet*. Only this case will be considered further at this point.

As was stated in the last section, the values of J belonging to an *LS* term include all integrally spaced values that satisfy the inequality

$$|L - S| \leq J \leq L + S. \quad (210)$$

Thus, if $L = 0$, it follows that $J = S$, whereas if $S = 0$, $J = L$, and in either case the multiplet reduces to a singlet. Otherwise, it is easily seen that, whether S is integral or half-integral, if $L = S$, J takes on all the $2S + 1$ integrally spaced values from $L + S$ down to $L - S$, inclusive; whereas, if $L < S$, J takes on the $2L + 1$ integrally spaced values from $S + L$ to $S - L$.

The number of J levels composing an *LS* term is called the *multiplicity* of the term and is commonly written as a superscript at the

upper left-hand corner of the letter denoting the term. Thus ${}^2D$ denotes a doublet D term, with $L = 2$, $S = \frac{1}{2}$ and with two possible J values: $L + S = 5/2$, $L - S = 3/2$. To denote an individual level, the J value is added as a subscript. Thus a ${}^2D$ term contains the levels ${}^2D_{3/2}$, ${}^2D_{5/2}$. If $L = 2$, but $S = 1$, we have a triplet D term, with component levels ${}^3D_3$, ${}^3D_2$, ${}^3D_1$.

For convenience, even when $L < S$, the same multiplicity, $2S + 1$, is indicated in the notation, although the number of levels now equals $2L + 1$ and therefore is less than $2S + 1$. The reason for this practice is that, because of the selection rule for S , $\Delta S = 0$ (Sec. 137), the value of $2S + 1$ tends to remain the same in a radiative transition. The terms can thus be classified into more or less noncombining groups according to the values of $2S + 1$, and the superscript serves to indicate at once to which class a particular term belongs.

Because of these relations, the *observed* number of levels composing a given multiplet obviously furnishes important information in regard to the values of L or of S that should be assigned to the term. If the observed number of levels is r , this number must equal either $2L + 1$ or (more commonly) $2S + 1$, hence, either $L = (r - 1)/2$ or, more likely, $S = (r - 1)/2$. A further test may then be made by noting whether or not the value of L or S so inferred is consistent with the selection rules as applied to transitions between the given multiplet of levels and other multiplets for which L or S may be assumed to be known.

All allowed transitions between two multiplets of levels, taken together, give rise to a group of spectral lines which may be called a *spectral multiplet* of lines. The student should distinguish carefully between multiplets of levels and multiplets of lines. Often one can tell only from the context whether the word "triplet," for example, refers to three energy levels or to three spectral lines.

In Table VI are shown the levels composing some of the types of LS terms that are commonly met with. Only singlet, doublet, triplet, and

TABLE VI.—/ LEVELS COMPOSING SOME LS TERMS

$L =$	0	1	2	3
$S = 0$	${}^1S_0$	${}^1P_1$	${}^1D_2$	${}^1F_3$
$\frac{1}{2}$	${}^2S_{\frac{1}{2}}$	${}^2P_{\frac{1}{2}}$, ${}^2P_{\frac{3}{2}}$	${}^2D_{\frac{3}{2}}$, ${}^2D_{\frac{5}{2}}$	${}^2F_{\frac{5}{2}}$, ${}^2F_{\frac{7}{2}}$
1	${}^3S_1$	${}^3P_0$, ${}^3P_1$, ${}^3P_2$	${}^3D_1$, ${}^3D_2$, ${}^3D_3$	${}^3F_2$, ${}^3F_3$, ${}^3F_4$
$\frac{3}{2}$	${}^4S_{\frac{3}{2}}$	${}^4P_{\frac{1}{2}}$, ${}^4P_{\frac{3}{2}}$, ${}^4P_{\frac{5}{2}}$	${}^4D_{\frac{1}{2}}$, ${}^4D_{\frac{3}{2}}$, ${}^4D_{\frac{5}{2}}$, ${}^4D_{\frac{7}{2}}$	${}^4F_{\frac{3}{2}}$, ${}^4F_{\frac{5}{2}}$, ${}^4F_{\frac{7}{2}}$, ${}^4F_{\frac{9}{2}}$

quartet terms are shown, but quintet, sextet, septet, and octet terms also occur (for example, in the arc spectrum of manganese).¹

¹ Cf. WHITE, *op. cit.* Chap. XIV, 1934.

139. Spacing of the LS Multiplet Levels.—For the spacing of the levels in an *LS* multiplet, wave mechanics furnishes a simple and useful formula, which holds quite accurately so long as the entire width of the multiplet is small relative to its difference in energy from all other energy levels. The increase in the atomic energy due to the spin-orbit effect, for any one of the J levels of a given multiplet, is ΔW where

$$\Delta W = \frac{1}{2} B[J(J+1) - L(L+1) - S(S+1)]. \quad (211)$$

Here B is a constant that varies from one multiplet to another.

The relative spacing of the levels in a given multiplet is determined by the term $\frac{1}{2} BJ(J+1)$. The difference between the energies of a level for J and that for $J+1$ is the difference in the corresponding values of ΔW or

$$W_{J+1} - W_J = \frac{1}{2} B[(J+1)(J+2) - J(J+1)] = B(J+1). \quad (212)$$

This equation expresses Lande's interval rule: *The energy differences between two successive J levels are proportional, in a given LS term, to the larger of the two values of J .* The rule is of great help in determining the value of J that is to be assigned to the various levels.

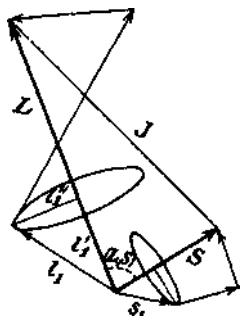
According to Eq. (211), ΔW is positive for some values of J and negative for others. The weighted average of ΔW , or ΔW for a weighted-average level, vanishes, provided each level is weighted in proportion to the number, equal to $2J+1$, of the M states composing it. The energy of the weighted-average level is then given by the formula

$$W_{l,s} = \frac{\sum_J (2J+1) W_J}{\sum_J (2J+1)}$$

It is to such a weighted-average level that a Rydberg formula really refers when it is written without regard to the fine structure of the terms. As an alternative, a separate Rydberg formula is sometimes written for each component of the multiplets forming a series.

Equation (211) can also be arrived at by the following argument based upon angular momenta, which is suggestive although not conclusive. According to expression (199) in Sec. 133, the spin-orbit energy for a classical electron is proportional to the cosine of the angle between its orbital momentum vector and its spin vector. Now, in a diagram illustrating *LS* coupling, such as that shown in Figs. 100 or 101, there exists an indefiniteness of azimuth similar to that described for 1 electron in Sec. 115(6); the angle between the l -vector for any particular electron and its s -vector is not fixed, since the composite I

line can be revolved around the L -vector without upsetting the coupling, and the composite s line can also be revolved independently around the S -vector. But the cosine of the angle between l_1 and s_1 for example, is proportional to the projection of h on s_1 ; and this projection is the sum of the projections on s_1 of the separate components l'_1, l''_1 , taken, respectively, parallel and perpendicular to L (see Fig. 103). Now, taken for all azimuths of the l -vectors around L , the average of l'_1 is zero. Hence, for a given position of s_1 the average projection of l_1 on s_1 is equal to the average projection of l_1 alone. Then, allowing the S -vector to revolve similarly around S , we may show that the average projection of l'_1 on s_1 is equal to its projection on S . Thus the average projection of l_1 on S is proportional to $\cos(L, S)$, where (L, S) denotes the angle between the L - and S -vectors; and, hence, $\cos(L, S)$ is also proportional to the average value of the cosine of the angle between l_1 and s_1 ; and so finally to the spin-orbit energy itself.



Now, by geometry,

$$\cos(L, S) = \frac{1}{LS}(L^2 + S^2 - J^2). \quad (214)$$

But in wave mechanics the magnitudes of the momenta are proportional, not to L , S , and J , but to $\sqrt{L(L+1)}$, $\sqrt{S(S+1)}$, and $\sqrt{J(J+1)}$. Making corresponding changes in (214), we find that the spin-orbit energy is proportional, for given L and S , to the expression given on the right in (211).

Many examples of LS coupling could be described. We have space in this book for only one or two.

140. The Arc Spectrum of Mercury.—The familiar arc spectrum of mercury presents spectroscopic features of great interest. The principal levels and many of the lines are shown in the usual way in Fig. 104, wave lengths being given in angstroms. The levels that are considered to form a sequence are placed under each other, with an appropriate spectroscopic designation at the top.

The levels fall into two classes, singlets and levels that naturally group themselves into triplets, although the separation is mostly too small to be shown in the diagram. Thus the spectrum evidently presents a fair example of LS coupling for 2 electrons. For the singlet terms $S = 0$; for the triplet terms $S = 1$. A number of intercombina-

tion lines, between singlet and triplet terms, are found, however, so that the *LS* coupling is not perfect in mercury.

Evidence from the theory of the periodic table indicates that mercury should have 2 electrons outside five completed shells. It is

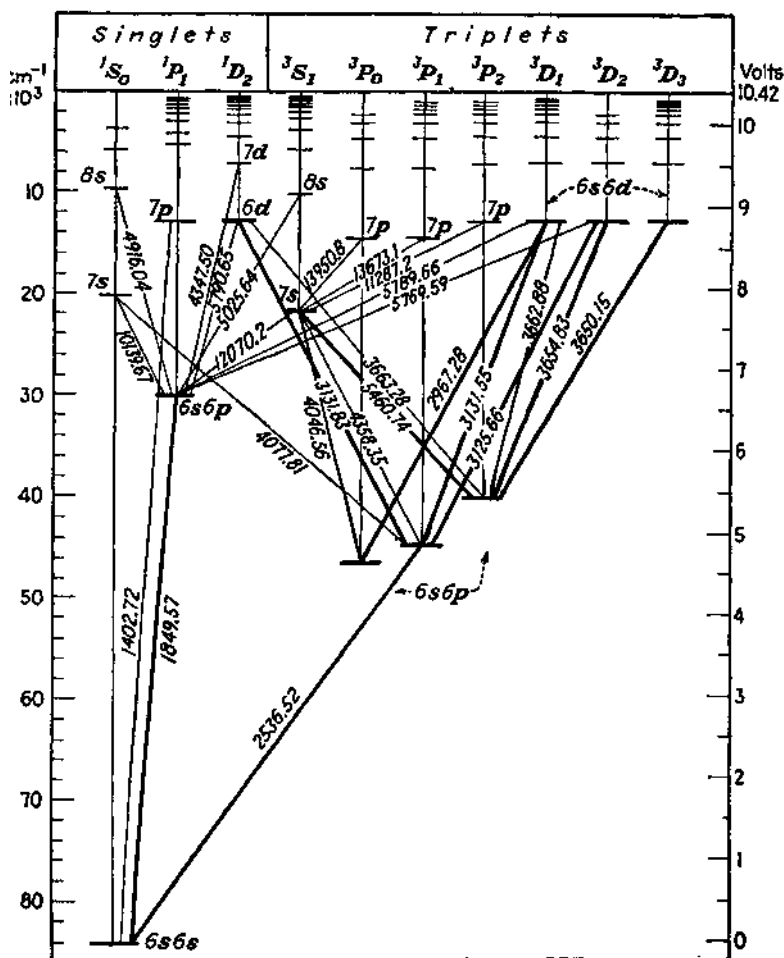


FIG. 104.—The most important energy levels and spectral lines for the neutral mercury atom.

assumed that these electrons normally form the closed subshell $6s^2$, but that in those excited atomic states which are represented in the diagram one of these electrons is displaced into an electronic state of higher energy. Since one electronic *I* is thus always 0, the value of *L* is always equal to that of the other *I*. Thus each electronic configura-

tion furnishes at most two LS terms, with $S = 0$ or 1 . The configuration in the normal state of the atom, $6s6s$, gives only one term, and, in fact, only one level, with $L = S = J = 0$; thus the normal state of the mercury atom is a $^1S_0$ state. In four of the most important terms, viz., $6s6p\ ^1P$, $6s6p\ ^3P$, $6s6d\ ^1D$, and $6s6d\ ^3D$, the mere change of an electronic l with no change in n results in a comparatively large increment of atomic energy.

The intercombination line, $6s6p\ ^3P_1 \rightarrow 6s^2\ ^1S_0$, $\lambda = 2,536.52$, is one of the principal ultraviolet lines in the mercury spectrum. Its strength, however, is not necessarily an indication of a high transition probability between these levels, arising from a high value of Einstein's A (Sec. 116). It is more likely that the strength of this line is due to a piling up of atoms in the $6s6p\ ^3P_1$ level. As the atoms are thrown into this state following collisions with electrons in the arc, or drop into it from higher levels, they have nowhere else to go, therefore they accumulate until through sheer numbers they are able to leave as fast as they enter. It is somewhat like water piling up in a pool until its depth becomes great enough to force a sufficiently rapid discharge from the pool through a small drain.

Atoms must accumulate also, of course, in the $^3P_2$ and $^3P_0$ levels of the $6s6p$ configuration. From these levels they cannot pass by a radiative jump to the normal state, for $^3P_2 \rightarrow ^1S_0$ would mean $\Delta J = 2$, and $^3P_0 \rightarrow ^1S_0$ would mean a jump from $J = 0$ to $J = 0$, both of which are forbidden by the selection rules for J . Levels lying above the normal state out of which radiative transitions are impossible are called "metastable" levels. An atom can stay in such a level for a long time if it is not disturbed by outside influences. It may eventually be brought back into the normal state, however, by a "collision of the second kind" [cf. Sec. 100(a), (c)], and it is presumably in this manner that the atoms in the $^3P_2$ and $^3P_0$ levels are eventually returned to their normal state.

Contrary to the selection rule for S , that for L is pretty well obeyed. With values of L assigned as in the diagram, S terms combine only with P terms, P with S and D , and so on. However, a few weak lines, not shown on the diagram, have been observed corresponding to $\Delta L = 2$.

All of the transitions shown in Fig. 104 are in harmony with the selection rule for configurations. Only 1 electron changes its n and l , and always $\Delta l = \pm 1$.

The spectral lines can be grouped into *series*, if desired. Thus, within the singlet system, all lines ending on the lowest $1s_0$ level form the singlet principal series. Only these and the line $\lambda = 2,536\text{ \AA}$

can be observed in absorption in mercury vapor. Of the lines ending on the lowest ^1P level, those originating from ^1S terms form a sharp series, those originating from ^1D terms, a diffuse series, just as in sodium; and so on. Similar series can be picked out within the triplet system. It is really not very interesting to group the lines of such a complex spectrum into series, however, especially when the "fine structure" is as coarse as it is in the mercury spectrum. Thus the great spectral sextet of ultraviolet lines, $X = 2,967$ to $X = 3,663$, from the lowest ^3D to the lowest ^3P term, would constitute together the first "line" of the triplet diffuse series.

The student will recognize the lines $^1D_2 \rightarrow 1P_1$ ($X\ 5,791$), $^3S_1 \rightarrow ^3P_2$ ($X\ 5,461$), and $^3S_1 \rightarrow ^3P_1$ ($X\ 4,358$) as the familiar yellow, green, and blue lines emitted from the mercury arc.

It may also happen that, when the atom is excited, both valence electrons are displaced into higher electronic states. A few levels ascribed to the $6p^2$ configuration have been discovered.

Among other neutral atoms which have spectra similar to that of mercury may be mentioned helium; then the alkaline earths, beryllium, magnesium, calcium, strontium, and barium; and the close relatives of mercury, zinc and cadmium. Ions with similar electronic exteriors are C^{++} , Al^+ , Si^{++} , Pb^{++} (the number of plus signs indicating the number of positive charges on the ion).

A second subclass of atoms with two optically active electrons is formed by those which, in their normal states, contain two s and two p valence electrons. In such cases the two s electrons usually (but not always) stay put, only the two p electrons being active. Examples of such atoms are neutral carbon, silicon, germanium, tin, and lead. Furthermore, certain observed spectra of the same type have been ascribed to singly (positively) ionized atoms of nitrogen, phosphorus, and bismuth.

141. Equivalent Electrons.—Hitherto, it has been assumed that all the electrons in a configuration have either different n 's or different V 's. When 2 electrons have the same n and also the same l , they are called by spectroscopists *equivalent electrons*. In configurations containing equivalent electrons, such as $5s^2$ or $5s7p^2$, certain LS terms that might otherwise occur are excluded through the operation of the exclusion principle.

Consider, for example, the simple configuration $1s^2$ in helium, or $6s^2$ in mercury. Four different combinations of the $n\ l\ m\ s$ electronic states can be made for the 2 electrons; these may be indicated, in an obvious notation, as follows:

$$(n00\frac{1}{2}, n00\frac{1}{2}), \quad (n00\frac{1}{2}, n00 - \frac{1}{2}), \quad (n00 - \frac{1}{2}, n00\frac{1}{2}), \\ (n00 - \frac{1}{2}, n00 - \frac{1}{2}).$$

where $n = 1$ for helium or $n = 6$ for mercury. (All these states have $I = 0$.) Since the total angular momentum is the sum of the orbital and spin momenta, the first and last of the four combinations suggest atomic states with $M = 1$ or -1 , and hence, with $J = 1$. But these particular combinations, in which both electrons are in the same electronic state, are ruled out by the exclusion principle. Furthermore, the other two combinations differ only in that the 2 electrons exchange quantum states, and according to another aspect of the exclusion principle, as stated in Sees. 118 and 120(6), it makes no difference which electron is in which state; thus each combination of $n l \lambda /u$'s, regardless of the order of these numbers, furnishes just one quantum state for the atom as a whole. Hence we obtain out of the configuration under discussion a single atomic state, represented in wave mechanics by a single wave function ψ . This must be a $1S$ state, with $L=S = M = J = 0$. The ^3S set of three states, with $M = 1, 0$, or -1 , is missing.

The same kind of reduction occurs for any closed subshell. When so many electrons are present with given values of n and of I that, in assigning them to different electronic states, every allowed value of X and j must be used, only a single combination of electronic states is possible, and it can lead to only a single atomic state. This is necessarily a state with $J = 0$, since any other value of J requires the existence of several atomic states with different values of M . Thus atoms whose electronic exterior is formed of closed subshells are necessarily in a ^1S state. This conclusion is confirmed by the spectroscopic evidence, as a check of the table in Appendix III will show.

In the simple example just discussed, the reduction in the number of states happens to be in accord with the predictions of the vector diagram for LS coupling. In more complicated cases, some of the LS terms that are predicted by the diagram are themselves missing. General rules can be given for determining which states are allowed for any given configuration containing identical electrons.¹ All the J levels belonging to a given LS term always appear or drop out together.

142. "jj" Coupling.—The validity of LS coupling is limited to atoms that are not too heavy. As the atomic number increases, the spin-orbit effects become rapidly larger; as a consequence, the J levels tend usually less and less to group themselves into LS multiplets, and the selection rules for L and S fail more and more. Finally, in very heavy atoms the spin-orbit effects tend to predominate over the residual electrostatic effect (*i.e.*, over the part of the electrostatic effect

¹ WHITE, C. F., "Introduction to Atomic Spectra," 1934.

that is not included in the central field). Then an approximation to another type of "coupling," known as "jj coupling," occurs.

This term has reference to an alternative path of approach in applying the perturbation theory. Let us ignore at first, not the spin-orbit effects, but the *electrostatic repulsion* of the electrons, except in so far as this repulsion is allowed for in the modified form of the central field. Let full account be taken of the spin-orbit effects. Since the spin magnetic moment of each electron reacts with its own orbital motion, independently of the other electrons, each electron can be assumed to occupy one of the $n\ l\ j\ m$ electronic states that were described in Sec. 133. In each of these states there is a definite value of $\sqrt{j(j+1)}\ \hbar/2\pi$ for the total angular momentum of the electron, and a value $m\hbar/2\pi$ for its component in the direction of a chosen axis.

A quantum state for the whole atom is then specified by assuming a set of the $n\ l\ j\ m$ quantum numbers for each of the N electrons in the atom:

$$n_1\ l_1\ j_1\ m_1\ n_2\ l_2\ j_2\ m_2, \cdot \cdot \cdot\ n_N\ l_N\ j_N\ m_N. \quad (215)$$

Electrons occupying closed shells may, however, be ignored provided these shells remain closed in all of the radiative transitions that are under consideration. In developing the wave-mechanical theory, a wave function for the atom is then constructed out of the N corresponding electronic wave functions, in such a manner that it makes no difference which electron is assigned to each of the electronic states.

The energy of the atom, however, is independent of the m 's and depends, like the individual electronic energies, only upon the electronic quantum numbers

$$n_1\ l_1\ j_1, n_2\ l_2\ j_2, \cdot \cdot \cdot\ n_N\ l_N\ j_N. \quad (216)$$

Let us choose for consideration only atomic states belonging to a particular configuration as specified by the still more limited array of electronic quantum numbers

$$n_1\ l_1, n_2\ l_2, \cdot \cdot \cdot\ n_N\ l_N.$$

Only the j 's will vary among these states; hence, in discussing a single configuration, only the j 's need be written down. The order in which the j 's are written is important, however, except when two j 's refer to equivalent electrons; for each j is associated in general with different values of n or l . The atomic energy will then depend upon the values of the j 's. Thus the atomic states composing the chosen configuration fall into groups, each characterized by a certain sequence of the j 's and by a certain value of the energy. The energy level or term associ-

ated with each set of j 's may be called a jj level, or term. The number of jj levels belonging to a given configuration cannot exceed 2^N ; for each j can assume at most 2 values, $I + \frac{1}{2}$ and $I - \frac{1}{2}$. This form of electronic coupling is known as " jj coupling"; the jj terms correspond to the LS terms in LS coupling.

In addition to the selection rule for configurations, there is now a selection rule for j : $\Delta j = 0$ or ± 1 . The rules for jj coupling can thus be summarized in the following form:

1. Only one $n I j$ set of quantum numbers can change in a radiative transition—"only 1 electron jumps at a time."

2. For the jumping electron,

$$\Delta l = \pm 1, \quad \Delta j = 0 \text{ or } \pm 1.$$

Each jj level contains numerous atomic states corresponding to different possible choices for the electronic m 's. These degenerate

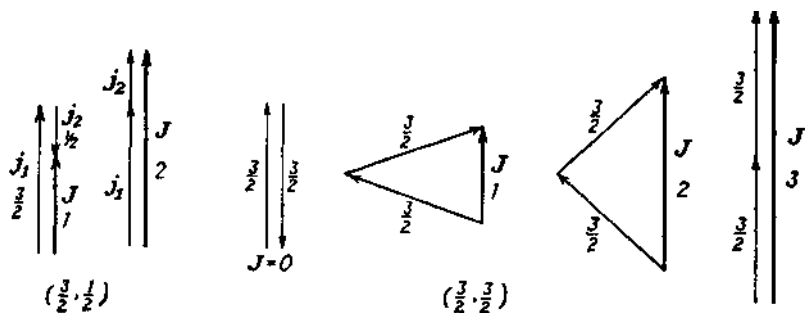


FIG. 105.—Diagrams illustrating jj coupling for two nonequivalent electrons.

states can now be reconstructed so as to quantize, not the individual electronic momentum components, but the total angular momentum of the atom, in harmony with the general scheme described in Sec. 130. In wave mechanics this is done by starting with atomic wave functions based on the $nljm$ electronic functions and making suitable linear combinations of these. The new atomic states are characterized by quantum numbers J and M referring to the total angular momentum; but the energy, as usual, is independent of M .

For a given jj level the possible values of J , which must be integrally spaced, may be ascertained by constructing in the usual way a vector diagram representing the addition of the various electronic j -vectors into a resultant J -vector, in analogy with the classical vector addition of the resultant angular momenta of the electrons to form the total vector angular momentum of the atom. Two cases, for a group of 2 electrons, are illustrated in Fig. 105.

In addition to the selection rules just stated, the usual selection rules then hold for J and M .

In this way we obtain atomic energy levels that are characterized by (1) a set of pairs of values of n and l , specifying a configuration, (2) a sequence of values of the j 's specifying a jj level or term within the configuration, and (3) a value of J .

So long as the residual electrostatic repulsion is omitted, all J levels belonging to the same jj term correspond to the same atomic energy. When the full effects of electrostatic interaction are taken into account, the J levels composing a jj term become separated in energy, forming a jj multiplet. If the residual electrostatic repulsion is actually strong, it tends also to destroy the significance of the electronic quantum numbers $n l j$, and the associated selection rules tend to fail. The quantum numbers J and M , however, and their selection rules, persist under all conditions, so long as the atom is free from external force.

A notation for the final energy levels in jj coupling is easily invented. Thus, analogous designations of a J level with $J = 3$ might be, in two alternative forms of coupling,

$$\begin{array}{cc} jj & LS \\ (n_1 l_1 \cdots n_N l_N)(j_1 \cdots j_N)_{J'} & N_1 l_1 \cdots n_N l_N {}^5D_3. \end{array}$$

An interesting example exhibiting the transition from LS to jj coupling is shown in Fig. 106. The relative positions of certain levels are shown for carbon ($z = 6$), germanium (32), and lead (82), corresponding levels being connected by dotted lines; and all transitions that are observed to occur between these levels in carbon and in lead are indicated by arrows. In silicon (14), the corresponding set of levels is observed to be arranged much as in carbon, whereas tin (50) resembles lead in this respect. The number of closed subshells underlying the active electrons is different in each case, but this difference is immaterial to the relative arrangement of the levels.

In Fig. 106, we note that the J value of any level remains the same in all three spectra; but in carbon the levels group themselves by their energy values into good LS multiplets, whereas in lead they form jj groups, for which the j values are given in parentheses. The observed radiative transitions as indicated in the figure for carbon and for lead include all that are allowed by the selection rules. The student will find it instructive to verify this statement in detail and to note which lines occur with one form of coupling and not with the other.

Many other examples of the same sort could be given.

For further information concerning complex spectra, the student is referred to special treatises on the subject.¹

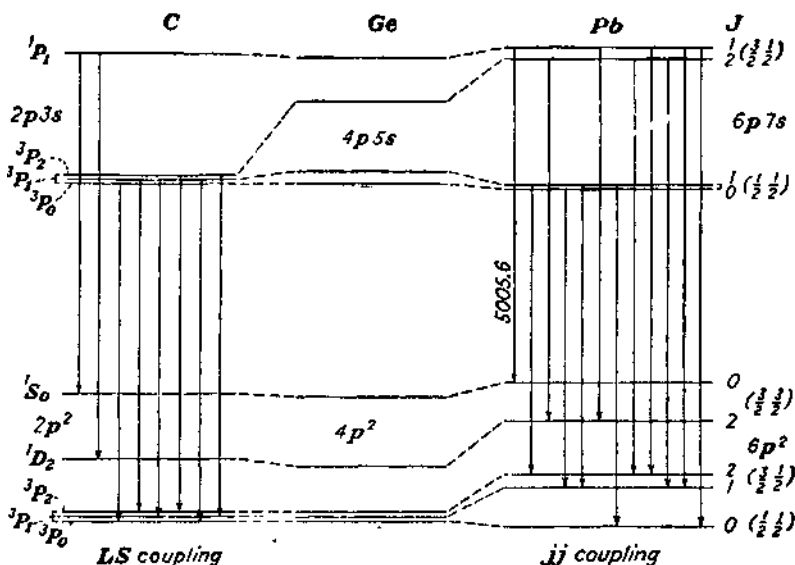


FIG. 106.—A transition from *LS* to *JJ* coupling. Within each multiplet of levels the spacing is drawn to scale, except that levels joined by a brace should be drawn closer together. The ranges covered by the levels, in cm^{-1} , are: carbon, $2p^2$, 69,231.1 to 90,878.3; $2p_3s$, 28,898 to 30,547; germanium, total range (term values uncertain), $4p^2$, 16,367; $4p_5s$, 2,568; lead, $6p^2$, 30,365 to 59,821; $6p_7s$, 10,383 to 24,863. One wave length is given in angstroms; the other lines lie in the ultraviolet.

143. The Effect of a Magnetic Field on an Atom.—When an atom is subjected to the action of *external forces*, its quantum states are modified. States which were formerly degenerate, corresponding to the same energy, may be separated, so that the degeneracy is reduced. The spectral lines emitted by the atom may consequently be split into several components.

The most interesting case of this sort is the *Zeeman effect*. In Sec. 40, Chap. III, we discussed it in terms of classical theory, and we remarked that in some cases the predictions of classical theory are in agreement with experiment. Such cases are sometimes referred to as the *normal Zeeman effect*. More commonly, however, the observed pattern of lines is quite different from that predicted by classical theory; this is called the *anomalous Zeeman effect*. Some anomalous patterns are illustrated in Fig. 107. Such patterns are fascinating in their variety.

¹ WHITE, *op. cit.*; HERZBERG, G., "Atomic Spectra and Atomic Structure," 1937.

No explanation of the anomalous effect was found until electron spin was introduced. Furthermore, in the modern wave-mechanical treatment of the Zeeman effect, there is, naturally, nothing closely resembling the electronic motion as described in the classical theory. Nevertheless, certain features of the classical picture still retain an interest, either because they are actually preserved in the new theory, or because they furnish a useful picture in terms of which the phenomenon can readily be visualized.

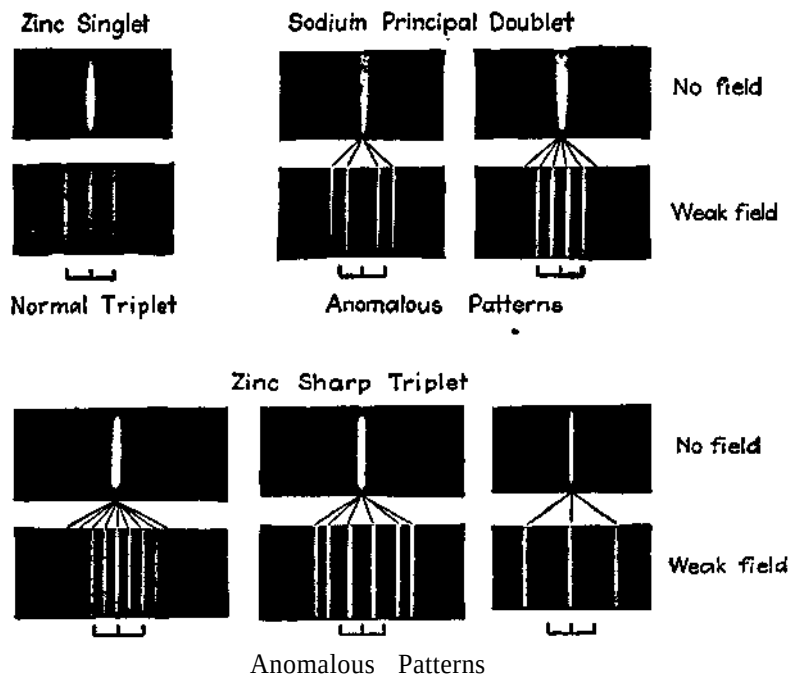


FIG. 107.—Photographs of some Zeeman patterns in a weak magnetic field, viewed perpendicularly to the field. The brackets show in each case the position of a normal triplet in the same field. {From H. E. White, "Introduction to Atomic Spectra."}

In the new theory, as in the old, the changes produced in the energy levels by a magnetic field can be regarded as resulting from the possession by the atom of a *magnetic moment*. Classical theory indicates that there should be a close relation between the magnetic moment due to motion of an electron in an orbit and the angular momentum resulting from the same motion. It will be instructive to work out this relation.

Suppose a particle carrying algebraic charge q revolves in a fixed circular orbit of radius a at a velocity v (Fig. 108). Then it makes

$v/(2\pi a)$ revolutions per second and so is equivalent to a current I flowing around the same circle of magnitude

$$I = \frac{qv}{2\pi a}.$$

It is well known that a current I , flowing in any plane loop enclosing area A , is equivalent to a magnet placed with its axis perpendicular to the loop and having a magnetic moment

$$\mu = \frac{IA}{c}.$$

(Here q and I are in electrostatic units, hence the factor $1/c$). In the present case $A = \pi a^2$, therefore,

$$\mu = \frac{I\pi a^2}{c} = \frac{qva}{2c}.$$

The moment (μ , can be treated as a vector perpendicular to the plane of the orbit; let us write for it μ_l .

The particle will also have angular momentum about the center of the circle of magnitude

$$G_l = mva,$$

m being its mass. As a vector, the angular momentum, G_l , is likewise perpendicular to the orbit.

Comparing these two equations, we see that *the magnetic moment and angular momentum are related by the vector equation*

$$\mu_l = \frac{qG_l}{2mc}. \quad (217)$$

This result, which is independent of a , can be shown to hold for motion in an orbit of any shape. If the particle is an *electron* with numerical charge e , we have $q = -e$ and

$$\mu_l = -\frac{eG_l}{2mc}. \quad (218)$$

Thus, for an electron, the vectors μ_l and G_l are oppositely directed.

A similar expression must then be added for the effect of *electron spin*. According to classical electromagnetic theory, a sphere, uniformly charged with electricity on its surface and set spinning, behaves as if it were covered with electric currents and possessed a correspond-

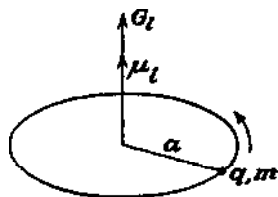


FIG. 108.—Diagram illustrating the angular momentum G_l and magnetic moment μ_l due to a charge revolving in a circle.

ing magnetic moment; and, because of the momentum in the electromagnetic field around it, it also possesses angular momentum (Sec. 30). For the relation between its vector magnetic moment $\mathbf{u}_s$ and its vector angular momentum $\mathbf{G}_s$ classical theory gives

$$\mathbf{u}_s \approx \frac{q\mathbf{G}_s}{mc}, \quad (219)$$

where q/c is the algebraic charge on the sphere in electromagnetic units and m is its electromagnetic rest mass, due to the inertia of its electric field (*cf.* Sec. 36). For an electron this becomes

$$\mathbf{u}_s = -\frac{e}{mc}\mathbf{G}_s. \quad (220)$$

Comparing Eqs. (218) and (220), we note that relatively twice as great a magnetic moment is associated with angular momentum of spin as with orbital angular momentum.

Now, there is no reason to suppose that the vectors $\mathbf{u}_l$ and $\mathbf{u}_s$ would necessarily have fixed directions for an electron in an atom. If a magnetic field is present, however, the components of the orbital and spin momenta in the direction of the field will have certain average values, which we may denote by G_{lH} and G_{sH} , respectively. According to Eqs. (218) and (220), this average momentum will be accompanied by an average magnetic moment in the direction of the field of magnitude

$$\bar{\mu} = -\frac{e}{2mc}(\bar{G}_{lH} + 2\bar{G}_{sH}) \quad (221)$$

(in which e stands for the *numerical* electronic charge). Or, if the atom contains more than 1 electron, its magnetic moment in the direction of the field will be

$$\bar{\mu} = -\frac{e}{2mc} \sum (\bar{G}_{lH} + 2\bar{G}_{sH}),$$

the Σ indicating a summation over all of the electrons in the atom.

These same expressions for $\mathbf{u}$ are given by wave mechanics. The occurrence of a magnetic moment associated with spin follows from the Dirac relativistic equation in much the same way as does the angular momentum of spin (Sec. 117). The quantum-mechanical definitions of the averages are

$$\bar{G}_{lH} = \int \psi^* G_{lH} \psi dq, \quad \bar{G}_{sH} = \int \psi^* G_{sH} \psi dq,$$

G_{lH} and G_{sH} in the integrals standing for certain operators and ψ^* for the complex conjugate of the wave function ψ ; the integrations are to

be extended over all values of the coordinates of all electrons in the atom.

In passing, it may be remarked that in Eq. (222) we have the wave-mechanical basis for the treatment of paramagnetism and diamagnetism. *Paramagnetism* is ascribed to a tendency of an atom to occur more frequently in those quantum states in which it has a magnetic moment in the direction of the magnetic field than in states in which the moment is opposed to the field. *Diamagnetism* results from a slight modification of the properties of the quantum states themselves, by which a slight additional magnetic moment is produced, always in a direction opposed to the field, in accordance with Lenz's law. *Ferromagnetism* is a more complicated phenomenon, believed to be due to a lining up of electronic spins in *different atoms* under special conditions.

Here, however, we are interested chiefly in effects upon the *atomic energy levels*. In classical theory, a magnet of moment M in a uniform magnetic field H behaves as if it possessed a potential energy of magnitude

$$-MH \cos (M, H),$$

the symbol (M, H) standing for the angle between the directions of M and H . The product $M \cos (M, H)$ can also be regarded as representing the component of the moment in the direction of the field. Similarly, wave mechanics indicates that the effect of the magnetic field upon an atomic energy level is to change the energy by the amount

$$W_H = -\mu H = \frac{eH}{2mc} \sum (\tilde{G}_{1H} + 2\tilde{G}_{3H}), \quad (223)$$

by Eq. (222). If the field H is not too strong, a sufficiently accurate value of WH can be obtained by using the values of G_{1H} and G_{3H} that belong to the various quantum states in the absence of a field.

The ideas and equations thus developed are adequate for an understanding and visualization of the theory of the Zeeman effect. It may be instructive, however, to investigate a little further the nature of the magnetic energy WH due to the classical motion of an electron in an orbit. The student who prefers to do so can omit the next section entirely without embarrassment in reading the rest of the book.

144. Classical Theory of the Magnetic Energy.—The change in the energy of a revolving electron due to the presence of a magnetic field cannot be of the nature of an ordinary *potential* energy, for the force exerted by the magnetic field upon the electron acts always perpendicularly to its velocity and so never does any work upon it.

If, therefore, owing to some change in conditions, the energy decreases, this decrease must occur in the kinetic energy of the electron. (Actually there is also a reaction on the magnetizing system itself, but further study shows that this can be left out of consideration.)

To take a definite case, suppose an electron is revolving in a circular orbit of fixed radius a , under the action of a force directed toward a fixed center (*cf.* Fig. 108). Let a magnetic field of H oersteds, perpendicular to the plane of the orbit, be created very slowly. If there were a conducting wire lying along the orbit, there would be an induced electromotive force in this wire of magnitude

$$\epsilon = -\pi a^2 \frac{dH}{dt}.$$

According to the elementary definition of e.m.f., this means that work of amount $-\epsilon e/c$ will be done on the electron every time it goes round the orbit, $-e/c$ being its charge in electromagnetic units. This work must be done by a force acting on the electron. Hence, if F is the average component of this force in the direction of the motion, we have for the work done in a revolution

$$\begin{aligned} \bar{F} \times 2\pi a &= -\frac{\epsilon e}{c} = \frac{e}{c} \pi a^2 \frac{dH}{dt}. \\ \therefore \bar{F} &= \frac{ea}{2c} \frac{dH}{dt}. \end{aligned}$$

The average moment of force about the center, on the other hand, is aF ; and this must be equal to the rate of change of the angular momentum G_{IH} about an axis parallel to the field. Thus

$$\frac{dG_{IH}}{dt} = \frac{ea^2}{2c} \frac{dH}{dt};$$

and, integrating, we have for the whole angular momentum imparted to the electron during the creation of the field

$$\Delta G_{IH} = \frac{ea^2}{2c} H. \quad (224)$$

Two cases may now be distinguished.

If in the beginning the electron had no angular momentum, and hence also no orbital magnetic moment, it acquires angular momentum of magnitude ΔG_{IH} as given by Eq. (224), and also, by Eq. (218), a magnetic moment

$$\Delta \mu_H = -\frac{e}{2mc} \Delta G_{IH} = -\frac{e^2 a^2}{4mc^2} H.$$

A similar expression is obtained from wave mechanics for the magnetic moment developed in an atom, when it is in a given quantum state, by the application of a magnetic field:

$$\Delta\mu_H = -\frac{e^2 H}{4\pi mc^2} \sum R^2, \quad (225)$$

where R^2 is the average of the square of the distance of any electron from an axis drawn through the nucleus parallel to the field, calculated as are the quantities GIH and $G_\theta H$ that occur in Eq. (221), and 2 indicates summation for all electrons in the atom. This added magnetic moment, necessarily opposite in direction to the field, gives rise to the phenomenon of diamagnetism.

The electron is also given a certain amount of kinetic energy, but this is proportional to H^2 and hence is easily shown to be negligible, in the classical analog of actual atomic cases.

If, on the other hand, the electron is already revolving, then, although the same changes in momentum and moment occur, they are in actual cases small as compared with the initial values. A change in the *kinetic energy* now occurs, however, that is of the *first order* in H . Let us denote the angular momentum about the direction of H by Gl , taking it to be positive, as is ΔG_l , when the associated rotation is clockwise to an observer looking in the direction of the magnetic field. Then, since $|G_l| = mva$, the total kinetic energy of the electron is

$$K = \frac{1}{2}mv^2 = \frac{G_l^2}{2ma^2}.$$

Thus, when the angular momentum is changed from GIH to $GIH + \Delta GIH$, GIH now standing for the initial momentum, the kinetic energy is increased by

$$\Delta K = \frac{1}{2ma^2} [(G_{lH} + \Delta G_{lH})^2 - G_{lH}^2] = \frac{1}{ma^2} \left[G_{lH} \Delta G_{lH} + \frac{1}{2} (\Delta G_{lH})^2 \right].$$

Here the last term is of the second order in H and is negligible. Hence, using (224), to the first order in //

$$\Delta K = \frac{e}{2mc} G_{lH} H.$$

ΔK may be positive or negative according to the sign of GIH .

Comparing the value just found for ΔK with the first term on the right in Eq. (223) with the Σ omitted (for one electron), which represents the change in energy due to the field and may be denoted by

WIH, we see that $W_{IH} = \Delta K$. Thus W_{IH} is accounted for in classical theory by the change that occurs in the kinetic energy of the revolving electron during the introduction of the magnetic field.

In obtaining this result, we supposed the plane of the orbit to remain fixed in position while the field varied. The same results are obtained, however, if, instead of creating the field, we rotate the plane of the orbit from a position parallel to the field into a position perpendicular to it. Furthermore, the results are readily shown to hold for an orbit of any shape.

In developing the older quantum theory of the Zeeman effect, the method just described for finding W_{IH} was especially advantageous because it obviated all doubt as to the proper conditions for determining the allowed quantum states in the presence of a magnetic field. For it was accepted as a general principle that, if a system is in a certain quantum state to begin with and if the external conditions are changed very slowly, the system must remain in an allowed quantum state; whatever state it passes into must, therefore, be an allowed state under the new conditions. This principle, known as Ehrenfest's *adiabatic principle*, holds in wave mechanics as well.

The Bohr quantum states for a 1-electron atom subject to a magnetic field are more simply described, however, from a distinctly different point of view, which lies much closer to the classical picture of the Zeeman effect. For this purpose we make use of a celebrated theorem due to Larmor, which for our purposes may be stated in the following form: Given a system of electrons describing orbits, however complicated, around a nucleus located at the origin of a set of coordinate axes to which the motion is referred. The behavior of the system when subjected to a magnetic field H may be described by saying that, neglecting small quantities of second order and higher, the electrons retain their original motions with respect to the coordinates, but by the action of the field the whole system, electrons and coordinate axes, is given a rotation about an axis parallel to the field and passing through the nucleus, the angular velocity of the rotation being¹

$$\omega_L = \frac{eH}{2mc}$$

(e = numerical charge and m = mass of an electron, c = speed of light). This rotation is called the *Larmor precession*.

¹ LARMOR, "Æther and Matter," p. 341, Cambridge University Press, London, 1900. See also RUARK and UREY, "Atoms, Molecules and Quanta," p. 759; JEANS, "Electricity and Magnetism," p. 565.

In view of this theorem we should expect to find the electron executing the same Bohr orbits relative to a frame of reference undergoing the Larmor precession as it does relative to a fixed frame in the absence of the field. It is easily shown that this assumption leads to the correct value for the energy.

The Larmor precession has no effect upon the potential energy of the electron. To find the resulting change in the kinetic energy, let us resolve the velocity of the electron at any instant into three rectangular components, of which one, v_1 is perpendicular both to the magnetic field and to the radius drawn from the nucleus to the electron. Only this component of the velocity is affected by the precession; and it is also only this component that contributes to G_{IH} , the angular momentum about the direction of the field. The precession adds to v_1 a component of velocity R_{wL} R being the perpendicular distance of the electron from a line drawn through the nucleus parallel to H . The change in kinetic energy is, therefore,

$$\Delta K = \frac{1}{2}m[(v_1 + R\omega_L)^2 - v_1^2] = m(v_1 R\omega_L + \frac{1}{2}R^2\omega_L^2).$$

But

$$G_{IH} = mv_1 R, \quad \therefore \Delta G_{IH} = mR \Delta v_1 = mR^2\omega_L.$$

Hence,

$$\Delta K = \frac{1}{mR^2} \left[G_{IH} \Delta G_{IH} + \frac{1}{2} (\Delta G_{IH})^2 \right].$$

This is obviously a generalized form of the first expression obtained above for ΔK , which was shown to represent the effect of a magnetic field upon the energy.

We are now fully prepared for a discussion of the Zeeman effect on the basis of quantum theory.

145. Zeeman Effect in a Huge Field.—The simplest type of Zeeman effect should be produced in a field so strong that complications due to all other sources can be ignored. Such a field we shall call "huge."

Consider first 1 electron in a central field, of Coulomb type or otherwise, upon which is superposed a uniform magnetic field H . Let us take an axis parallel to the magnetic field in constructing electronic quantum states characterized by the quantum numbers $n \mid n$ (Sec. 118). When the electron is in one of these quantum states, its component of orbital angular momentum about the axis, G_{IH} , and its component of spin momentum, G_{sH} , have respective magnitudes

$$G_{IH} = \frac{\lambda \hbar}{2\pi}, \quad G_{sH} = \frac{\mu \hbar}{2\pi}.$$

Substitution of these values for G_{IH} and G_{sH} in Eq. (223), in which we

omit the summation sign because we are dealing here with a 1-electron atom, gives the change in the energy of the electron due to the presence of the magnetic field. Adding W_0 , the energy when the field is absent, we thus obtain for the total energy

$$W = W_0 + \frac{eh}{4\pi mc} (\lambda + 2\mu) H \quad (227)$$

(e being here the numerical electronic charge).

The coefficient occurring in this formula,

$$\frac{eh}{4\pi mc} \quad 0.925 \times 10^{-20} \text{ electromagnetic c.g.s. unit,}$$

represents a natural unit of magnetic moment and is called the *Bohr magneton*. It is equal to the classical magnetic moment due to an electron revolving in the smallest Bohr orbit for hydrogen, with angular

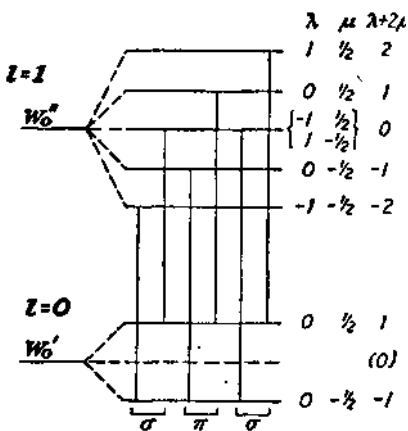


FIG. 109.—Energy levels for 1 electron in a huge magnetic field. (The levels for $l = 0$ are atypical in that the middle one is missing.) Transitions giving rise to the same ν are bracketed.

momentum $\hbar/2\pi$. The magnetic moment of an electron due to spin is also a whole Bohr magneton; for, although the angular momentum of spin is only half a unit or $\hbar/4\pi$, the magnetic moment, as we have seen in Sec. 143, is relatively twice as large. An electron in the quantum state number $n\lambda/x$ has thus a magnetic moment about the axis of $(\lambda + 2\mu)$ Bohr magnetons. Here $\lambda + 2\mu$ is always integral since $\mu = \pm 1/2$ and λ is an integer or zero.

From Eq. (227) we see that the magnetic field splits the *single* original level with energy W_0 into *several magnetic levels*, each characterized by a value of $\lambda + 2\mu$. An example is illustrated in Fig. 109. Since $\lambda + 2\mu$ will obviously take on all integral values from $I + 1$, when $\lambda = I$ and $\mu = 1/2$, down to $-(I + 1)$, when $\lambda = -I$ and $\mu = -1/2$ there are in general $2I + 3$ magnetic levels in all; if, however, $I = 0$, there are only two levels, with $\lambda + 2\mu = \pm 1$. All of the levels except the upper two and the lower two are double. The levels are *equally spaced*, the difference in energy between successive ones being

$$\frac{eh}{4\pi mc} H. \quad (228)$$

The cause of the separation of the levels can be visualized, if desired, with the help of a vector diagram similar to that used for the angular momentum. Let a vector of length I be drawn from a fixed point in such a direction that its component along a line parallel to the field is λ ; from another fixed point let another vector of unit length be drawn either parallel to the field or opposed to it (Fig. 110). Each pair of positions of these vectors then represents one of the quantum states characterized by λ and μ . The vectors can be regarded as representing magnetic moments due to orbital motion and to spin, of respective magnitudes I and 1 Bohr magnetons, the directions of these moments being opposite to those of the vectors because of the negative charge on the electron.

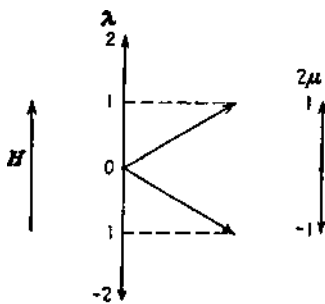


FIG. 110.

Consider now *transitions* between two such sets of energy levels. We may write Eq. (227) for the two sets thus:

$$W' = W'_0 + \frac{en}{4\pi mc} (\lambda' + 2\mu') H,$$

$$W'' = W''_0 + \frac{eh}{4\pi mc} (\lambda'' + 2\mu'') H.$$

As selection rules we have, besides $\Delta l = \pm 1$,

$$\Delta\lambda = 0 \text{ or } \pm 1, \quad \Delta\mu = 0.$$

For comparison with other cases we may note that these rules require also that

$$\Delta m = 0 \text{ or } \pm 1, \quad (229)$$

where $m = \lambda + \mu$ and represents, in units of $h/2\pi$, the total component of angular momentum in the direction of the field.

Calculating frequencies as $(W - W'')/h$ and writing

$$\frac{(W'_0 - W''_0)}{h} = \nu_0$$

for the frequency of the line emitted in zero field, we find just three possible lines when the field is present:

$$\lambda'' - \lambda' = \Delta\lambda = 0: \quad \nu = \nu_0 \quad (\pi)$$

$$\Delta\lambda = -1: \quad \nu = \nu_0 + \frac{e}{4\pi mc} H \quad (\sigma)$$

$$\Delta\lambda = 1: \quad \nu = \nu_0 - \frac{e}{4\pi mc} H. \quad (\sigma)$$

Each of these lines results, in general, from several different jumps, so that it might be described as consisting of several lines superposed. An example is illustrated in Fig. 109, already mentioned. The student will find it instructive to construct a similar diagram for transitions between levels with $l = 2$ and $l = 1$.

Comparing these results with those obtained from classical theory in Sec. 40, we see that we have here exactly the *classical triplet*. According to further conclusions from wave mechanics, even the polarization features should be the same as those deduced from classical theory. The undisplaced line, corresponding to $\Delta m = 0$, should be polarized with the electric vector in a plane parallel to the magnetic field, and it should be invisible when the radiation is viewed longitudinally (i.e., in a direction parallel to the field); such lines are called π lines (i.e., "parallel" as to polarization) or, sometimes, p lines. The other two lines, when the radiation is viewed transversely, should be plane polarized with the electric vector perpendicular to the magnetic field, whereas, when viewed longitudinally, they should be circularly polarized with opposite directions of rotation, as in the classical case; such lines are called σ lines (from the German "senkrecht" = perpendicular) or, sometimes, s lines. A Zeeman triplet agreeing thus in all respects, even in spacing, with the classical theory is called a *normal triplet*. An example is shown in Fig. 107.

^y Concerning the *relative intensity* of the lines, wave mechanics makes the following predictions. Suppose all of the initial quantum states are excited equally, i.e., equal numbers of atoms are thrown into each of these states per second by the process of excitation. Under these conditions, the two σ lines will each be half as bright as the central π line, as seen transversely to the magnetic field; whereas, the σ lines will be twice as bright when viewed along the field as they are when viewed transversely. This latter result is easily seen to be in agreement with the classical theory of the Zeeman effect; each circular motion is equivalent to two linear vibrations at right angles to each other, but, in the transverse direction, radiation is received from only one of these, the one that lies in the line of sight being invisible. Furthermore, as much light is polarized in one way as in the opposite way, so that on the whole the emitted radiation is unpolarized; and the total intensity is equal in all directions.

The quantity $eH/(4\pi mc)$, representing the separation in frequency of each outer line from the central line in a normal triplet, is often called a *Lorentz unit*, symbol L . Measured in waves per cm., its value is .

$$\tilde{L} = \frac{e}{4\pi mc^2} H = 4.67 \times 10^{-5} H \text{ cm.}^{-1} \quad (230)$$

Thus, the normal separation is $4.67 \times 10^{-5} \text{ cm.}^{-1}$ per oersted. We note that L does not contain Planck's constant h . It is because of this circumstance that classical theory gave correct results for certain cases of the Zeeman effect.

Up to this point, we have completely neglected the spin-orbit interaction. The effect of this interaction, when the magnetic field is very intense, is to shift the levels somewhat, in such a way as to separate slightly in energy those quantum states which have the same value of $\lambda + 2\mu$ but different values of $m = \lambda + \mu$. For example, states with $\lambda = 1$ and $u = -\frac{1}{2}$ and with $\lambda = -1$ and $u = \frac{1}{2}$ both have $X + 2U = 0$ and so fall together if the spin-orbit effect is omitted entirely; but $m = \frac{1}{2}$ for the first and $m = -\frac{1}{2}$ for the second, so that the energy levels for these two states are in reality slightly separated because of the spin-orbit effect. The number of lines in the Zeeman pattern is thereby increased, but, so long as the field is sufficiently intense, the lines fall into three groups in the approximate positions of the normal triplet.

Experimental confirmation of these conclusions from the theory is not extensive for 1-electron atoms, because very intense fields are required. However, the first three Balmer lines of hydrogen were observed as approximate normal triplets by Paschen and Back, in a field of 26,900 oersteds.¹ The fine structure due to the spin-orbit effect was not resolved, but it was presumably responsible for the broadness of the observed lines.

On an atom containing more than 1 electron, the effect of a huge magnetic field is to split each energy level W_0 for zero field according to the more general equation

$$W = W_0 + \frac{en}{4\pi mc} [\Sigma(\lambda + 2\mu)] H, \quad (231)$$

which constitutes an obvious generalization of Eq. (227); here the summation extends over all electrons in the atom. The normal triplet occurs as with 1 electron, since only 1 electron can jump at a time. The selection rules now include the rule that

$$\Delta M = 0 \text{ or } \pm 1, \quad (232)$$

where $M = \Sigma(\lambda + \mu)$ and denotes the total component of angular momentum in the direction of the field.

To obtain a good approximation to this type of Zeeman pattern with an atom containing more than 1 electron, the magnetic field must be so strong that it swamps all other influences except those for which allowance is made in the central field. The necessary

¹ PASCHEN and BACK, *Ann. d. Physik*, vol. 39, p. 897 (1912).

strength runs to millions or even hundreds of millions of oersteds, so that Zeeman patterns of this type cannot be observed at present.

146. Zeeman Effect in a Weak Field.—The Zeeman effect is comparatively simple when the magnetic field is either *tremendously large* or *extremely small*. It will be convenient to take up next the case of a "weak" field, *i.e.*, a field weak enough so that the Zeeman splitting is small as compared even with the separations in the ordinary fine structure. In such a field, J is a good quantum number and the effect of the field is merely to separate the states of different M that compose each J level.

According to Eq. (223), the effect of a magnetic field upon the energy is determined by the average values of the electronic orbital and spin momenta in the direction of the field. The most important

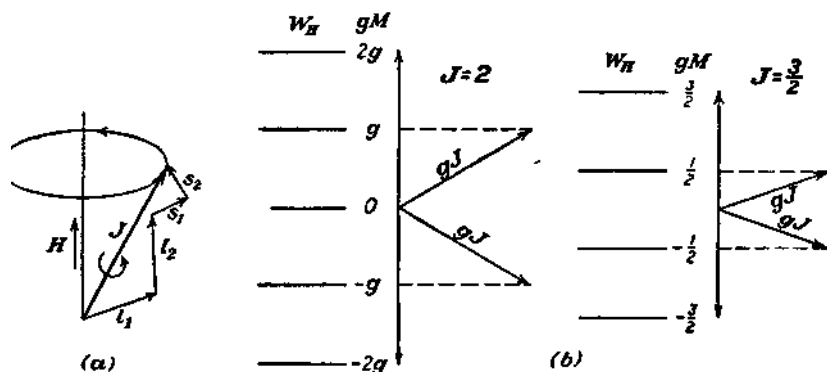


FIG. 111.—Diagram illustrating (a) vector addition in a magnetic field and (b) the separation of the magnetic levels in a weak field.

results obtained from wave mechanics in regard to these average values can be inferred readily from an appropriate vector diagram interpreted semiclassically, in analogy with the treatment of the spin-orbit effect in Sec. 133. We may think of the various electronic angular momenta as added together vectorially into a resultant vector angular momentum; and this may be represented in a certain sense by the addition of the electronic l - and s -vectors into a resultant J -vector, in analogy with similar operations in Sec. 137 [see Fig. 111(a)]. The components of the l - and s -vectors perpendicular to J may be thought of as averaging zero, as in Sec. 138, since the diagram could be revolved around J without affecting the J -vector. Thus, in the average, only components parallel to J will remain effective; and the projection of these components on the direction of the magnetic field H will then represent the average values of the components of the electronic momenta in the direction of the field.

Now the projection of any vector that is parallel to J upon the direction of H is proportional to $\cos \theta$, where θ is the angle between J and H , since J is of fixed length. Furthermore, in a comparison of the various quantum states, $\cos \theta$ is proportional to the component of J along H , and hence to the quantum number M . Thus the average components of the orbital and the spin momenta of any electron in the direction of H , which have been denoted by G_{lH} and G_{sH} , respectively, should both be proportional to M ; and we may write, in terms of constants of proportionality g_1 and g_2

$$\bar{G}_{lH} \approx g_1 \frac{Mh}{2\pi}, \quad \bar{G}_{sH} = g_2 \frac{Mh}{2\pi},$$

and

$$\sum \bar{G}_{lH} + 2 \sum \bar{G}_{sH} = g \frac{Mh}{2\pi},$$

where g_1 and g_2 are constants of proportionality (independent of M), and $g = \sum g_1 + 2 \sum g_2$, 2 indicating summation over all electrons in the atom.

Substituting the value given by (233) for the quantity in parentheses in Eq. (222), we have for the magnetic moment of the atom in the direction of the field, when it is in one of the M states,

$$\mu = - \frac{eh}{4\pi mc} gM. \quad (234)$$

Making the same substitution in the expression given in Eq. (223) for the energy, and adding W_{0J} , representing the energy for the J level in the absence of a magnetic field, we have for the energy

$$W_{JM} = W_{0J} + \frac{eh}{4\pi mc} gMH. \quad (235)$$

(Here, as usual, e stands for the numerical electrohic charge.)

The last equation shows that the "magnetic" states composing a given J level and corresponding to different values of M become separated and equally spaced in a weak magnetic field, but the spacing differs from that corresponding to classical theory, as represented by Eq. (231). The value of the factor g may vary from one J level to another. It is often called the *Lande splitting factor* or *g-factor*. The splitting can be visualized by means of an obvious modification of the vector diagram, as in Fig. 111(b), where vectors are drawn of length gJ instead of J .

A more convenient equation for spectroscopic purposes is obtained if we divide all energies by hc so as to express them in equivalent

wave numbers. At the same time, let us introduce the Lorentz unit defined above,

$$\tilde{L} = \frac{cH}{4\pi mc^2} = 4.67 \times 10^{-5} H \text{ cm.}^{-1}, \quad (236)$$

representing the normal separation of either a line from the w line. Then (235) becomes

$$\tilde{W}_{JM} = \tilde{W}_{0J} + gM\tilde{L}. \quad (237)$$

For the values of W_{JM} corresponding to two different J levels, we may then write

$$\tilde{W}_{0J'} + g'M'\tilde{L}, \quad \tilde{W}_{0J''} + g''M''\tilde{L}.$$

Subtracting these two expressions, we have for the frequencies of the Zeeman lines between the two J levels, all of which would coalesce into one line if the magnetic field were absent:

$$\tilde{\nu} = \tilde{\nu}_0 + (g'M' - g''M'')\tilde{L}, \quad (238)$$

where $\nu_0 = W_{0J'} - W_{0J''}$ and represents the frequency when $H = 0$. If J' , J'' are the corresponding values of J for the two levels, in Eq. (238) M' takes on integrally spaced values from $-J'$ to J' , and M'' from $-J''$ to J'' , subject only to the selection rule

$$\Delta M = M'' - M' = 0 \text{ or } \pm 1.$$

As in Sec. 145, transitions in which $\Delta M = 0$ give rise to π lines; those in which $\Delta M = \pm 1$, to σ lines. The arrangement of the components into which a spectral line is split by a magnetic field is called the "Zeeman pattern" of the line.

All observed Zeeman patterns in weak fields are found to agree with Eq. (238) if suitable values of g' and g'' are inserted in it. If $g' \neq g''$, the observed Zeeman pattern leads to a knowledge not only of g' and of g'' but also of the values of J for both levels, since the number of magnetic states for given J is $2J + 1$. (The student may be interested to invent a method for inferring values of J and g from such observations.) Thus, the Zeeman effect is of great utility in the analysis of spectra. A mathematical description of the anomalous Zeeman effect, in terms of an equation equivalent to (238), was worked out empirically by Lande before the discovery of wave mechanics.

To calculate a *theoretical* value of g requires, in general, a knowledge of the wave functions for the atom. If *LS coupling occurs*, however, wave mechanics furnishes a simple general formula for g ; and it

happens that this formula can be inferred by resorting again to the vector diagram.

For this purpose we draw the diagram in fuller detail than in Sec. 137, so as to represent the addition, (1) of the electronic Z-vectors into a resultant L-vector and of the electronic s-vectors into a resultant S-vector; (2) the addition of L and S into a resultant J ; and (3) the projection of J on an axis in the direction of the magnetic field H . See Fig. 112, which is drawn for 3 electrons, with $l_3 = 0$. This procedure represents in a certain sense the vector addition of the electronic orbital and spin angular momenta into separate resultants and then the addition of these orbital and spin resultants into the total vector angular momentum. The magnetic effect is then determined by the average values of the projections of the electronic l - and s -vectors on the direction of H , which we shall calculate indirectly.

An argument similar to that used in deducing the LS multiplet spacing in Sec. 138 indicates that the components of the l -vectors perpendicular to L will average to zero, when all possible azimuths about the L -vector are considered. The sum of the projections of the l -vectors on H is given, therefore, by the projection of the sum of their components along L , which is just L itself. The

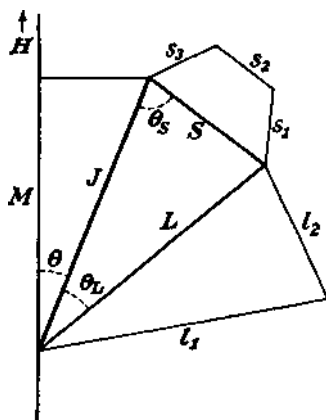


FIG. 112.—Diagram illustrating LS coupling in a magnetic field H .

$$\Sigma l_H = L \cos \theta_L \cos \theta$$

where θ_L is the angle between L and J and θ is the angle between J and H . Similarly, the sum of the average projections of all s -vectors on H is

$$\Sigma s_H = S \cos \theta_S \cos \theta$$

where θ_S is the angle between S and J . But, by geometry, $J \cos \theta = M$ and

$$2JL \cos \theta_L = J^2 + L^2 - S^2, \quad 2JS \cos \theta_S = J^2 + S^2 - L^2.$$

Hence

$$\sum l_H = \frac{M}{2J^2} (J^2 + L^2 - S^2), \quad \sum s_H = \frac{M}{2J^2} (J^2 + S^2 - L^2).$$

Now the vector diagram represents angular momenta in a certain special way. The average components of electronic orbital and spin momentum in the direction of H will be, respectively,

$$\sum \tilde{G}_{lH} = \frac{\hbar}{2\pi} \sum l_H, \quad \sum \tilde{G}_{sH} = \frac{\hbar}{2\pi} \sum s_H; \quad (239a,b)$$

but the squares of the momenta, in units of $\hbar^2/4\pi^2$, are given, not by J^2 , L^2 , S^2 , but by $J(J+1)$, $L(L+1)$, $S(S+1)$ (see Sec. 137). Making corresponding changes in the expressions just found for $2l_H$ and $2s_H$ and then substituting in (239a,b), we obtain

$$\begin{aligned} \sum \tilde{G}_{lH} + 2 \sum \tilde{G}_{sH} &= \frac{M\hbar}{4\pi J(J+1)} [3J(J+1) + S(S+1) - L(L+1)] \\ &= \frac{M\hbar}{2\pi} \left[1 + \frac{J(J+1) + S(S+1) - L(L+1)}{2J(J+1)} \right]. \end{aligned}$$

Comparing this with Eq. (233), we see that, in case LS coupling is valid,

$$g = 1 + \frac{J(J+1) + S(S+1) - L(L+1)}{2J(J+1)}. \quad (240)$$

Theoretical formulas for g in case jj coupling holds can also be worked out.¹ To calculate g from theory in other cases is more difficult.

The values of g inferred from observation agree, in general, fairly well with the theoretical formulas whenever the coupling is shown by other considerations to approximate to the type assumed.

147. Zeeman Patterns of LS Multiplets in a Weak Field.—Equation (238) shows that the splitting of a spectral line in a weak field will depend upon the values of g for both of the levels involved. A wide variety of patterns is possible. Since the various component lines of an LS spectral multiplet arise from levels varying in the value of J , these levels will also have different g values, in general, and the lines of the multiplet will, therefore, exhibit different Zeeman patterns. *Corresponding lines* in the spectral terms of the *same series*, however, should exhibit the same type of pattern; for such lines have a common final level, and their initial levels have the same J , S , and L , and hence the same value of g . This conclusion from the theory is in agree-

¹Cf. WHITE, *op. cit.*

ment with observation; it is known as "Preston's rule," having been discovered empirically by Preston in 1898. Preston's rule is sometimes of use in deciding what lines belong together in a series.

The following particular cases deserve comment.

(a) $g' = g'' = 0$. If $g = 0$ both for the initial and for the final state, we see from Eq. (238) that the line is *not split at all* in a weak field. Such cases can arise in several ways. *Every level with $J = 0$ can be supposed to have $g = 0$.* Actually, in Eq. (240) the fraction is indeterminate; but if $J = 0$, there is only one magnetic level, hence no splitting of the J level; and the single level with $M = 0$, by Eq. (235) is not even displaced in a weak field. It is customary to write $g = 0$ when $J = 0$, but it seems simpler just to write $g = 0$. A jump between two levels both having $J = 0$ is excluded, to be sure, by the usual selection rule; but it can also happen that $g = 0$ when $J = 0$, e.g., for a $^1D_{3/2}$ level ($J = 1/2, S = 3/2, L = 2$) or a $^1F_1$ ($J = 1, S = 2, L = 3$). A line of the type $^3F_1 \rightarrow ^5D_0$ should, therefore, show no Zeeman effect at all. This is confirmed by experiment (e.g., the line $\lambda = 5,713$ Å in the arc spectrum of titanium).

(b) If $g' = g'' = g \neq 0$, or if $J = 0$ for one level only and $g = 0$ for the other, the Zeeman pattern is a *triplet*. For, if $J = 0$ for the second level, in Eq. (238), $M'' = 0$ and

$$\nu = \nu_0 + gM'L.$$

Since $M' - M'' = 0$ or ± 1 , $M' = 0$ or ± 1 here, and three lines are obtained. Again, if both levels have the same $g = 0$,

$$\nu = \nu_0 + g(M' - M'')L;$$

the selection rule for M gives again three lines, although in this case some of the lines arise each from several different atomic transitions, as in the huge-field effect. The outer or r lines are displaced a distance gL from the central T line.

If also ($7 = 1$, the triplet is a *normal* one (Sec. 145), spaced as in classical theory. The value $g = 1$ always occurs when $S = 0$, so that $J = L$. Thus all lines in a *singlet spectrum* exhibit the *normal* Zeeman effect in a weak field. An example is the singlet spectrum of mercury (Sec. 140), or the singlet zinc line shown in Fig. 107. In singlet levels the electron spins are opposed and so produce no magnetic effect, whereas the orbital effect by itself produces the classical pattern, as we found to be the case in a huge field (Sec. 145). Zeeman had the good fortune to succeed first in resolving the components in **singlet** lines, e.g., in the spectra of zinc, cadmium, or tin.

A few of the theoretical values of g for LS levels are shown in Table VII. Some typical spectral patterns, as predicted by the theory, are shown in Fig. 113. All but one are examples of the anomalous Zeeman effect. The lengths of the vertical lines as drawn indicate the relative theoretical intensities of the spectral lines as

TABLE VII.—THE LANDE SPLITTING FACTOR g FOR LS COUPLING

S	$J \searrow L =$	$S \quad 0$	$P \quad 1$	$D \quad 2$	$F \quad 3$	$G \quad 4$
0	L	0	1	1	1	1
$\frac{1}{2}$	$L - \frac{1}{2}$	..	$\frac{2}{3}$	$\frac{4}{3}$	$\frac{6}{7}$	$\frac{8}{9}$
	$L + \frac{1}{2}$	2	$\frac{4}{3}$	$\frac{6}{5}$	$\frac{8}{7}$	$1\frac{1}{9}$
1	$L - 1$	..	0	$\frac{1}{2}$	$\frac{3}{2}$	$\frac{5}{2}$
	L	..	$\frac{3}{2}$	$\frac{7}{6}$	$1\frac{1}{2}$	$2\frac{1}{2}$
	$L + 1$	2	$\frac{3}{2}$	$\frac{4}{3}$	$\frac{5}{4}$	$\frac{6}{5}$

seen in a direction perpendicular to the field, which are usually in agreement with experiment; T lines are drawn above, a lines below the horizontal axis. Three dots indicate the normal triplet in the same field. In each case the two J levels involved in the transition are indicated.

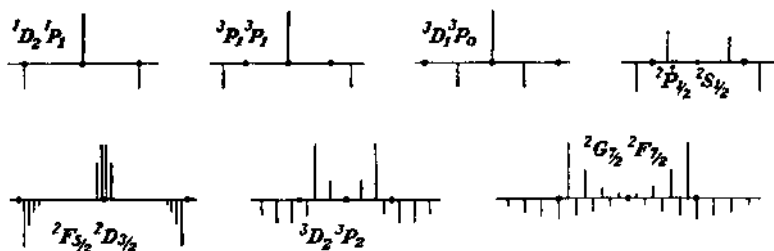


FIG. 113.—Some Zeeman patterns for LS coupling in a weak field. Heights of lines indicate roughly the relative intensities of the lines, T lines are drawn above, a lines below. The three dots indicate the position of the normal triplet in the same field.

The student will find it instructive to check the spacing from the formulas in a number of cases. Other diagrams of the same sort are shown in H. E. White's book.¹

To illustrate the use of the theoretical equations, consider the familiar D lines of sodium, which arise from transitions of the type

¹ Op. cit., pp. 161, 222.

$^2P_{3/2} \rightarrow ^2S_{1/2}$ and $^2P_{1/2} \rightarrow ^2S_{1/2}$. For the $^2P_{3/2}$ state, we insert in Eq. (240) $J = 3/2$, $S = 1/2$, $L = 1$, obtaining

$$g = 1 + \frac{\frac{3}{2}(\frac{3}{2} + 1) + \frac{1}{2}(\frac{1}{2} + 1) - 1(1 + 1)}{2 \times \frac{3}{2}(\frac{3}{2} + 1)} = \frac{4}{3}.$$

Since M may take on any one of the four values $3/2, 1/2, -1/2, -3/2$, gM in Eq. (235) or (237) may have any one of the four values

$$H, \quad \frac{6}{3}, \frac{2}{3}, \quad -\frac{2}{3}, \quad -\frac{6}{3}.$$

The level $^2P_{3/2}$ is thus split into four magnetic levels removed from the

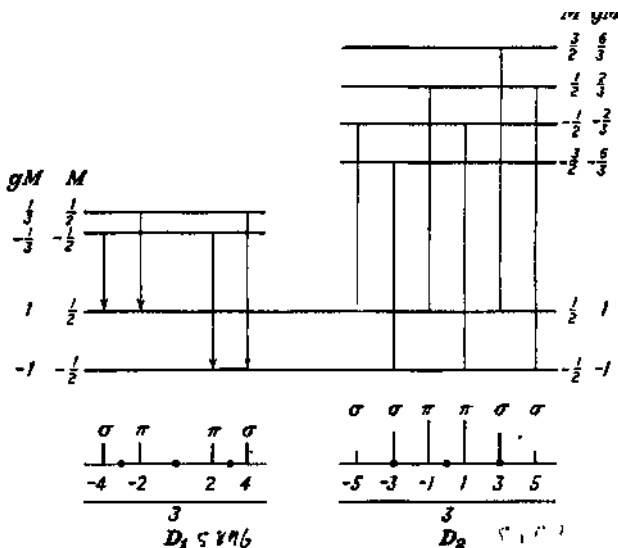


FIG. 114.—Weak-field Zeeman effect for the D lines of sodium. The same scale is used for all of the Zeeman separations, also for the theoretical relative intensities of the component lines as shown by the relative heights of the lines as drawn. The position of the normal triplet in the same field is shown by three dots,

position of the $^2P_{3/2}$ level in zero field by gM Lorentz units or $gML \text{ cm.}^{-1}$. Similarly, the state $^2P_{1/2}$ is split into two levels for which

$$g = 1 + \frac{\frac{1}{2}(\frac{1}{2} + 1) + \frac{1}{2}(\frac{1}{2} + 1) - 1(1 + 1)}{2 \times \frac{1}{2}(\frac{1}{2} + 1)} = \frac{2}{3}, \quad gM = \pm \frac{1}{3};$$

and the $^2S_{1/2}$ level is split into two with

$$g = 1 + \frac{\frac{1}{2}(\frac{1}{2} + 1) + \frac{1}{2}(\frac{1}{2} + 1)}{2 \times \frac{1}{2}(\frac{1}{2} + 1)} = 2, \quad gM = \pm 1$$

Figure 114 shows these levels, and also the transitions between them that are allowed by the selection rule $\Delta M = 0$ or ± 1 . The resulting

line patterns are shown at the bottom of the figure. The expected displacement of each component line from the position for zero field is calculated, in Lorentz units, by subtracting the two gM values; *e.g.*, for the left-hand σ component of D_2 , we have $-\frac{2}{3} - 1 = -\frac{5}{3}$ Lorentz units, which is shown below the line in an obvious notation. Further explanation is added below the figure.

We note that D_1 or ${}^2P_{1/2} \rightarrow {}^2S_{1/2}$ should be resolved into four components, two T lines ($AM = 0$) polarized parallel to the field and two r lines ($Ailf = \pm 1$) polarized perpendicularly. D_2 or ${}^2P_{3/2} \rightarrow {}^2S_{1/2}$ should be resolved into six components of which the outer four are σ lines. A glance at the photograph in Fig. 107 shows that the theoretical predictions are confirmed by experiment, as to the patterns themselves, and at least roughly also as to intensities. Examination of the lines through a Nicol prism shows that the polarizations also are correctly predicted.

148. The Paschen-Back Effect.—In the discussion in the last two sections the emitting atom has been assumed to be in a "weak"

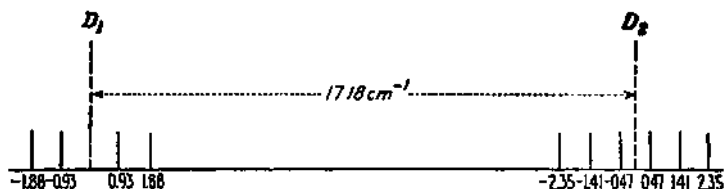


FIG. 115.—Relative separations of the D lines of $\text{Na}(11)$ and their Zeeman components in a field of 30,000 oersteds.

magnetic field. The term "weak" is, of course, relative, since fields in excess of 10,000 oersteds are usually necessary to produce observable Zeeman patterns. A magnetic field is said to be "weak" (or strong) according as the separations of the Zeeman components of each of the several lines of a multiplet group are small (or large), compared with the separation of the lines of the multiplet from one another. Figure 115, for example, shows diagrammatically to scale the D lines of sodium with their respective Zeeman patterns in a field of 30,000 oersteds. The Zeeman separations are seen to be small compared with the separation between D_1 and D_2 . For the D lines of sodium, therefore, a field of 30,000 oersteds is a "weak" field.

Consider, however, the Zeeman effect in lithium. The separation of the two components of the first line of the principal series¹ is of the order of 0.3 cm.^m¹ A field of 30,000 oersteds would be expected to produce Zeeman separations several times larger than this doublet

¹See Table V, p. 338.

separation [according to (230), if $H = 30,000$, $L = 1.4$ cm.¹]. For lithium, therefore, a field of 30,000 oersteds is a "strong" field.

It was found by Paschen and Back¹ that, whatever the Zeeman patterns of a given LS multiplet of lines may be in a weak field, the Zeeman pattern of the group as a whole in a strong field always approximated a normal triplet. The transition from weak to strong field is an interesting phenomenon, exhibiting the following stages. As the field is increased to such a point that the magnetic levels into which one J level is split begin to approach those belonging to another J level, departures from the weak-field arrangement of the lines begin to occur. As the field increases further, magnetic sublevels that differ both in J and in M may cross one other. Eventually, when the field has become strong enough, the magnetic levels are observed to group themselves more or less closely into a set of equally spaced composite levels, and the spectral pattern then approximates a normal triplet! i.e., we have then the case of a huge field as described in Sec. 145. I

The transition from one type of Zeeman effect to the other is accompanied by changes in the angular momenta. As long as the field is weak, the resultant orbital momentum and the resultant spin momentum remain added approximately into a resultant momentum of fixed magnitude, characterized by a certain value of the quantum number J . As the field grows stronger, however, this coupling is progressively broken down, and eventually, in a strong field, it disappears. Then A and 2 become good quantum numbers, along with L and S ; the resultant orbital momentum by itself has a fixed component in the direction of the field of magnitude $\Delta h/2\pi$, and the resultant spin momentum has similarly a component $\Sigma h/2\pi$. The total angular momentum in the direction of the field is thus $Mh/2\pi$, where $M = A + 2$. Under these conditions the quantum number J ceases to have any significance at all. The theoretical treatment of the energy then parallels closely the theory for 1 electron in a huge field, as described in Sec. 145. The energy can be written, approximately, in analogy with Eq. (227),

$$W = W_0 + \frac{eh}{4\pi mc}(\Lambda + 2\Sigma)H,$$

W_0 being a constant for a given LS term. The selection rules are, besides those for L and S ,

$$\Delta\Lambda = 0 \text{ or } \pm 1, \quad \Delta M = 0 \text{ or } \pm 1, \quad \Delta\Sigma = 0.$$

¹ *Ann. d. Physik*, vol. 39, p. 897 (1912); vol. 40, p. 960 (1913).

The normal spectral triplet is produced in essentially the same way as for 1 electron.

In this discussion, however, we have ignored the spin-orbit and additional electrostatic effects. These effects actually separate slightly those levels which have the same value of $A + 2$ but different values of M and thereby introduce a sort of fine structure into the Zeeman pattern.

A corresponding theory is easily developed for the case of jj coupling. In this case, however, or in any case other than that of LS coupling, there is an intermediate stage in the progressive change of the Zeeman pattern.

All of these changes of pattern as the strength of the magnetic field increases are known collectively as the *Paschen-Back effect*. The effect is said to be *incomplete* when the magnetic field is strong enough to make the Zeeman splitting for the energy levels larger than the mutual separation of the J levels comprised in a given LS or jj term, or comprised in a similar group in other cases, but yet small as compared with the separations between the terms or groups as wholes. Such a field is called merely "strong." If the field could be increased still further, until the magnetic levels of two different *terms*, such as a ^1P and a ^1D , approached each other, the pattern should change again, approaching the normal triplet as the field becomes so large that its effect predominates over everything else except the central field. This stage, described in Sec. 145 as the effect in a "huge field," is known as the *complete* Paschen-Back effect. Unfortunately, it can scarcely be realized experimentally except in 1-electron atoms, in which the first two stages of the Zeeman effect are practically unobservable.

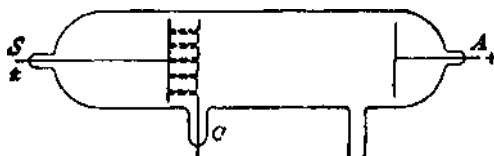
149. The Stark Effect—An effect of an *electric field* on spectral lines, somewhat analogous to the Zeeman effect, was discovered by Stark in 1913 and is known as the "Stark effect." It is observed in the well-known "canal rays," when these are moving in an electric field of several thousand volts per centimeter. An arrangement for producing the Stark effect is shown diagrammatically in Fig. 116. The anode and cathode in a glass tube are A and C , respectively, in which the gas is maintained at such a pressure that the Crookes dark space in front of C is several centimeters long. The cathode C is perforated with small holes through which pass, in cylindrical bundles, luminous streams of atoms of the gas that have acquired a positive charge immediately in front of C . These streams of atoms are the canal rays. A third terminal S is placed behind C , at a distance of a few millimeters, and an electric field of several thousand (20,000 and up) volts per centimeter is maintained between S and C . For the "transverse" Stark

effect, light from the canal rays enters the spectrometer in the direction of the arrow.

Stark observed that the spectral lines which are emitted by the canal rays when no field exists between S and C are split up, when the field* is applied, into numerous components; these components are polarized, some of them parallel to the field (p- or π -components), others perpendicularly (s- or σ -components). The splitting is usually proportional, at first, to the square of the electric field strength; in an intense field, however, the pattern becomes quite different, in analogy with the Paschen-Back effect in a magnetic field.

Classical theory could not account for the Stark effect, but a good account of it was developed by Epstein, using an extension of Bohr's

To Specframeier



To Pump

FIG. 116.—Arrangement for producing the Stark effect.

theory; and it is completely explained by wave mechanics. The treatment of the Stark effect constituted, in fact, one of the most striking successes of quantum theory. In spectroscopy, however, this effect is of minor interest because it does not, like the Zeeman effect, serve to illuminate other parts of the subject. Examples of Stark patterns are given in White's book.¹

150. The Stern-Gerlach Experiment.—The space quantization of atomic angular momenta, upon which we have based the discussion of the Zeeman effect, constitutes a characteristic feature both of the older quantum theory and of wave mechanics, in contrast with classical theory. Its occurrence is not definitely confirmed by the existence of the Zeeman effect, however, in view of the partial success of classical theory in accounting for these phenomena. An experiment which directly reveals the space quantization itself was proposed by O. Stern in 1921 and was carried out by him in collaboration with Gerlach.² -

In elementary magnetic theory, it is shown that a magnet tends to move so as to increase the magnetic flux through it in the direction of

¹ *Op. cit.*; cf. also HERZBERG, *op. cit.*

² STERN, *Zeits. f. Physik*, vol. 7, p. 249 (1921); GERLACH and STERN, *Zeits. f. Physik*, vol. 8, p. 110; vol. 9, p. 349 (1922); *Ann. d. Physik*, vol. 74, p. 673 (1924).

its magnetic axis. In a uniform field, the only result is that the magnet experiences a torque tending to line it up with the field. In a *nonuniform* field, however, the magnet experiences a *translatory force* as well.

Suppose now that a slender beam of atoms having magnetic moments travels in the direction of the x-axis across a magnetic field whose lines are approximately parallel to the y-axis but whose magnitude increases rapidly in the direction of the z-axis. After passing through the field, let these atoms be collected on a suitable target. Then atoms which have a component of their magnetic moment in the direction of the field will be deflected sideways; if the component of the moment has the same direction as the field, they will be deflected toward $+z$; if it has the opposite direction, toward $-z$.

So far, the predictions of classical and quantum theory coincide. According to *classical* theory, however, each atom will enter the field

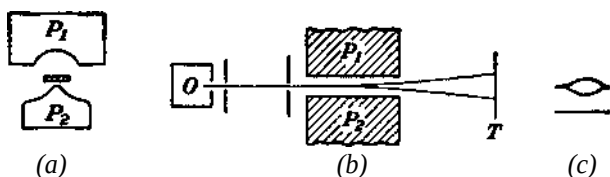


FIG. 117.—Illustrative of the Stern-Gerlach experiment, (a) and (c) represent views seen in the direction of the beam; (c) shows the trace on the target (exaggerated), with and without the magnetic field.

with its magnetic axis inclined at some angle θ to the field, and the axis will then execute a Larmor precession about the field at the fixed angle θ . Since all values of θ will occur among the atoms, their deflections will be distributed in continuous fashion, and the atoms, instead of forming a small spot on the target, will be drawn out into a continuous band.

According to *quantum* theory, on the other hand, each atom will enter the field in a certain quantum state, defined with the direction of the field as an axis. Its magnetic moment in the direction of the field will be gM Bohr magnetons (if the field is not too strong), where g is the Lande splitting factor and M has one of several integral values [cf. Eq. (234)]. The beam will be broken up, therefore, into separate beams and will form on the target a series of distinct spots, one for each possible value of M .

The arrangement used in the experiment of Stern and Gerlach is shown diagrammatically in Fig. 117. The nonhomogeneous field was produced between pole pieces of which one had a sharp edge, so that near it the field was much stronger than elsewhere [cf. Fig. 117(a)], A

strap-shaped beam of silver atoms was formed by evaporating silver in a heated oven O and allowing atoms from the vapor to stream out through collimating slits; the beam [shown by the small rectangle in Fig. 117(a)] traveled closely past the sharp edge of pole piece P_2 and was condensed on a plate at T . With no field, the beam formed a narrow line on the plate [Fig. 117(c), lower line]. When the magnetizing current was turned on, the line was not widened continuously! but was divided into two lines, as shown in the upper part of Fig. 117(c), except at the ends, which were produced by atoms passing at some distance from the sharp edge.

Space quantization of the silver atoms was thus clearly revealed. From careful measurements of the separation of the two lines and of the gradient of the magnetic field strength, it was calculated that each silver atom had a magnetic moment in the direction of the field of 1 Bohr magneton, with an error of 10 percent at most.

These results, and others obtained subsequently, are in complete agreement with the predictions of wave mechanics, silver atom is normally in a $^2S_{1/2}$ state, for which $g = 2$; thus, half of the atoms should have $M = \frac{1}{2}$ and a moment in the direction of the field of $2 \times \frac{1}{2} = 1$ Bohr magneton; whereas, for the other half, $M = -\frac{1}{2}$ and the moment is -1 ." Similar results were obtained subsequently, with greater precision, by Taylor¹ and by Leu² for sodium and potassium, which are likewise in $^2S_{1/2}$ states. Atoms of zinc and cadmium, in Leu's experiments, were unaffected by the magnetic field. Atoms of thallium gave a double line corresponding to a moment of $\frac{1}{3}$ magneton. Zinc and cadmium are normally in $^1S_0$ states, which have $M = 0$ and no magnetic moment. For thallium the normal state is inferred from spectroscopic evidence to be $^2P_{1/2}$; thus $M = \frac{1}{2}$ or $-\frac{1}{2}$ again, but $g = \frac{2}{3}$ instead of 1 [by Eq. (240), in which $J = \frac{1}{2}$, $S = \frac{1}{2}$, $L = 1$].

The method of magnetic deflections has been extended very ingeniously by Rabi for the measurement of the magnetic moments of *nuclei*. The student will find it worth while to look up his papers on this subject in the *Physical Review*, vol. 49 onward.

151. Isotope Structure and Hyperfine Structure.—The ordinary fine structure due to spin-orbit interaction does not exhaust the possibilities of fine details in spectral lines. Even before 1900, Michelson and others had shown, by means of the interferometer, that many spectral lines possess a further structure much finer still. This came to be known as hyperfine structure.

¹ TAYLOR, *Phys. Rev.*, vol. 28, p. 276 (1926).

² LEU, *Zeits.f. Physik*, vol. 41, p. 551 (1927).

After the discovery that many chemical elements consist of several isotopes (Sec. 195), it was believed for a time that each component of a hyperfine pattern of lines was emitted by a different isotope. But later, hyperfine structure was discovered in the spectra of some elements which, as shown by the mass spectrograph, consist of only 1 isotope. An example is bismuth; the line $\lambda = 3,596 \text{ \AA}$ contains six hyperfine components spread over a range of $0.3 \text{ \AA}$ or 2.3 cm^{-1} . It was suggested by Pauli in 1924 that this effect might be due to an occurrence in the nucleus of angular momentum and an associated magnetic moment. It appears that both causes are operative. To distinguish between them, a tendency has arisen to restrict the term *hyperfine structure* to that which is associated with the angular momentum of the nucleus, the other type being called *isotope structure*.

A photograph illustrating *isotope structure* in the spectrum of

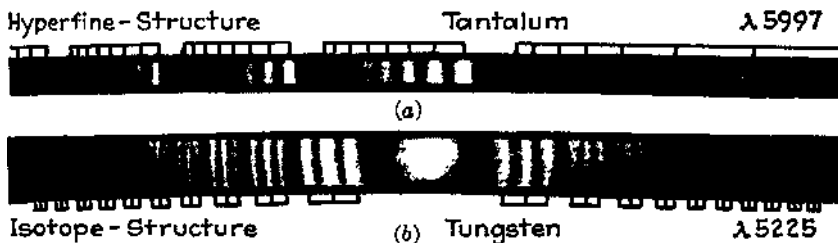


FIG. 118—Photographs illustrating (a) hyperfine structure in a spectral line of tantalum and (b) isotope structure in a line of tungsten. (After Grace, More, Mac-Afillan, and White; from H. E. White, "Introduction to Atomic Spectra.")

tungsten is shown in Fig. 118. The spectrum was formed with a Fabry-Perot etalon, hence the same pattern appears repeated many times in different orders. Tungsten consists of 4 isotopes, with mass numbers 182, 183, 184, 186, the first two being somewhat less abundant than the last two. The three lines observed were ascribed to the 3 isotopes of even number, the line due to 183 being assumed to be masked by the others.

Isotope shifts of spectral lines are found, in general, to be proportional, for a given element, to the differences in atomic mass. As a rule, however, they do not seem to be purely the result of nuclear motion, as is the difference in the positions of those lines in the spectra of hydrogen and of ionized helium which would otherwise fall together (Sec. 97). The shift associated with an increase in mass from one isotope to another is even sometimes in opposite directions in different elements. The cause of most isotope shifts is thought to lie, perhaps, in some departure from the Coulomb type of field very close to the

nucleus, the departure being different for different elements and even for different isotopes of the same element.

An interesting example in which the isotope shift probably is a simple mass effect is afforded by the spectrum of hydrogen, in which the isotope shift was used as a guide by Urey and his collaborators in the discovery of heavy hydrogen or deuterium.¹ The Rydberg constants for the two kinds of hydrogen atoms can be found by substituting $M = M_1$ and $M = 2M_1$ successively in Eq. (146a) in Sec. 97, which gives

$$R_1 = \frac{R_\infty M_1}{M_1 + m}, \quad R_2 = \frac{2R_\infty M_1}{2M_1 + m},$$

M_1 being the mass of an ordinary hydrogen atom and m the mass of an electron. The difference in wave length for a given line of wave length λ will then be, nearly enough,

$$\Delta\lambda = - \frac{\lambda(R_2 - R_1)}{R_\infty} = - \frac{\lambda m}{2M_1}.$$

Since $M_1/m = 1,837$, this means that, for example, the $H\beta$ line for deuterium should lie $4,861/3,674 = 1.32$ Å on the violet side of that for ordinary hydrogen atoms. This line was observed to be very faintly visible, in the expected position, in the spectrum from a sample of common hydrogen, and it increased in strength as the hydrogen was subjected to operations which should increase the relative concentration of deuterium.

The theoretical treatment of *true hyperfine structure*, on the other hand, resembles closely the treatment of *LS* fine structure due to the spin-orbit effect, as described in Sec. 138. The total angular momentum of the *nucleus* is assumed to be fixed, so that it can be represented by a quantum number J . The value of I may be 0 or integral or half-integral, and it may be different for different isotopes of the same chemical element. The angular momentum in question is presumably the resultant of spin momenta of the protons and neutrons of which the nucleus is composed; hence, it is commonly said to be due to nuclear "spin." I remains constant for a given atom because in ordinary physical processes the nucleus remains always in its lowest energy level. This level is degenerate, however, if $J > 0$; there is a second quantum number M_I , restricted to integrally spaced values such that $|M_I| \leq I$. Thus, I and M_I in nuclear theory correspond to J and M in atomic theory.

If the action of the nucleus on the electrons were purely that of a

¹ UREY, BRICKWEDDE, and MURPHY, *Phys. Rev.*, vol. 40, p. 1 (1932).

central field, the only effect of the nuclear angular momentum would be a $(2I + 1)$ -fold degeneracy of all quantum states, with the selection rule $\Delta M_I = 0$. But there may exist also an associated nuclear magnetic moment, as in atoms. Equation (227) in Sec. 145 and its physical interpretation suggest that a single particle with a mass M equal to that of a proton, if, like an electron, it has spin $\frac{1}{2}$, might have an effective magnetic moment of magnitude

$$\frac{e\hbar}{4\pi Mc} = \frac{1}{1,837} \text{ Bohr magneton.}$$

This is called a *nuclear magneton*. Observation indicates, however, that the magnetic moment of the proton or hydrogen nucleus is actually nearer to 3 nuclear magnetons.¹

If such a nuclear moment exists, wave mechanics indicates that the angular momentum of the nucleus is to be added to the resultant angular momentum of the electrons by a process of *IJ* coupling exactly

TABLE VIII.—HYPERFINE STRUCTURE OF THE COMPONENTS OF THE DOUBLET OF THALLIUM: $6S^2 6p^2 P_{1/2}, ^3P_{2/2} > 6S^2 7s^2 S$

λ	$\Delta\tilde{\nu}$, cm. ⁻¹	Intensity	λ	$\Delta\tilde{\nu}$, cm. ⁻¹	Intensity
3,776.888 Å		3	5,352.184 Å		8
	0.40			0.60	
3,776.830		10	5,352.014		3
	0.71			0.14	
3,776.729		8	5,351.974		10

analogous to the *LS* coupling of resultant electronic orbital and spin momenta (Sec. 138). The result is then, in general, that each of the quantum states of the atom as dealt with hitherto is broken up into a group of quantum states characterized by two additional quantum numbers F and M_F , which have reference to the total angular momentum of the atom, that of the nucleus included.

The interaction between the nuclear magnetic moment and the electronic motions and spins then slightly separates those states which have different values of F . The number of hyperfine energy levels thus produced may be different for electronic states of different J . If $J > I$, their number is $2I + 1$; but, if $J < I$, it is $2J + 1$. For the spacing of the hyperfine levels the analog of Eq. (211) for *LS* multiplets is obtained; the separation of successive levels is proportional to the larger of the values of F for the two levels [cf. Eq. (212) in Sec. 138]. As selection rules, in addition to all others, we have

$$\Delta F = 0 \text{ or } \pm 1, \quad \Delta M_F = 0 \text{ or } \pm 1.$$

¹ KELLOGG, RABI, and ZACHARIAS, *Phys. Rev.*, vol. 50, p. 472 (1936).

A *hyperfine spectral multiplet* results from transitions between the hyperfine levels composing two ordinary J levels. Usually it happens that the spacing of the hyperfine sublevels in one of the two J levels is much larger than that in the other, so that the former spacing stands

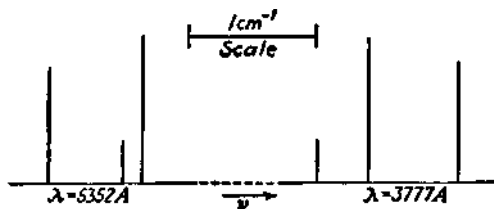


Fig. 119.—Hyperfine structure of the thallium doublet $\lambda = \frac{5,352}{3,777}$ angstroms.

out in the spectral multiplet as observed, the finer structure due to the other J level being frequently not resolved at all. The result is then an easily recognizable "flag" type of pattern, a good example of which is shown in Fig. 118.

As an illustration of the theory, consider the following doublet in the spectrum of thallium (Tl 81):

$$6s^2 6p \, ^2P_{1/2} \rightarrow 6s^2 7s \, ^2S_{1/2} (\lambda = 3,777 \text{ Å}; \bar{\nu} = 26,478 \text{ cm.}^{-1}) \quad (a)$$

$$6s^2 6p \, ^2P_{3/2} \rightarrow 6s^2 7s \, ^2S_{1/2} (\lambda = 5,352 \text{ Å}; \bar{\nu} = 18,684 \text{ cm.}^{-1}) \quad (b)$$

Under high resolution¹ each line of this doublet is seen to be made up of three components. Their wave lengths and estimated intensities are given in Table VIII; their relative positions on a frequency scale are shown in Fig. 119.

To produce three component lines, as observed, there must be two hyperfine sublevels in each J level, just as in the production of the $D \rightarrow P$ triplets in the alkali spectra there are two J levels in each LS term (Sec. 138). Hence, here $2 = 2I + 1$, and so $I = \frac{1}{2}$, for the thallium nucleus. Then for $J = \frac{1}{2}$, $F = I \pm \frac{1}{2}$ or $F = 1$ or 0 ; for $J = \frac{3}{2}$, $F = 2$ or 1 . The spacing of the hyperfine levels is easily determined from the observed separations in the spectrum. The level diagram is shown for one line of the doublet in Fig. 120.

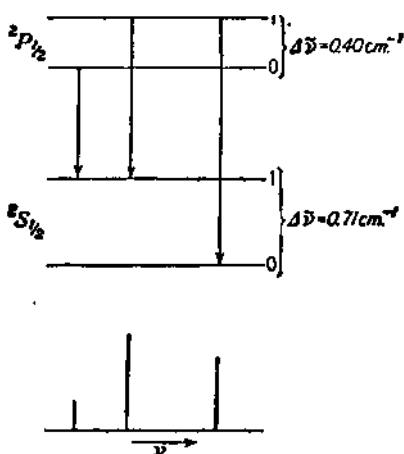


Fig. 120.—Energy levels and transitions for the hyperfine structure of the thallium line $\lambda = 3,777$ angstroms.

¹ BACK, *Ann. d. Physik*, vol. 70, p. 367 (1923).

It may be remarked that the same pattern would result if the doublet separations in Fig. 120 were to be interchanged. According to Back and Wulff, the Zeeman pattern for the hyperfine components unambiguously assigns the energy-level differences.¹

In Table IX are shown a few values of I and of the nuclear magnetic moment μ expressed in terms of nuclear magnetons. M is the mass number of the isotope in question (Sec. 195). Where no references

TABLE IX.—NUCLEAR SPIN NUMBERS I AND MAGNETIC MOMENTS μ

	Z	M	I	μ		Z	M	I	μ
H	1	1	$\frac{1}{2}$	2.790*	K	19	41	$\frac{3}{2}$	0.215†
H	1	2	1	0.856*	Rb	37	85	$\frac{5}{2}$	1.340‡
Li	3	6	1	0.821§	Rb	37	87	$\frac{3}{2}$	2.733‡
Li	3	7	$\frac{3}{2}$	3.253§	Cd	48	111, 113	$\frac{1}{2}$	-0.6
N	7	14	1	0.403**	Cs	45	133	$\frac{7}{2}$	2.558**
Na	11	23	$\frac{3}{2}$	2.215**	Hg	80	199	$\frac{1}{2}$	0.5
Al	13	27	$\frac{5}{2}$	3.630††	Hg	80	201	$\frac{3}{2}$	-0.6
K	19	39	$\frac{3}{2}$	0.391**	Tl	81	203, 205	$\frac{1}{2}$	1.6 ±
K	19	40	4	-1.290††	Bi	83	209	$\frac{9}{2}$	3.8 ±

* MILLMAN and KUSCH, *Phys. Rev.*, vol. 60, p. 91 (1941).

† KUSCH, MILLMAN, and RABI, *Phys. Rev.*, vol. 57, p. 765 (1940).

‡ KUSCH and MILLMAN, *Phys. Rev.*, vpl. 56, p. 527 (1939).

§ RABI, MILLMAN, KUSCH, and ZACHARIAS, *Phys. Rev.*, vol. 55, p. 526 (1939).

** KUSCH, MILLMAN, and RABI, *Phys. Rev.*, vol. 55, p. 1176 (1939).

†† MILLMAN and KUSCH, *Phys. Rev.*, vol. 56, p. 303 (1939).

tt ZACHARIAS, *Phys. Rev.*, vol. 61, p. 270 (1942).

are given, the data are taken from Condon and Shortley's book² and were originally calculated from the observed spectroscopic separations by means of perturbation theory.

It was remarked in 1940 by Mattauch³ that hyperfine structure had not been found in the spectrum from any atom whose nucleus was composed of an even number of protons and an even number of neutrons, so that both its atomic number and its mass number are even (Sees. 195 and 206). Presumably all such nuclei have $I = 0$ and $\mu = 0$. Examples are He, C, O, Ne, Mg.

162. The Breadth of Spectral Lines.—No spectral line as observed is perfectly sharp, no matter how great the resolving power of the spectrometer. A line devoid of structure appears densest in the center and fades out symmetrically on the edges. It is of interest to inquire as to the causes of this distribution of the energy within a line.

¹ BACK and WULFF, *Zeits. f. Physik*, vol. 66, p. 31 (1930).

² CONDON and SHORTLEY, "Theory of Atomic Spectra," p. 424, 1935.

³ MATTAUCH, *Zeits. f. Physik*, vol. 117, p. 246 (1940).

For convenience in comparing lines, the *width* Δ of a line is defined to be the distance between two points, one on each side of the center, at which the intensity is half as great as it is in the center of the line (Fig. 121). (This is sometimes called the half-intensity width.) The principal causes of line broadening are the following:

(a) *Doppler Effect*.—The *observed* frequency of a spectral line may be slightly changed by the motion of the radiating atom in the line of sight, owing to Doppler's principle, the apparent frequency *increasing* if the motion is *toward* the observer and *decreasing* if the motion is *away* from the observer. Only for those atoms which have no component of velocity in the direction of the observer will the observed frequency of the emitted light be equal to the actual frequency. As is well known, the measurement of the velocity of stars in the line of sight is based on this principle.



FIG. 121.—Illustrating definition of Δ , the width of a line.

In a luminous gas, such as the mercury vapor in the mercury arc lamp, the atoms are moving with a maxwellian distribution of velocity, the average velocity increasing with temperature. It is obvious, therefore, that a spectral line emitted by a gas must comprise, as observed, a *range* of frequencies *symmetrically* distributed about the frequency emitted by the atom when at rest and, further, that this range should increase with increasing temperature. The distribution of intensity throughout the line is determined by Maxwell's distribution of velocities. According to Rayleigh,¹ the brightness of the line at a distance $\pm \Delta \nu$ wave numbers from the center is proportional to

$$e^{-k\Delta\nu^2}$$

where k is a constant which depends on the temperature and on the mass of the atom, and e is the Napierian base of logarithms. From this formula it follows that the width Δ defined as above, if it is due entirely to the Doppler effect, should be

$$\Delta = 0.72 \times 10^{-6} \lambda \sqrt{\frac{T}{M}}$$

wave-length units, where T is the absolute temperature and M is the atomic or molecular weight of the radiating atom or molecule.²

Careful studies of the breadth of lines in the spectra of the rare gases were made by Fabry and Buisson,³ using an interferometer

¹ RAYLEIGH, *Phil. Mag.*, vol. 27, p. 298 (1889).

² Cf. WHITE, H. E., "Introduction to Atomic Spectra," p. 419, 1935.

³ FABRY and BUISSON, *J. de Physique*, vol. 2, p. 442 (1912).

method. They confirmed the formula for λ as written above. At liquid-air temperatures the breadth of the krypton line $\lambda = 5,570 \text{ \AA}$ was found to be only $0.006 \text{ \AA}$, practically all of which could be ascribed to the Doppler effect resulting from thermal agitation at that temperature.

(b) *Pressure*.—Using the light from a single unresolved spectral line, it is possible, under favorable conditions, to produce interference fringes when the difference in the path of the two beams is as much as several hundred thousand wave lengths. In terms of classical theory, this fact was interpreted to mean that the wave train sent out by any particular atom is continuous, *i.e.*, is without change of phase, for at least that number of vibrations. In order that the atom may emit wave trains of this length, it must be "free from interruptions" for a corresponding period of time. In the terminology of the kinetic theory of gases, this means that the *mean free time* between collisions with other atoms must, on the average, exceed the time required to emit a complete wave train, since it may be assumed that a collision with another atom would cause either a change of phase or excessive damping or other disturbance. Analogous conclusions are deduced, although as a result of rather different reasoning, from wave mechanics.

Now, collisions between atoms become more frequent the higher the pressure of the gas for a given temperature. The higher the pressure, the shorter, therefore, should be the wave trains and the more frequent the abrupt changes of phase. An increase of temperature, at given density, also increases both the collision rate and the pressure. Thus in general, at higher gas pressure, not only should there be a broadening of the line due to the Doppler effect but also an additional broadening due to the increasing frequency of phase changes resulting from collisions. Michelson confirmed this by showing, from measurements with the interferometer, that below a pressure of the order of a millimeter, the breadth of the hydrogen line $\lambda = 6563 \text{ \AA}$ is almost entirely due to the Doppler effect; but that at higher pressures the line becomes considerably broader.

The observed broadening is enhanced by direct disturbances of the energy levels, or of the radiation process itself, when another molecule comes close to the radiating atom.

(c) *Natural Line Breadth*.—According to wave-mechanical theory, a line ought also to exhibit a small "natural" width even when emitted by an atom at rest. As an analogy, a classical oscillator, radiating energy, would decrease continually in amplitude; it would emit, therefore, a damped wave train of finite effective length. It can be shown that a damped train of sine waves is equivalent to the superposition of a large number of perfectly regular trains of great length, with

slightly differing frequencies. Such radiation, observed as a spectral line, would, therefore, be broadened slightly. In the visible region, the natural line breadth is mostly much less than 0.001 Å and so is not detectable. It is easily observable, on the other hand, in the case of X-rays.

(d) *Incipient Stark Effect*.—In the spectra emitted by discharge tubes, a common cause of broadening is the production of a small Stark effect by strong electric fields, which are not great enough to produce an observable splitting of the line but are sufficient to make it appear perceptibly broader than it otherwise would.

Thus there are many very different factors that may be responsible for the broadening of spectral lines.

153. Molecular Spectra.—In preceding sections of this chapter we have considered only spectra emitted by *atoms*. There are other spectra in vast variety which are believed on good evidence to be emitted by *molecules* containing 2 or more atoms.

Thus, in the visible part of the spectrum emitted by a discharge tube containing hydrogen, only three or four lines belonging to the Balmer series are emitted by free atoms, which have been produced by the dissociation of molecules. Many other lines are observed, however, mostly fainter, which are ascribed to emission by the undissociated molecules themselves. Again, if one looks at the spectrum of the carbon arc with a spectroscope of moderate resolving power, one will observe, at the extreme (violet) edge of the visible part of the spectrum, "bands," or very broad lines, which are sharply defined and brightest on the long-wave-length edge and which fade out gradually toward shorter wave lengths. With higher resolving power, these "bands" are seen to be composed of a large number of lines which are crowded together at the long-wave-length edge, called the "head" of the band, and are separated farther and farther toward the short-wave-length side, the lines, however, being so close together as to appear, under low resolving power, like a *continuous* spectrum. These bands are ascribed to molecules of cyanogen, CN. There are also many groups or bands of lines in the infrared which have been found to have a molecular origin.

In the space available here we can discuss only a few of the features of molecular spectra, referring the student to other books or to the literature for further information.¹

In general, the spectrum emitted by any given kind of molecule

¹ HERZBERG, G., "Molecular Spectra and Molecular Structure," 1939; RTJARK and UREY, "Atoms, Molecules and Quanta"; WEIZEL, W., "Bandenspektren," Akademische Verlagsgesellschaft m.b.H., Leipzig, 1931.

can be divided into *three spectral ranges* which correspond to different types of transition between molecular quantum states. The principal exception is that the spectrum of some molecules is confined to only one of these ranges. Simple reasoning in terms of familiar classical assumptions concerning molecular structure leads us to expect such a feature in molecular spectra, and the reasoning needs only to be translated into wave-mechanical terms in order to constitute a correct theoretical approach to the subject.

(a) *Rotation Spectra*.—Suppose a molecule were a rigid structure but contained electrical charges so disposed that the molecule possessed an electric moment. If such a molecule were to rotate, according to classical theory, it would emit radiation for essentially the same reason that an electron revolving in a circle would radiate. It is readily seen that the radiation would consist of sine waves having a single frequency, viz., the frequency of rotation. Conversely, radiation falling upon such a molecule would tend to set it into rotation, energy of the radiation being at the same time absorbed.

A few spectral lines corresponding to this simple picture have been observed in the far infrared and are said to constitute "rotation spectra."

(b) *Vibration-rotation Spectra*.—If the molecule were not rigid but contained atoms capable of vibration under elastic forces about equilibrium positions, and if the chemical binding were of the ionic type (Sec. 125), so that some atoms contained an excess of positive charge and others an excess of negative, then according to classical theory radiation would be emitted by the vibrating atoms as they move back and forth. Unless the molecule were at the same time rotating, the frequency emitted would be that of the atomic vibration. If the molecule were rotating, however, the emitted line would be divided into two lines having frequencies respectively greater or less than the frequency of the atomic vibration, in essentially the same way as, in the classical theory of the Zeeman effect (Sec. 40), the precession caused by a magnetic field modifies the frequencies emitted by a vibrating electron.

Furthermore, it would be anticipated that, if the amplitude of vibration became large, the atomic vibrations, although still periodic, would no longer be simple harmonic. This is true even in the familiar example of the vibrations of a pendulum. The radiation emitted could then be resolved by Fourier analysis into wave trains with frequencies representing the fundamental and the harmonic overtones of the atomic vibrations. Each of these separate frequencies would then be split up further by rotation of the molecule.

Many spectra corresponding roughly to this classical picture are known in the infrared and are called "vibration-rotation spectra."

(c) *Electronic Spectra*.—Finally, according to classical ideas, an electron in the molecule might vibrate by itself and so radiate. The emitted radiation would be affected, however, both by the vibrations of the atoms in the molecule and by the rotation of the molecule as a whole. It would probably be one of the outer electrons that radiated in the optical region of the spectrum, and its frequency would be much affected by the instantaneous position and motion of the nuclei. The rotation of the molecule would tend to split up the emitted lines as in the emission of the vibration-rotation spectrum. Molecular bands in the visible and ultraviolet, such as the cyanogen bands described above, are believed to correspond roughly to this third classical picture.

The three types of molecular spectra thus characterized will be taken up in succession for a brief discussion.

154. Rotation Spectra.—As a simple model to illustrate certain features of the behavior of actual molecules, we may imagine a molecule to consist of several mass points held rigidly at fixed distances from each other. The quantum states for such a molecule, according to wave mechanics, would be characterized by fixed values for the angular momentum, in the same way as are the states of an atom (Sec. 130). The corresponding quantum number J , however, is confined here to integral values (zero included). The discussion will be restricted hereafter almost entirely to *diatomic molecules*. If there are only two mass points in the molecule, the line joining them is an axis of symmetry, and only rotation about an axis perpendicular to this line has significance; furthermore, the moment of inertia about all such perpendicular axes will have the same value.

The relation between angular momentum and energy is found to be the same according to wave mechanics as in classical theory. For the angular momentum G and energy W , we shall have, therefore, in terms of the angular velocity ω and moment of inertia I ,

$$G = I\dot{\omega}, \quad W = \frac{1}{2} I\omega^2, \quad \therefore W = \frac{G^2}{2I}.$$

Inserting here the wave-mechanical value of G^2 (Sec. 130), we have

$$\begin{aligned} G^2 &= \frac{J(J+1)\hbar^2}{4\pi^2}, \\ W &= J(J+1) \frac{\hbar^2}{8\pi^2 I}. \end{aligned} \quad (241)$$

Such a molecule can radiate only if it possesses an electrical moment, which will be the case, for example, if one mass point has associated with it a positive charge and the other an equal negative charge. The selection rule for J is then found to be the same as that for the quantum number l of a single electron in a central field:

$$\Delta J = \pm 1. \quad (242)$$

Since in the present case W and J increase or decrease together, $\Delta J = -1$ will correspond to emission of energy and $\Delta J = +1$ to absorption. Thus, in a transition from state J to state $J - 1$, the emitted frequency will be, since

$$J(J+1) - (J-1)(J-1+1) = 2J, \\ \nu = \frac{\Delta W}{h} = 2BJ, \quad B = \frac{h}{8\pi^2 I}.$$

Thus a molecule of the type under consideration will emit a spectrum consisting of *equally spaced lines, with frequencies equal to a multiple of a fixed number, B* . For an emission line, J refers to the initial state for the molecular transition; for an absorption line, to the final state. The corresponding type of energy-level diagram is illustrated in Fig. 122. The arrows pointing downward refer to transitions for emission, those pointing upward to transitions for absorption.

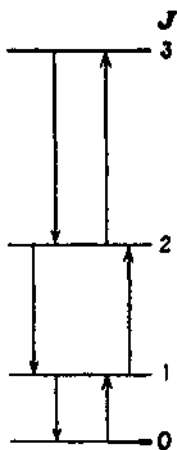


FIG. 122.—
Energy-level di-
agram for a rota-
tion spectrum.

We may form some idea of the spectral region in which such lines may be expected to occur by estimating the probable magnitude of I from other considerations. The molecule HCl , for example, might be expected to have an electric moment, since it is strongly ionized in solution. In the formula $I = 2mr^2$, in which r denotes distance from the center of mass of the molecule, the Cl atom will contribute little because it is so close to the center of mass. The mass of a hydrogen atom will be approximately the reciprocal of the Avogadro number (Sec. 79) or the number of

atoms in a gram-atom. Hence, if we assume $r = 10^{-8}$ cm. for the H atom, we have

$$I = 10^{-16}/(6 \times 10^{23}) = 1.7 \times 10^{-40} \text{ c.g.s. unit.}$$

Thus, for the constant B in Eq. (243a), if we also divide by c to obtain it in wave-number units, we find

$$B = \frac{6.6 \times 10^{-27}}{8\pi^2 \times 1.7 \times 10^{-40} \times 3 \times 10^{10}} = 17 \text{ cm.}^{-1},$$

roughly. For $J = 10$, we find $\tilde{\nu} = 2BJ = 340 \text{ cm.}^{-1}$, or
 $\lambda = 0.029 \text{ mm.} = 29\mu$.

Now all of the hydrogen halides in the gaseous state show broad absorption lines in the far infrared which are very nearly equally spaced and whose wave numbers are very nearly multiples of a constant quantity. These lines are believed to result from transitions in which the rotational state of the molecule alone changes, very nearly according to the simple theory just described. Thus for H O, Czerny¹ found a series of absorption maxima at wave lengths ranging from 120u to 44u. The corresponding wave numbers are listed as ν_{obs} in Table X.

TABLE X.—ABSORPTION SPECTRUM OF HCl IN THE FAR INFRARED

J	$\tilde{\nu}_{obs}$	$\Delta\tilde{\nu}$	$\tilde{\nu}_{calc} = fJ - gJ^2$
4	83.03		83.06
5		(41.27)	103.75
6	124.30		124.39
7	145.03	20.73	144.98
8	165.51	20.48	165.50
9	185.86	20.35	185.94
10	206.38	20.52	206.30
11	226.50	20.12	226.55

Under the heading J is given the larger of the two values of J assigned to each transition. We note that the spacing of the maxima is almost uniform but shows a slight trend. In the last column of the table are shown values of ν calculated from the empirical formula, $\nu = fJ - gJ^2$, with $f = 20.79$, $g = 0.0016$; according to this formula

$$\Delta\nu = f - g[J^2 - (J - 1)^2] = f - (3J^2 - 3J + 1)g \text{ cmr}^{-1}$$

If we have sufficient confidence in the theoretical interpretation, such data can be used as a basis for the calculation of molecular moments of inertia. Substituting $B = cf/2 = 20.79 c/2$ in Eq. (2436), we find for the moment of inertia of the HCl molecule, in a state of minimum rotation, $I = 2.45 \times 10^{40}$ gram cm.². This value agrees satisfactorily with our rough estimate, 1.7×10^{40} . The slight decrease in $\Delta\nu$ as J increases, and hence also in the apparent value of B , implying an increase in J , is what would be expected if the atoms were not tightly bound together but became slightly pulled apart by centrifugal action as the speed of rotation increases.

Thus, wave mechanics accounts satisfactorily for the rotation type of molecular spectrum.

¹ CZERNY, *Zeits. f. Physik*, vol. 34, p. 227 (1925).

165. Vibration-rotation Spectra, (a) *Approximate Theory of a Vibrating Diatomic Molecule.*—To obtain a simple model of a diatomic molecule in which atomic vibrations can occur, suppose the atoms are themselves point masses but that, instead of being rigidly bound together, they are held by a force that varies with the distance between them. Let this force correspond to a potential energy V , which, plotted as a function of the distance r between atomic centers, is represented by a curve of the type of that marked V in Fig. 123. The potential has been arbitrarily taken to be zero at $r = \infty$. The force exerted by either atom on the other will be proportional to the slope of this curve; it is represented by the curve marked F in Fig. 123. From the point r_0 at which V has its minimum value outward, the force is attractive; for $r < r_0$, it is repulsive, rising rapidly to high values. Under the influence of such a force, according to classical theory, the atoms could be at rest and in equilibrium at the distance r_0 . If disturbed moderately, they would vibrate about this point; if given kinetic energy exceeding $-V_0$, however, where V_0 is the value of V at $r = r_0$, they would

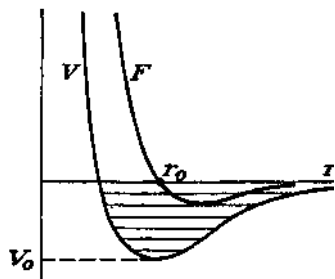


FIG. 123.—The mutual potential energy V and force F , measured positively as a repulsion, between 2 atoms in a diatomic molecule.

fly apart entirely, i.e., dissociation of the molecule would ensue. Such a picture corresponds to the observed properties of molecules, and it is also suggested by wave-mechanical theory.

According to wave mechanics, however, there would be a set of *discrete quantum states* for the positions of the atoms relative to each other, with energies as suggested by the horizontal lines in Fig. 123. The lowest of these quantum states would correspond to an energy a little greater than V_0 ,—it might be said, therefore, that even in their lowest state the atoms have a certain amount of kinetic energy, which is analogous to the zero-point energy $h\nu/2$ of a harmonic oscillator. Thus the interatomic distance is not quite fixed even when the molecule is in this state, although the most probable value of the interatomic distance will be close to r_0 . Let us number the quantum states in the order of increasing energy, denoting the number of a state by $v > 0$. Let the corresponding energy be denoted by W_v . The values of W_v are all negative, zero energy belonging to a state in which the atoms are at rest at infinity. Thus, $-W_v$ represents the energy of dissociation of the molecule when it is in state **number v ; and not rotating.**

The total number of the discrete states may be finite or infinite, depending upon the form of the potential curve.

The wave functions and energies for the first few states should resemble those for a harmonic oscillator. For, if we expand V in a Taylor series about $r = r_0$ and note that $dV/dr = 0$ at $r = r_0$, we obtain

$$V = V_0 + \frac{1}{2} \left(\frac{d^2V}{dr^2} \right)_{r=r_0} (r - r_0)^2 + \frac{1}{6} \left(\frac{d^3V}{dr^3} \right)_{r=r_0} (r - r_0)^3 - \dots$$

and the first two terms of this series represent a potential function of the same type as that for a harmonic oscillator (Sec. 114).¹ The wave functions and energies for the higher states, however, must depart considerably from those for an oscillator, owing to the influence of the remaining terms of the series. Thus the selection rule for the harmonic oscillator, $\Delta v = \pm 1$, cannot be expected to hold, in general, for a diatomic molecule, even $\Delta v = 0$ being possible. The probability of a jump may be expected, however, to fall off rather rapidly as Δv increases.

To obtain quantum states for the whole molecule, allowance must then be made for rotation of the line joining the atoms. If we suppose that, to a sufficient approximation, the energies of rotation and of vibration are additive, we may write for the total energy, by Eq. (241) above,

$$W = W_v + J(J + 1)Bh, \quad B = \frac{h}{8\pi^2 I}. \quad (244a,b)$$

The various values of W_v are called *vibrational* levels of the molecule. The energies represented by both terms on the right in Eq. (244a) are called *vibration-rotation* levels, or simply *rotational* levels. In actual cases Bh is usually very small relative to the difference between successive values of W_v ; hence, the rotational levels belonging to each vibrational level form a closely spaced group. The general arrangement of the rotation-vibration levels is illustrated in Fig. 124, in which, however, the relative spacing of the rotational levels is enormously exaggerated, and only a few of these levels are shown.

Because of the relative smallness of B , all lines arising from transitions between two given vibrational levels will lie close together; they are said to constitute a *band*, because with low resolving power they appear as a continuous streak in the spectrum. Bands arising from

¹ The wave functions for the first few states should be practically zero for values of $(r - r_0)$ that are large enough to give appreciable magnitude to terms of the series beyond the first two.

transitions in which only the vibrational and perhaps rotational energies of the molecule change are called *vibration-rotation bands*.

For a molecule composed of two point masses, the selection rule for J is found to be, as for the pure rotation lines,



$$\Delta J = \pm 1.$$

For many molecules of more complicated structure, $\Delta J = 0$ is also allowed. A few transitions allowed by the rule $\Delta J = \pm 1$ are indicated in Fig. 124.

From the approximate expression for the energy in Eq. (244a), we find, since

$$J(J+1) - (J-1)(J-1+1) = 2J,$$

for the frequencies emitted in a transition between levels $v', J-1$ and v'', J , or between

$$v', J-1; v'', J: \quad \nu = \nu_{v'v''} - 2BJ,$$

$$J = 1, 2, 3, \dots;$$

$$v', J; v'', J-1: \quad \nu = \nu_{v'v''} + 2BJ,$$

$$J = 1, 2, 3, \dots;$$

$$\nu_{v'v''} = \frac{W_{v'} - W_{v''}}{h}.$$

It is assumed here that $v' > v''$ and $W_{v'} > W_{v''}$.¹

According to Eqs. (245a,b), the lines in a given vibration-rotation band emitted by a diatomic molecule should be equally spaced in frequency, with a separation $2B$. It is clear, however, that the value $\nu = \nu_{v'v''}$ cannot occur, so that there is no line corresponding to pure vibration. The nearest frequencies to this are $\nu = \nu_{v'v''} + 2B$ and $\nu_{v'v''} - 2B$. The central line of the band should thus

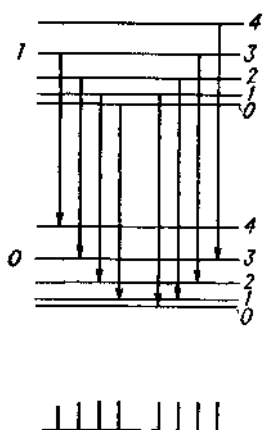


FIG. 124.—The energy levels for a rigid diatomic molecule (the relative magnitude of the rotational separations being enormously exaggerated), and a few of the transitions for the first band.

appear to be missing, as is illustrated in Fig. 124.

(6) *Observed Absorption Bands of Vibration-rotation Type.*—Absorption bands answering the foregoing description were observed long ago in the near infrared regions of the spectra of the hydrogen

¹ In absorption, therefore, v' refers to the final state into which the molecule is raised by the absorption of a quantum. In considering the spectrum, however, it is simplest to think only of numerical energy differences, without bothering to remember which state is the initial or final one in the radiative process.

halides. The band of HC1 at 3.5 μ is shown, as observed by Imes,¹ in Fig. 125, percent of absorption being plotted against wave length. In order to obtain this curve, Imes passed the radiation from a Nernst glower through 15 cm. of HC1 gas at atmospheric pressure and then through a spectrometer containing a prism of rock salt, and measured the transmitted radiation with a thermopile and a galvanometer. The band in question had been observed by Burmeister in 1913 as a continuous band with a double top and was thought to be capable of explanation in terms of classical theory. The molecules were assumed to possess a maxwellian distribution of rotational velocities, so that according to classical theory the splitting described in Sec. 152(a) would be expected to spread the vibrational line out into a broad,

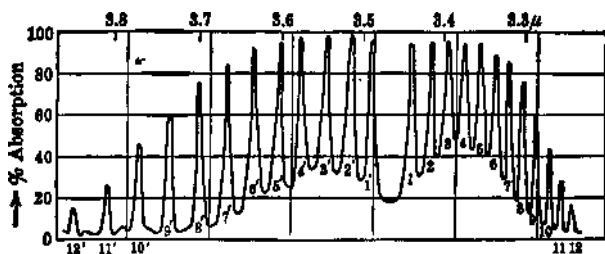


FIG. 125.—The principal absorption band of HC1 in the near infrared {after E. S. Imes²}. The numbers on the lines give the larger J for the transition, those for which AJ is opposite in sign to Δv being primed. {Reprinted by courtesy of Prentice-Hall Inc., New York; from G. Herzberg, "Molecular Spectra and Molecular Structure," 1939.)

continuous band. When later workers observed a succession of maxima in the band, it was realized that an explanation could only be found in terms of a discrete set of rotational quantum states.

The anticipated absence of the central line in Fig. 125 is very evident. The lines show some approach to equal spacing, but a considerable systematic variation occurs; this is well brought out by the following table of frequency differences, in which those values of J are primed that refer to transitions in which $\Delta v = v'' - v'$ and AJ have opposite signs, so that the change in rotational energy is opposite in sign to the change in vibrational energy and the frequency is less than $\nu_{v'v''}$.

J	12'	11'	10'	9'	8'	7'	6'	5'	4'	3'	2'	1'
Δf	26.62	26.15	25.91	25.03	25.30	23.67	23.85	23.04	22.49	22.48	21.75	

$$v', J; v'', J - 1$$

J	(1')	1	2	3	4	5	6	7	8	9	10	11	12
Δf	41.60	20.24	18.35	18.62	17.38	17.62	16.30	14.92	15.89	14.47	13.27	12.85	

¹ IMES, *Astrophys. J.*, vol. 50, p. 251 (1919).

² *Loc. cit.*

(c) *The More Accurate Empirical Theory of Vibration-rotation Bands.*—The variation in $\Delta\tilde{\nu}$ as observed indicates that the theory-developed above is too simple and requires modification. , Even on classical assumptions, complications are to be expected in an actual case. When the molecule is in rotation, the position of relative equilibrium for the 2 atoms should shift outward from r_0 in Fig. 123, because of centrifugal action; the result would be an increase in the moment of inertia I . An effect of the spreading upon the vibrational motion of the molecules is also to be anticipated.

The true formula for the energy levels must, therefore, be more complicated than that expressed in Eq. (244a), although the quantum numbers ν and J can undoubtedly be retained as labels. Various empirical equations have been proposed from time to time, alongside efforts to improve the treatment based on quantum theory. A form of equation equivalent to the following has become accepted as standard, all terms being expressed in cm^{-1} :

$$\tilde{W} = \tilde{W}_\nu + B_\nu J(J+1) - D_\nu J^2(J+1)^2 \cdots \quad (243)$$

The coefficients B_ν and D_ν vary somewhat with ν and are often written as power series in this variable:

$$B_\nu = B_0 - \alpha\nu + \gamma\nu^2, \quad D_\nu = D_0 + \beta\nu^2 \cdots$$

In these series, wave-mechanical considerations indicate that B_0 and D_0 should be positive, and that D_ν as a whole should be comparatively small. If we omit the D_ν term entirely, we obtain, in place of (245a) and (245b), changing to wave numbers instead of frequencies:

$$\begin{aligned} \nu', J-1; \nu'', J: \quad \tilde{\nu} &= \tilde{\nu}_{\nu'\nu''} - (B_{\nu'} + B_{\nu''})J \\ &\quad + (B_{\nu'} - B_{\nu''})J^2 \cdots \end{aligned} \quad (247a)$$

$$\begin{aligned} \nu', J; \nu'', J-1: \quad \tilde{\nu} &= \tilde{\nu}_{\nu'\nu''} + (B_{\nu'} + B_{\nu''})J \\ &\quad + (B_{\nu'} - B_{\nu''})J^2 \cdots \end{aligned} \quad (247b)$$

These equations yield the lines in order of increasing frequency if we take decreasing values of J in the first equation and increasing values in the second. For $\Delta\nu$ between successive lines, we find, denoting by J_m the larger of the two values of J for each transition:

$$\begin{aligned} \nu', J-1; \nu'', J: \quad \Delta\tilde{\nu} &= B_{\nu'} + B_{\nu''} - (B_{\nu'} - B_{\nu''})(2J_m - 1) \cdots, \\ \nu', J; \nu'', J-1: \quad \Delta\tilde{\nu} &= B_{\nu'} + B_{\nu''} + (B_{\nu'} - B_{\nu''})(2J_m - 1) \cdots. \end{aligned}$$

In view of the direction in which J changes throughout the band, these equations imply that, if $B_{\nu'} < B_{\nu''}$, $\Delta\tilde{\nu}$ will decrease from one line to the next, as it actually does. The observed rate of decrease, however, is greater at higher than at lower frequencies. This is

easily seen to be accounted for by the term in D_v in Eq. (246), provided $D_{v'}$ and $D_{v''}$ are both positive. The term in D_v represents essentially the stretching effect due to centrifugal action, which was referred to above; this stretching, by increasing the moment of inertia J , should decrease the energy, as is evident from the approximate equations (244a,b); hence, in Eq. (246), D_v should be positive.

(d) *Relation between Rotation and Vibration-rotation Bands.*—It is obvious that there should exist a close relation between the vibration-rotation bands and the pure rotation spectrum of a given substance; indeed, the rotation spectrum is simply a band arising from transitions in which $\Delta v = 0$. If we put $v_{v'v''} = 0$ in Eq. (2456), we obtain $v = 2BJ$, in agreement with (243a) in the last section. Thus, according to the simplest type of theory, the *separations of the lines should be the same in the rotation and in the vibration-rotation parts* of the spectrum, being equal to $2B$ in both cases. This conclusion is easily tested. According to (245a) and (2456), the separation of the innermost two lines ($J = 1$) in a vibration-rotation band is $4B$. In the HC1 band just described, this separation is 41.60 cm^{-1} . Half of this, or 20.8 , happens to agree exactly with the value obtained for the spacing of the lines at $J = 0$ from the empirical formula that was cited in the last section for the rotation spectrum of HC1, viz., $\Delta v = f = 20.79$. The comparison will not be pressed more closely here, because of insufficient accuracy in the data for the vibration-rotation band.

(e) *Bands of Higher Order, and the Heat of Dissociation.*—Besides the relation between the two kinds of bands, another theoretical point that can be tested is the prediction that, although transitions for $\Delta v > 1$ are to be expected, the resulting bands should be relatively weak. If the vibrational levels were equally spaced, as for the harmonic oscillator [Eq. (184a,b) in Sec. 114], the vibrational frequencies of the bands for $\Delta v > 1$ would be proportional to $1, 2, 3, \dots$, just as they are for the harmonic overtones of a classical vibrating system. Thus, if we assume that the HC1 band at 3.46μ arises from a transition between the lowest two vibrational levels ($v'' = 0, v' = 1$), we should expect to find bands diminishing progressively in intensity at about $1.73\mu, 1.15, 0.865$, etc. Actually, by observing the absorption through very thick layers of HC1 gas, bands have been observed at $1.76, 1.20$, and 0.916μ , the last mentioned being 10,000 times weaker than the 3.46 band.¹

The occurrence of these "harmonic" bands with the expected variation in strength, and the agreement of the spacing in all bands,

¹ HERZBERG and SPINKS, *Zeits. f. Physik*, vol. 89, p. 474 (1934); CLEAVES and EDWARDS, *Phys. Rev.*, vol. 48, p. 850 (1935).

including the rotation spectrum, with theoretical expectations, constitute strong evidence that our interpretation of these infrared absorption spectra is essentially correct. A further check is furnished by a connection with the *heat of dissociation* of the molecule. The heat of dissociation represents the energy that must be added to a molecule in its lowest quantum state in order to separate the atoms and leave them at rest an infinite distance apart. Obviously, the energy of the molecule in any possible quantum state must be less than the energy of the separated atoms, else the molecule would dissociate spontaneously. Hence, the quantity $h\nu$ for any molecular line, representing the difference in energy between two molecular states, must be less than the heat of dissociation. For the HC1 band at $X = 0.916u$,

$$h\nu = 6.6 \times 10^{-27} \times \frac{3 \times 10^{10}}{0.906 \times 10^{-4}} = 2.2 \times 10^{-12} \text{ erg.}$$

In the "International Critical Tables,"¹ the heat of dissociation of a gram-molecule of HC1, for dissociation into H₂ and Cl₂, is given as 92 kilojoules, whereas the heats of combination of a gram-atom of H or Cl into H₂ or Cl₂ are, respectively, 211 and 120.3 kilojoules. Hence to dissociate 1 molecule of HC1 into H and Cl requires

$$(92 + 211 + 120.3) \times \frac{10^{10}}{6.02 \times 10^{23}} = 7.0 \times 10^{-12} \text{ erg.}$$

This is more than three times the value of $h\nu$ as just calculated for the highest-frequency vibration band of HC1 that has been observed, so that the theoretical expectation is confirmed.

(f) *Effect of Thermal Agitation upon Infrared Bands.*—No absorption bands are observed for which $\nu'' > 0$ in the spectrum of HC1. This is explained as a consequence of the wide spacing of the vibrational levels relative to the quantity kT . Molecules are seldom thrown by thermal agitation into a vibrational level which can serve as the initial level for an absorption band with $\nu'' > 0$. Even for $X = 3.5u$, $h\nu = 5.7 \times 10^{-18}$ erg; whereas at $T = 288^\circ\text{K.}$ (15°C.),

$$kT = 288 \times 1.38 \times 10^{-16} = 4 \times 10^{-14} \text{ erg.}$$

If T_0 , W_1 are the energies of the two states involved in the production of $X = 3.5u$, the ratio of the number of atoms in the upper of these two states to that in the lower will be the ratio of their Boltzmann factors or, according to Eq. (148a) in Sec. 101

$$\frac{n_1}{n_0} = e^{-(W_1 - W_0)/kT} = e^{-h\nu/kT} = e^{-5/4} = 10^{-0.2}.$$

¹ Vol. V. p. 176.

This is so small that practically all molecules of HC1 will be in their lowest vibrational states ($v = 0$) at ordinary temperatures. Hence only the absorption band for which $v'' = 0$ can be observed; and the observed rotation spectrum is the $0 - 0$ vibration-rotation band ($v = 0$ to $v = 0$).

Among the various *rotational* levels, on the other hand, a wide distribution of the molecules occurs even at room temperature. For the experimental value of $5, 20.6/2 \text{ cm.}^{-1}$, makes Bh equal to

$$6.6 \times 10^{-27} \times 3 \times 10^{10} \times \frac{20.6}{2} = 2.0 \times 10^{-15} \text{ erg.}$$

Thus, for these states, if the energy of the lowest is taken as zero, $W = J(J + 1) \times 2 \times 10^{-15} \text{ erg}$, which is much smaller than kT . For $J = 10$, we find $e^{W/kT} = e^{-5.5} = 0.0041$. Furthermore, there are $2J + 1$ fundamental states for each J (with $M = -J, \dots, J$), so that even states with J as large as 10 will occur

$$(2 \times 10 + 1) \times 0.0041 = 0.08$$

times as often as states with $J = 0$. This result agrees nicely with the number of lines that are actually observed in the HC1 band at 3.5M.

In some other substances, such as I_2 , an appreciable fraction of the molecules are normally in higher vibrational states, and other bands can occur in absorption or in the rotation spectrum.

(g) *Infrared Bands of Other Types of Molecules.*—Vibration-rotation spectra are known for many other molecules than those which have been mentioned. Their structure is often more complex than that described above; in many cases transitions for $\Delta J = 0$ are permitted, so that an additional sequence of lines occurs and there is no gap in the center of the band. Thus the absorption spectrum of CO_2 shows many vibration-rotation bands in the region from 1.46μ to 15.50μ , and that of water vapor shows many from 0.69μ to 6.26μ *; these bands, together with the rotation lines of water vapor, are responsible for the marked infrared absorption of the earth's atmosphere.

In order to exhibit vibration-rotation and rotation spectra of appreciable intensity, a molecule must possess an electrical moment. *Homonuclear* diatomic molecules, such as O_2 , N_2 , H_2 , Cl_2 , possess no moments and, hence, have no spectra of these two types. Gases composed of such molecules are entirely transparent in the infrared.

156. General Theory of Molecular Quantum States.—Up to this point, the *electrons* in the molecule have been ignored. Actually, the

* WEIZEL, *loc. cit.*

wave function for a molecule must contain the coordinates and spins of all of the electrons as well as the coordinates of the nuclei. When atoms unite into a molecule, the inner electrons in each atom can be regarded as remaining associated, more or less, with the nucleus of that atom; but the outer electrons come to belong to the molecule as a whole rather than to any individual nucleus.

(a) *Approximate Separation of Electronic and Nuclear Motions.*—

In many cases, especially for the diatomic molecules, the wave function can be separated approximately into two factors, of which one has reference to the electrons, whereas the second factor represents vibrations of the nuclei and rotation of the molecule as a whole. The first factor then represents the electrons as being in a certain *electronic quantum state*. This electronic factor is analogous to the wave function for the electrons in an atom. Both the energy and the wave function of the electronic state, however, will depend upon the relative positions of the nuclei. In this respect the situation in a molecule is quite different from that in the atom. The mutual electrostatic energy of the nuclei is commonly included in the energy of the electronic state. The electronic energy as so defined possesses a minimum value for certain relative positions of the nuclei; if they move closer together, or move farther apart, the electronic energy rises and, hence, so does the energy of the molecule. Thus the energy of the electronic state is equivalent to a potential energy tending to hold the atoms together in definite relative positions of equilibrium. It is in this way that an interatomic potential energy arises in a diatomic molecule such as was sketched in Fig. 123.

In this same approximation, the energy W of the molecule can be written as the sum of two parts, a negative part W_n representing roughly the average of the electronic energy, including also the electrostatic energy due to the mutual repulsion of the nuclei, and a much smaller positive part W_{rv} , which is associated with vibration of the atoms relative to each other and with rotation of the molecule as a whole. Thus

$$W = W_n + W_{rv}. \quad (248)$$

For a diatomic molecule, the energy W_{rv} is what was denoted by W in Sec. 155, as in Eq. (244a) or (246).

For a detailed discussion of electronic states we must refer the student to other books,¹ but something may be said here by way of explanation of the notation that the student will meet. Only *diatomic* molecules will be discussed.

¹ Cf. HERZBERG, "Molecular Spectra and Molecular Structure," Chaps. IV, V.

(b) *AS Coupling*.—As in the theory of atomic states, electronic spin may be treated in different ways according to circumstances. If the *electronic spin-orbit effects are small*, it is possible, as a first approximation, to assign a fixed value to the component of the orbital angular momentum of the electrons about the nuclear line, or line joining the two nuclei, which is an axis of symmetry. The magnitude of this momentum is denoted by $A\hbar/2\pi$, where A is some positive integer or zero. The quantum number A corresponds to A or M_L in atomic theory, which refers to the component of the angular momentum of the atom in the direction of the axis of coordinates. In a diatomic molecule a unique choice of axis is supplied by the nuclear line. The *total* orbital momentum of the electrons cannot have a fixed value in a molecule; hence there can be no quantum number L . Spectroscopically, however, A plays a role somewhat analogous to L , since states corresponding to different values of A have different energies. Hence, in imitation of the atomic notation for *LS* coupling, states with $A = 0, 1, 2, 3, \dots$ are indicated by the letters $2, \Pi, A, \Phi, \dots$, respectively.

The electronic spins are then combined, with the introduction of a quantum number S . The value of $2S + 1$ is often written as a superscript. Thus, we obtain a set of *AS* electronic states represented by such symbols as

$$^1\Sigma, ^3\Sigma, ^1\Pi, ^3\Pi, \dots; \quad ^2\Sigma, ^4\Sigma, ^2\Pi, ^4\Pi, \dots$$

As with atoms, an important significance of the quantum numbers lies in the associated selection rules. There is a strong tendency for transitions to be limited to those for which

$$\Delta A = 0 \text{ or } \pm 1, \quad \Delta S = 0.$$

Instead of introducing next the electronic spin-orbit effects, however, as is done in atomic theory, we now turn to the consideration of the *nuclear* motions, the effect of which upon the energy we are assuming to be much larger than the electronic spin-orbit effects. Each electronic state of the molecule may be combined with any one of the vibrational states for the nuclei, which may be numbered off as before with a quantum number $v = 0, 1, 2, \dots$. The orbital angular momentum of the electrons is then added vectorially to the angular momentum of rotation of the nuclei about axes perpendicular to the nuclear line. This results in the introduction of a rotational quantum number K , which may have any integral value such that

$$K \geq A.$$

The quantum number K has the significance that the square of the resultant orbital angular momentum of electrons and nuclei together has the value $K(K+1)\hbar^2/4\pi^2$. The general selection rule for K is that

$$\Delta K = 0 \text{ or } \pm 1.$$

We obtain in this way a set of *electronic-vibration-rotation states* for the molecule, numbered with quantum numbers $A S v K$. But the *electronic spin-orbit effect* has not yet been considered. In preparation for its introduction, we must now add the total orbital angular momentum of electrons and nuclei, represented by K , to the total angular momentum due to electronic spin, represented by S , thus obtaining a grand resultant angular momentum, for which we will introduce the usual quantum numbers J and M . The possible values of J are integrally spaced from $K + S$ down $|K - S|$. The effect of the electronic spin-orbit interaction is then, as in atomic LS coupling, to separate states having different values of J . Thus a *fine structure*, analogous to the LS multiplet structure, is introduced into the electronic-vibration-rotation levels. The superscript in such a symbol as $^3\Pi$ refers to the (normal) number of J levels in this fine structure.

In singlet levels, with $S = 0$, there is no fine structure, just as in atoms. For such levels, $J = K$, and the quantum number K need not be used. The HC1 molecule, for example, is normally in a ^1S electronic state; hence, the discussion of the vibrational and rotational states of this molecule as given above was adequate.

(c) *12 Coupling*.—When the electronic spin-orbit effects are not smaller than the effects of molecular rotation, other modes of approach by means of perturbation theory become appropriate. We shall mention only one of them, briefly. If the *spin-orbit* effect is actually *large*, we have the analog of *jj* coupling in atoms. In this case, the electronic orbital and spin momenta about the nuclear line are first added together, their magnitude being represented by a quantum number 12 . The values of 12 are integral or half-integral according as the number of electrons in the molecule is even or odd. Each electronic state, characterized by a value of 12 , is then combined with a vibrational state represented by quantum numbers v and J . Thus in this form of coupling, as when $S = 0$, the rotational levels correspond to the value of J .

(d) The *energy of a molecular state* thus depends, in the case of AS coupling upon five quantum numbers, $A S v K J$, or in the case of 12 coupling upon three, $12 v J$. Its variation with the various numbers presents, however, several different orders of magnitude. The varia-

tion with K or J is comparatively small. Spectral lines that differ only in the K values for the initial and final states lie close together and form a band, similar to the vibration-rotation bands that have been described. In such cases, if $J \gg K$, the energy varies still less with J , the effect of this variation being only to introduce a fine structure into the lines of the band. In the case of Q coupling, on the other hand, the lines of a band arise from differences in the values of J . The energy varies much more rapidly with v than it does with K or J . Each pair of values of v , one for the initial and one for the final state, gives rise to a possible band. The electronic states, finally, characterized by various values of A and S , or of Q , are separated in energy by differences of the order of those between atomic LS terms. These general features of the array of molecular levels are illustrated for a simple case in Fig. 126.

167. Electronic Bands.—The most general type of transition between molecular states is one in which changes occur in the electronic state of the molecule as well as in its nuclear vibration-rotation state. The spectra hitherto discussed, of rotational or vibrational type, represent special cases in which the electronic state does not change. Such spectra constitute, however, only a small fraction of all known band spectra. When the electronic state does change in a transition, the resulting change in energy is usually so large that the band lies in the visible or ultraviolet region of the spectrum. Such bands may be called *electronic bands*.

In transitions characterized by a given pair of electronic states and by given values v' and v'' of v , representing a fixed pair of vibrational states, various changes of J (and of K , if $K = J$) may occur. The resulting lines form a single *band*. All of the bands due to transitions between a given pair of electronic states, for all possible values of v' and v'' , are said to form a *band system*. Because various electronic jumps are possible, the band spectrum of any molecule consists of many band systems.

The lines in a given electronic band are limited by the selection rules

$$\Delta K = 0 \text{ or } \pm 1, \quad \Delta J = 0 \text{ or } \pm 1.$$

Even aside from the fine structure that exists when the rotational levels

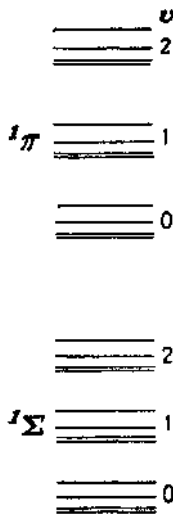


FIG. 126.—Diagram illustrating general arrangement of molecular levels. To make the scale correct, the rotational levels (numbered by $K = J$) should be drawn much closer together, and the ${}^1\Pi$ and ${}^1\Sigma$ groups should lie much farther apart.

are numbered by a quantum number K which is different from J , electronic bands are commonly more complicated than the simple type of vibration-rotation band described above because transitions for $AK = 0$ (or $AJ = 0$) are allowed. Lines for which AK (or AJ if there is no fine structure) has the opposite sign to Δv are said to constitute the *P branch* of the band; those for $AK = 0$ (or $AJ = 0$) constitute the *Q branch*; those for which AK (or AJ) is in the same direction as Δv form the *R branch*. The *Q branch* is frequently missing, however, for special reasons, e.g., in all bands arising from a $2 \rightarrow 2$ electronic transition. When $K = J$, what is regarded as a single band sometimes contains more than one branch of each type.

A good approximate expression for the energies of the molecular levels is often obtained if we employ an expression like that in Eq. (246) for the vibration-rotation part. Omitting the small D_v term, we have then for the molecular energy, in cm^{-1} ,

$$\tilde{W} = \tilde{W}_e + \tilde{W}_v + B_v K(K+1) \cdots \quad (249)$$

where B_v is positive. Thus, for the levels belonging to two different electronic states, we may write

$$\begin{aligned}\tilde{W}' &= \tilde{W}'_e + \tilde{W}'_v + B'_v K(K+1), \\ \tilde{W}'' &= \tilde{W}''_e + \tilde{W}''_v + B''_v K(K+1);\end{aligned}$$

and we then find for the three branches of the band, as a generalization of (247a) and (247b):

$$P: v', K-1; v'', K: \quad \tilde{\nu} = \tilde{\nu}_e + \tilde{\nu}_{v''v'} - (B'_v + B''_{v''})K + (B'_v - B''_{v''})K^2 \quad (250a)$$

$$Q: v', K; v'', K: \quad \tilde{\nu} = \tilde{\nu}_e + \tilde{\nu}_{v''v''} + (B'_v - B''_{v''})K(K+1) \quad (250b)$$

$$R: v', K; v'', K-1: \quad \tilde{\nu} = \tilde{\nu}_e + \tilde{\nu}_{v''v''} + (B'_v + B''_{v''})K + (B'_v - B''_{v''})K^2. \quad (250c)$$

Here K refers in each case to the larger of the two values of K concerned in the transition. If the quantum number K does not exist for the levels in question, K is to be replaced by J in all equations from (249) to (250c). The symbol $\tilde{\nu}_e = (\tilde{W}'_e - \tilde{W}''_e)/h$ represents the frequency that would arise from the electronic transition alone; whereas $\tilde{\nu}_{v''v''} = (\tilde{W}'_v - \tilde{W}''_{v''})/h$, and it is assumed that $v' \geq v''$.

Since the quantities B'_v and $B''_{v''}$ in Eqs. (250a) to (250c) refer to different electronic states, they may be expected to differ considerably. We may suppose the forces between the atoms to be quite different in the two cases; hence, their positions of equilibrium and the values of the moment of inertia of the molecule are also different. This is in contrast to the case of the vibration-rotation bands, where B_v

varies only a little from one level to another. The quadratic terms in (250a,b,c), therefore, will soon make themselves felt. As K increases, the trend of ν in one branch will soon be reversed, in the P branch if $B'_\nu > B''_{\nu''}$, in the R branch if $B'_\nu < B''_{\nu''}$. It is thus a general characteristic of electronic bands that one branch is folded back on itself and on top of the others. At the point in the spectrum where a branch turns back, the lines are crowded together, forming a *band head*. The band appears to shade away from the head; some bands are shaded in this way toward the red, some toward the violet. When a Q branch is present, it may form a second head, although such behavior is not obviously predicted by Eq. (2506).

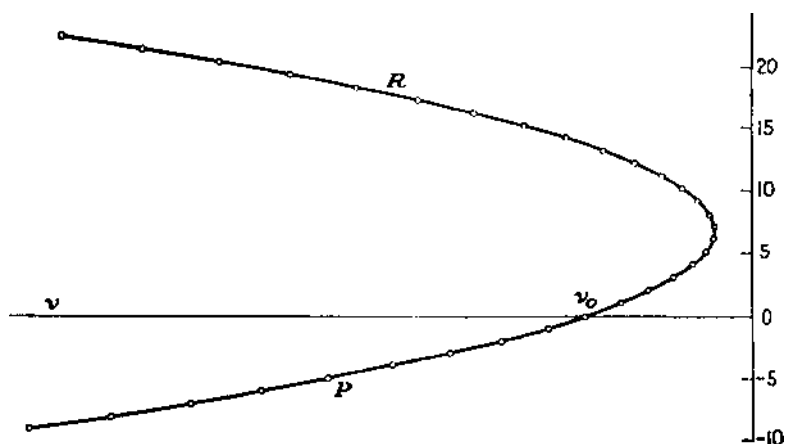


FIG. 127.—Fortrat diagram for the CuH band $\lambda 4,280 \text{ \AA}$. The lines are represented by circles, the ordinate representing the larger J value for the transition, $\nu_0 = \nu' - \nu''$.

A band of this type is often represented graphically by means of a Fortrat diagram, on which each line is represented by a point or circle, the ordinate representing K (or J) and the abscissa, P .

Such a diagram is shown in Fig. 127 for the CuH band $\lambda 4,280 \text{ \AA}$, which has no Q branch; a spectrogram of this band due to Professor R. Mecke is shown in Fig. 128. The band is shaded toward the red from a head which is plainly shown in the spectrogram. The electronic transition is $^1J \rightarrow ^1J$.

The number of different bands in a band system, all arising from the same electronic transition but with various values of ν' and ν'' , may be very large. The change in ν , $\nu' - \nu''$, however, although not confined to ± 1 as for a harmonic oscillator, tends to be restricted to moderate values. A spectrogram showing parts of three such band systems is reproduced in Fig. 129. Extremely low dispersion was used in order to include a wide spectral range, hence each *band* appears

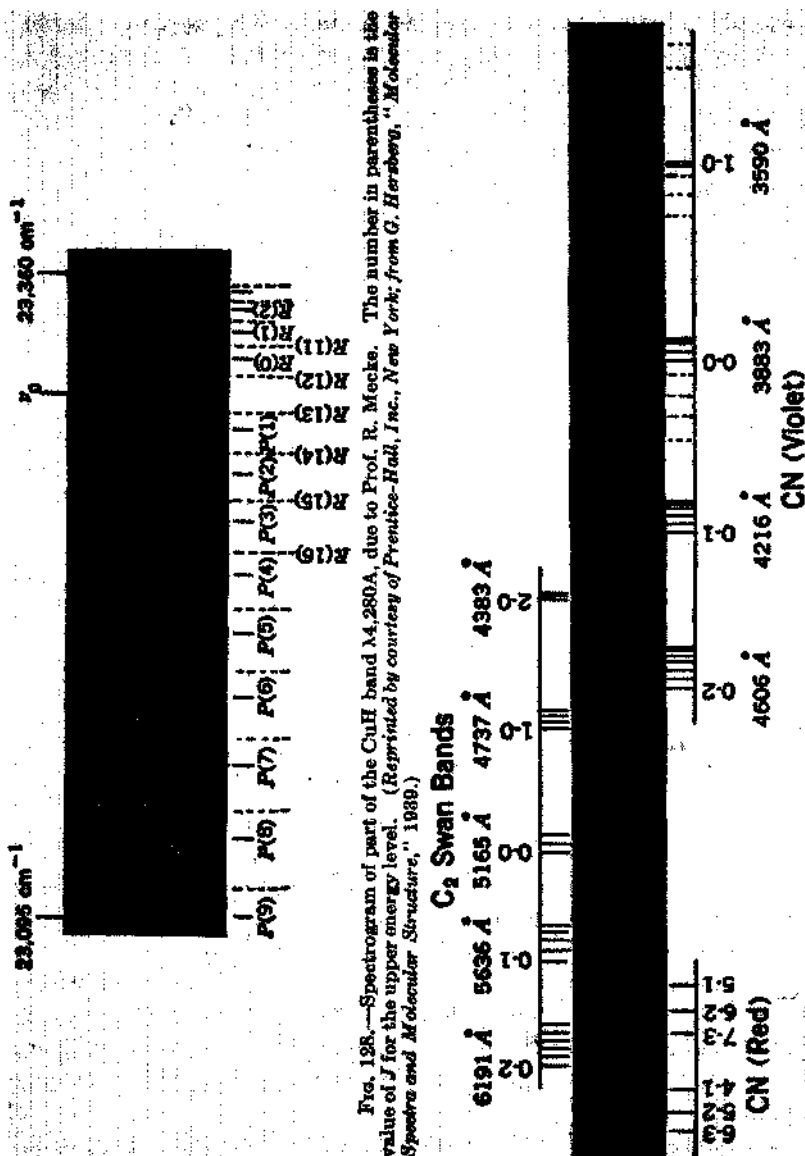


FIG. 128.—Spectrogram of part of the CuH band M4, 280Å, due to Prof. R. Mecke. The number in parentheses is the value of J for the upper energy level. (Reprinted by courtesy of Prentice-Hall, Inc., New York; from G. Herzberg, "Molecular Spectra and Molecular Structure," 1939.)

FIG. 129.—Bands of CN and C₂ (carbon arc in air). Three band systems are shown.

as a single *line* in the figure; the groups that seem to the eye to stand out are groups of bands having in each case the same value of the difference $\nu' - \nu''$. The values of ν' , ν'' (e.g., 6 — 3) are shown for a few bands. The Swan bands of C_2 result from a $^3II \rightarrow ^3II$ electronic transition; the violet CN bands, from a $^2S \rightarrow ^2\Delta$; the red CN bands, from a $^2II \rightarrow ^2S$ transition.

The *rotation* spectrum of a molecule like HC1 (Sec. 154) can be regarded as the 0 — 0 (i.e., $\nu = 0$ to $\nu = 0$) band, and the *vibration-rotation* band (Sec. 155) as the 0 — 1 band, of a band system in which the molecule remains in its normal state. In the halogen hydrides, like HC1, this latter is a 12 state; therefore, $J = K$ and, as already remarked, there is no fine structure. There is also no *Q* branch, and it is partly for this reason that the band exhibits a gap in the center. Furthermore, there is no band head because the *P* and *R* branches do not fold back but form a single straight sequence of lines. Thus, bands of this sort are exceptionally simple in form.

Many other interesting features of band spectra could be discussed if space permitted. We shall mention only one more fact. The vibrational and rotational frequencies are affected by the *masses* of the atoms, so that the band spectra of molecules containing *different isotopes* of the same element are slightly different. Extra lines in band spectra have sometimes led to the discovery of rare isotopes, such as the oxygen isotope of atomic weight 18, present in only about 1/500th as great concentration as O^{16} .¹

For further information about band spectra the student is referred to the treatises mentioned in a footnote to Sec. 153 above or to the literature.

168. The Raman Effect.—As a final topic in spectroscopy brief mention will be made of the Raman effect.

When light passes through a "transparent" substance, solid, liquid, or gaseous, a certain part of the light is scattered in all directions (the Tyndall effect). The most familiar example is the light from a clear sky. Rayleigh ascribed such effects to scattering by the individual molecules, or by groups of molecules much smaller in linear dimension than the wave length of the light. If the incident light is monochromatic, the scattered light is ordinarily observed to be unchanged in frequency, in accordance with Rayleigh's theory.

It was shown in 1925 by Kramers and Heisenberg,² however, that, according to classical electromagnetic theory, if the scattering electrons

¹ GIAUQUE and JOHNSTON, *Nature*, vol. 123, p. 318 (1929); MECKE and CHILDS, *Zeits. f. Physik*, vol. 68, p. 362 (1931).

² KRAMERS and HEISENBERG, *Zeits. f. Physik*, vol. 31, p. 681 (1925).

in an atom or molecule are in motion, the scattered light should contain other frequencies in addition to that of the incident light. Failure to notice this implication of classical theory had been due to the common assumption that the radiating particles were at rest except as disturbed by radiation. In 1928, independently of the theoretical prediction, Raman and Krishnan discovered the phenomenon experimentally,¹ in the course of an extensive study of scattering by liquids and solids. The scattering of light with a change in its frequency has been studied extensively since then and is commonly called the Raman effect.

To observe the Raman effect, the incident light should be monochromatic and very intense. The scattered light is then seen to contain, besides a line of the same frequency as the incident light, several weak lines of other frequencies. If the frequency ν of the incident light is varied, these other lines move along the frequency axis at the same rate, maintaining constant frequency differences from ν and not changing greatly in intensity. In these respects the Raman lines differ sharply in behavior from fluorescent lines, the frequencies of which are fixed by the scattering substance and which flash out only when the incident frequency falls upon an absorption line of the substance. Thus, in the Raman effect, it is *frequency shifts* in the scattered spectrum that are determined by the nature of the scatterer rather than the frequencies themselves.

The explanation of the Raman effect in terms of quantum theory is very simple. When a photon of frequency ν is "scattered" by an atom or molecule whose quantum state is not altered in the process, the scattered photon has the same frequency as had the incident photon. But it may happen that the atom or molecule is changed in the process from a state in which its energy is W_1 to a state of different energy W_2 . If this happens, conservation of energy requires that the frequency ν' of the scattered photon be modified so that

$$h\nu' + W_2 = h\nu + W_1;$$

$$\nu' = \nu + \frac{W_1 - W_2}{h}.$$

In terms of such ideas the Raman effect had been predicted, tentatively, at a much earlier date by Smekal.

In the expression just written for ν' , the term $(W_1 - W_2)/h$ can be interpreted as a frequency ν_{12} that the atom might conceivably emit

¹ RAMAN and KRISHNAN, *Nature*, vol. 121, p. 501 (1928); RAMAN, *Indian J.ourn. Phys.*, vol. 2, p. 387 (1928).

or absorb in the usual way, in jumping from the first state to the second. Thus for the Raman line we may write

$$\nu' = \nu + \nu_{11}.$$

The difference between the frequency of each Raman line and the frequency of the incident light is thus equal to the frequency of some conceivable emission or absorption line of the scattering atom or molecule.

The intensity of a Raman line has nothing to do, however, with the intensity of the emission or absorption line that is thus correlated with it. The selection rules for the two are quite different; transitions that are forbidden in ordinary spectra may occur freely in the Raman effect. This is one reason for its great theoretical interest. According to wave mechanics, *a Raman jump is possible between two atomic or molecular levels A and B only when there exists at least one third level, C, such that ordinary radiative transitions are allowed between A and C and between B and C.* It is almost as if the atom or molecule actually jumped first from A to C and then from C to B. The relative probabilities of the various processes do not correspond to this simple picture, however, nor can it be said that the atom or molecule remains any definite length of time in C, as it does in the production of fluorescence.

By means of these principles the array of lines to be expected in any given case is easily worked out. In the case of the scattering of light by molecules, for example, there may be Raman lines associated with the rotational or vibration-rotation spectra of the molecule, in case it possesses such spectra. In the common type of rotation spectrum, the selection rule is $\Delta J = \pm 1$ (Sec. 154). For the Raman lines associated with this spectrum, therefore, the selection rule will be either $\Delta J = 0$ (i.e., $J \rightarrow J \pm 1$ for a jump from A to C and $J \pm 1 \rightarrow J$ from B to C) or $\Delta J = \pm 2$ {e.g., $J \rightarrow J + 1$ and $J + 1 \rightarrow J + 2$ }. The case, $\Delta J = 0$, however, involves no change in the molecular energy and hence merely contributes to the ordinary or Rayleigh scattering. Thus, effectively, for Raman lines of purely rotational origin, we must have

$$\Delta J = \pm 2.$$

The incident line, as seen in the spectrum of the scattered light, should thus be accompanied on each side by several lines spaced twice as far apart on the frequency scale as the lines of the rotational spectrum. From Eq. (241), it is easily seen that the innermost lines will be distant from the unmodified line 1.5 times as far as the spacing between the Raman lines themselves [the frequency differences of the various lines being proportional to $2(2J - 1)$, with $J > 2$, J referring to the upper level].

Besides these *rotational* Raman lines, there should also appear, at a much greater distance and only on the long-wave side, a band corresponding to the ordinary *vibration-rotation* spectrum. This can arise from the following Raman transitions, connected to third levels by allowed transitions as shown, the value of ν being indicated numerically:

$$0, J \rightarrow \begin{cases} 1, J+1 \\ 1, J-1 \end{cases} \begin{cases} 1, J+2 \\ 1, J-2 \end{cases} \quad 0, J \rightarrow \begin{cases} 0, J+1 \\ 0, J-1 \end{cases} \begin{cases} 1, J+2 \\ 1, J-2 \end{cases}$$

Thus, for this band, the Raman selection rule is

$$\Delta\nu = 1, \quad \Delta J = 0 \text{ or } \pm 2.$$

Lines for which $\Delta J = 0$, involving almost no change in rotational energy, will coalesce into an intense line in the approximate position of the missing central line of frequency $\nu_{\nu\nu'}$ or ν_{10} in the vibration-rotation band. On each side of this intense line, there should then appear other much fainter lines for $\Delta J = \pm 2$ spaced twice as widely as in the rotational spectrum. In case the vibrational level for $\nu = 1$ lies so low that a considerable fraction of the molecules are maintained in this state by thermal agitation, a similar but weaker Raman band should be observed on the short-wave side of the exciting line as well, associated with the transition $\nu = 1 \rightarrow \nu = 0$.

The Raman spectrum as observed in light scattered by gaseous hydrogen halides such as HCl fulfills these predictions. The spectrum of the light scattered by HCl gas when strongly illuminated by light from a mercury arc in glass was studied by Wood and Dieke.¹ They found a line at 4,581.8 Å which they interpreted as the single intense line in a vibration-rotation Raman spectrum excited by the Hg line at 4,047 Å. The frequency difference between these two lines is 2,886.0 cm^{-1} , which is in excellent agreement with the frequency of the missing central line in the 3.5 μ band from HCl , viz., 2,885.4 cm^{-1} [Sec. 155(6)]. Much closer to the exciting line and on both sides of it there were also a number of lines which they interpreted as a rotational Raman spectrum.

The data obtained by Wood and Dieke for the rotational Raman lines excited by another mercury line, λ 4,358 Å, are reproduced in Table XI, the wave lengths λ of the lines being given in angstroms. Under " $\nu - \nu_0$ " is given the separation in wave numbers of each Raman line from the exciting line; under " $J \rightarrow J'$ " are given the initial and

¹ WOOD and DIEKE, *Phys. Rev.*, vol. 35, p. 1355 (1930).

final values of J ; as $(\nu_1 + \nu_2)_{\text{inf}}$ is given the sum of the frequencies of the two rotational lines that connect the two states (*e.g.*, for $1 \rightarrow 3$, the frequencies of the rotational lines for $1 \rightarrow 2$ and $2 \rightarrow 3$ are added together). The rotational frequencies are taken from Table X in Sec. 154; where observed values are missing, calculated values taken from the last column of Table X are used or others calculated from the same formula there employed. The agreement between $\nu - \nu_0$

TABLE XI.—ROTATIONAL RAMAN SPECTRUM FOR HC1. EXCITED BY X 4,358 Å

λ	$\tilde{\nu} - \tilde{\nu}_0$	$J \rightarrow J'$	$(\tilde{\nu}_1 + \tilde{\nu}_2)_{\text{inf}}$	λ	$\tilde{\nu} - \tilde{\nu}_0$	$J \rightarrow J'$	$(\tilde{\nu}_1 + \tilde{\nu}_2)_{\text{inf}}$
4,314.7	143.8	4 $\rightarrow$ 2	145.4			0 $\rightarrow$ 2	62.4
				4,377.6	101.1	1 $\rightarrow$ 3	103.9
23.8	183.3	5 $\rightarrow$ 3	186.8				
				85.6	142.7	2 $\rightarrow$ 4	145.4
31.2	232.2	6 $\rightarrow$ 4	228.0				
				94.2	187.5	3 $\rightarrow$ 5	186.8
				4,402.4	229.4	4 $\rightarrow$ 6	228.0
				10.4	271.0	5 $\rightarrow$ 7	269.3
				18.6	312.9	6 $\rightarrow$ 8	310.5
				26.5	353.0	7 $\rightarrow$ 9	351.4

and $(\nu_1 + \nu_2)_{\text{inf}}$ is sufficiently good, in view of the uncertainty of the infrared data, to constitute strong confirmation of the theory as to the origin of the Raman lines.

Many Raman spectra have been observed.¹ Commonly they are more complicated than the very simple type exhibited by HC1. Even homonuclear molecules, such as O_2 and N_2 , give Raman spectra corresponding to vibration-rotation and rotational spectra, although the latter spectra cannot be observed directly [Sec. 155(g)]. The Raman effect thus constitutes a valuable tool in the study of molecular quantum states.

¹ For a more extensive discussion see Glockler, *Rev. Modern Phys.*, vol. 15, p. 111 (1943).

CHAPTER IX

THE QUANTUM THEORY OF SPECIFIC HEATS

In the last few chapters, we have discussed the application of quantum theory to the processes of the emission and absorption of radiation. The next important application of quantum theory after the publication of Bohr's first paper was made in a different field, in relation to the theory of specific heat. The success of the theory in this field demonstrated beyond doubt the wide range of applicability of the new ideas. In this chapter we shall discuss very briefly the applications of quantum ideas to the theory of specific heats, mainly with the purpose of pointing out how the newer theory succeeded where the old one had largely failed.

For the benefit of students who are not at the moment interested in specific heats, it may be remarked that none of the material in this chapter is necessary for the understanding of later chapters.

159. Variation of Specific Heats of Solids with Temperature.—Over 100 years ago, Dulong and Petit, as a result of the determination of the specific heats of a number of elements, such as iron, lead, silver, gold, concluded that "the product of the atomic weight and the specific heat is the same for all elementary (solid) substances." This law is known as the "law of Dulong and Petit." The product, atomic weight by specific heat, is known as the *atomic heat*.

The law of Dulong and Petit represents at best only a rough approximation, and there are some notable exceptions. The atomic heats in terms of joules are given for 63 elements in crystalline form in the "International Critical Tables." Converted into calories, the atomic heats of 58 of these elements average 6.15 cal. per gram-atom; but they exhibit considerable variation, ranging from 5.38 to 6.93. The atomic heats of the remaining 5 of the 63 elements are in complete disagreement with the law: boron, 3.34; beryllium, 3.85; carbon (diamond), 1.46; hydrogen (solid), 0.57; silicon, 4.95. These 5, with the exception of hydrogen, are light elements having high melting points: boron, 2300; beryllium, 1350; silicon, 1420°C. The specific heats are taken at ordinary temperatures wherever possible.

The exceptions to the law of Dulong and Petit were the subject of extended research, and it was early discovered that at least one reason for the failure of the law in these cases was the fact that the

specific heat varies with temperature. Thus, in 1872, Weber observed that the specific heat of diamond (i.e., carbon) increases threefold between 0 and 200°C. The study of specific heats of substances as a function of temperature has been greatly facilitated in recent years by the availability of liquefied gases for making measurements at low temperatures. It has been found that below a certain temperature characteristic of each substance the specific heat of any solid decreases rapidly with decreasing temperature, ultimately reaching a value almost zero. Figure 130 shows the variation with temperature of the

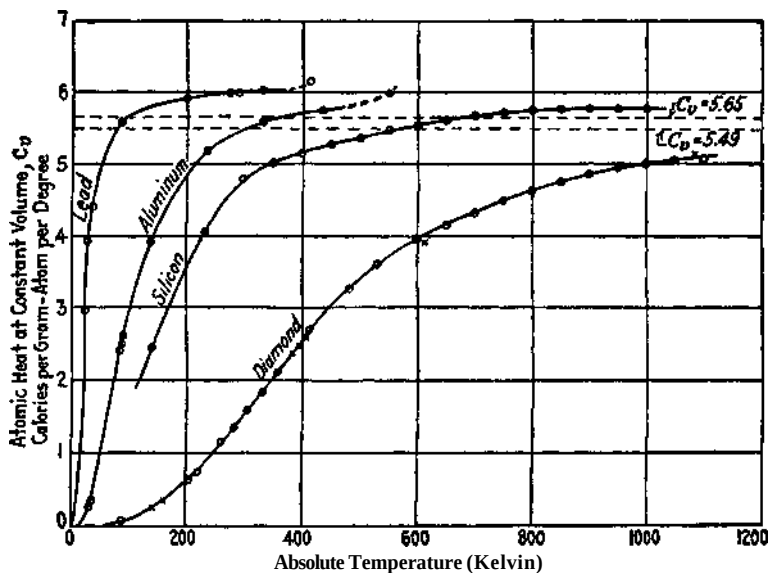


FIG. 130.—Variation of atomic heat at constant volume with temperature.

atomic heats at constant volume of four characteristic substances: diamond, silicon, aluminum, and lead. The atomic heat of lead at ordinary temperatures is observed to approximate the value required by the law of Dulong and Petit, but below a temperature of about 100°K. it drops very rapidly toward zero. The curves for the other three substances have the same general form as that for lead. They differ mainly (1) in the temperature above which they seem to obey the law of Dulong and Petit and (2) in the rate at which the atomic heat decreases below that temperature. In the case of diamond, the curve suggests that a value of about 6 cal. per gram-atom per degree would not be reached until the temperature exceeded 2000°K.

The curves for nearly all substances are similar to those shown in Fig. 130; they lie between the curves **for lead** on the one side and for

diamond on the other, very commonly even between those for aluminum and for lead. Since most of the measurements on which the law of Dulong and Petit was based were made at *room* temperature, it is obvious why the law was so nearly obeyed.

The curves for different substances are not only similar in appearance but can be brought (almost) into coincidence by suitable change in the temperature scale for each curve. Thus, if the abscissas of the curve for aluminum be multiplied by a factor of 4.6, this curve will nearly coincide with that for diamond. The crosses (X) on the curve for diamond represent the points from the aluminum curve shifted in the way described. This fact indicates that the curves are closely related to each other by some fundamental underlying law. What is this fundamental law which governs the variation of specific heat with temperature and which, at sufficiently high temperatures gives the law of Dulong and Petit? Classical physics failed to find the law. The quantum theory was more successful.

160. Classical Theory of the Specific Heats of Solids.—According to classical conceptions, the atoms of a solid element at the absolute zero of temperature would be at rest in positions of equilibrium under the action of their mutual attractions and repulsions. Let us consider the energy of the solids in this state to be zero. When the temperature is raised, the atoms will be set into vibration about their positions of equilibrium; but as long as the amplitude of vibration is not too large, the restoring force will be proportional to the displacement and the vibrations will be of the simple harmonic type.

Each atom of the solid, like a monatomic gas molecule, has obviously 3 degrees of freedom. According to the principle of the equipartition of energy, discussed in Sees. 78 and 79 of Chap. V, when the solid is at an absolute temperature T , each degree of freedom will have an average amount of energy given by

$$E = kT,$$

k being Boltzmann's constant. That is, for each degree of freedom there is average kinetic energy equal to $\frac{1}{2}kT$, and, because the motion is simple harmonic, there is also potential energy averaging $\frac{1}{2}kT$ as well. The total average energy per atom is thus $3kT$. The total thermal energy of agitation of a gram-atom of the solid, containing N_0 atoms, is, therefore,

$$W_A = 3N_0kT = 3RT, \quad (251)$$

since $N_0k = R$, the gas constant for a gram-atom.

Now the specific heat of a substance is equal to the energy required to raise the temperature of 1 gram 1 degree; hence, the atomic heat is equal to the energy required to raise the temperature of 1 gram-atom 1 degree. If, during this rise in temperature, the substance is not allowed to expand, all of the added energy goes to increase the energy of vibration of the atoms. Calling the value of the atomic heat at constant volume, measured in mechanical units, C_v , it follows that

$$C_v = \frac{dW_A}{dT} = 3R. \quad (252)$$

According to classical theory, then, the atomic heat at constant volume of a monatomic solid should be (1) constant and *independent of temperature* and (2) equal to three times the gas constant R . Since the numerical value of R is 1.9856 cal. per gram-atom per degree, it follows that the numerical value of C_v , on the classical theory of the equipartition of energy, should be

$$C_v = 5.96 \text{ calories per gram-atom per degree.}$$

The values ordinarily given for the specific heat of a solid refer, to be sure, to the specific heat under constant pressure. The difference, however, is at most a few percent.¹ Hence, to a first approximation, the value just found for C_v is the numerical value of the atomic heat as required by the law of Dulong and Petit.

The value of C_v given in Eq. (252) is obviously very close to the value which the curves in Fig. 130 approach asymptotically as the temperature increases. But Eq. (252) predicts no variation in atomic heat with temperature, as is exhibited by these experimental curves.

161. Einstein's Theory of the Atomic Heats of Solids.—An important advance in the theory of the specific heats of solids was made by Einstein,² who applied the quantum theory to the vibrations which atoms were assumed to execute about their positions of equilibrium. Let it be assumed (1) that the atoms of a monatomic solid vibrate with a frequency ν , the same for all atoms, which depends on the mass of the atom and on the restoring forces brought into play when the atom is displaced, and (2) that the average energy associated with each degree of freedom is that given by the quantum theory for a harmonic oscillator, or, by Eq. (116) in Sec. 83,

$$\bar{U} = \frac{h\nu}{e^{h\nu/kT} - 1}.$$

¹See any book on Heat, e.g., J. M. Cork, "Heat," p. 233, 1933; J. K. Roberts, "Heat and Thermodynamics," p. 157, 1933.

²Ann. d. Physik, vol. 22, p. 180 (1907).

Since each atom has 3 degrees of freedom, the average energy per atom E_a is, then,

$$\bar{E}_a = \frac{3n\nu}{e^{h\nu/kT} - 1},$$

and the total energy W_A of the N_0 atoms in a gram-atom is

$$W_A = N_0 \bar{E}_a = \frac{3N_0 h\nu}{e^{h\nu/kT} - 1}. \quad (253)$$

As before, the atomic heat at constant volume is obtained by differentiating W_A with respect to T . Therefore,

$$C_v = \frac{dW_A}{dT} = 3N_0 h\nu \frac{1}{(e^{h\nu/kT} - 1)^2} e^{h\nu/kT} \frac{h\nu}{kT^2},$$

or, after rearranging and remembering that $N_0 k = R$,

$$C_v = 3R \left[\frac{e^{h\nu/kT}}{(e^{h\nu/kT} - 1)^2} \left(\frac{h\nu}{kT} \right)^2 \right].$$

This is Einstein's equation, based upon the older quantum theory, for the atomic heat of a solid at constant volume. It differs from Eq. (252) in that the term $3R$ is multiplied by the quantity within the square bracket, which is a function of temperature. Thus according to the quantum theory C_v is a function of temperature.

Now the experimental curves in Fig. 130 show three main characteristics: (1) they have essentially the same form, as we have seen, for all substances; (2) at low temperatures, the atomic heat approaches zero; (3) at high temperatures, *i.e.*, sufficiently high for each substance, the atomic heat approaches the constant value $3R$.

Einstein's equation, (254), agrees with experiment as regards these three characteristics. " (1) According to Eq. (254), the atomic heats of different substances will differ only because of differences in the characteristic frequency ν . At corresponding temperatures, such that the values of the ratio ν/T are the same, the value of C_v will be the same for all substances. (2) If $T \rightarrow 0$, the quantity $h\nu/kT$ becomes infinite, and it may readily be shown that the square bracket in Eq. (254) becomes zero. Einstein's equation, therefore, predicts correctly that the atomic heat should become zero at absolute zero. (3) If kT becomes very large with respect to the product $h\nu$, $h\nu/kT$ approaches zero, and the square bracket approaches unity. At high temperatures, therefore, the atomic heat should approach $3R$, in agreement with experiment and with the law of Dulong and Petit.

For intermediate temperatures, Eq. (254) predicts *qualitatively* the correct variation of atomic heat with temperature, but it yields too low values at low temperatures. This is shown in Fig. 133 below, in which the lower curve is for the atomic heat of aluminum as given by Einstein's formula, whereas the circles are the observed values of C used in plotting the curve for aluminum in Fig. 130. The frequency ν in Einstein's formula was chosen so as best to fit the data. In spite of the disagreement, really comparatively small, between theory and experiment, it is remarkable that Einstein's simple theory should predict values of C_ν so nearly correct.

162. Characteristic Frequencies.—Einstein's formula for the atomic heat of a solid, given in Eq. (254), contains, in the frequency ν , a single constant characteristic of a given substance. Experimental values of this constant can be obtained from the atomic-heat data by choosing ν so as to obtain the best fit between Einstein's formula and the data. But several independent methods of estimating ν have also been proposed. They lead to values in substantial agreement with those computed from the specific-heat curves.

(a) *From the "Reststrahlen" or "Residual Rays."*—One of the most convincing of these methods is that due to Rubens¹ and collaborators, in which they used the phenomenon of selective reflection for isolating long waves. The reflectivity² R of a substance for radiation of wave length λ depends on both the coefficient of absorption of the substance and its refractive index n at that wave length, the value of R being given for normal incidence by³

$$R = \frac{(1 - n)^2 + n^2 \kappa^2}{(1 + n)^2 + n^2 \kappa^2};$$

here $\kappa = \lambda \mu / 4\pi$, being the coefficient of absorption for radiation of wave length λ in the material.⁴ Any substance "which exhibits *selective absorption* in any spectral region presumably has resonating mechanisms the natural periods of which coincide with the central part of the absorption band. In those regions, the value of K and, therefore, of n may become large enough compared with $(n - 1)$ to raise the value of the reflectivity R considerably above its value for neighboring

¹ RUBENS and NICHOLS, *Ann. d. Physik*, vol. 60, p. 418 (1897); RUBENS and KURLBAUM, *Ann. d. Physik*, vol. 4, p. 649 (1901); RUBENS and HOLLNAGEL, *Phil. Mag.*, vol. 19, p. 761 (1910); RUBENS and VON WARTENBERG, *Konigl. Preuss. Akad., Ber.*, p. 169 (1914).

* For definition, see Sec. 70(e).

* See WOOD, R. W., "Physical Optics."

⁴ Thus M is not the "absorptivity" as defined in Sec. 70(d). The fraction of the radiant energy that is absorbed in a short distance Δx is $\mu \Delta x$.

spectral regions. The substance then exhibits *selective reflection*. Confirmation of this relation between selective absorption and selective reflection is found in the work of Nichols and of Rubens on quartz. Quartz is transparent, or nearly so, up to about 7.5M but has two very strong absorption bands near together at about 8.5u. At 8.5u, it is found to reflect 80 percent, while at 4u, where it is quite transparent, it reflects only a few percent.

If a continuous spectrum is reflected from a substance-which shows selective reflection at a wave length X_0 , the reflected beam will be relatively much richer in radiation of that wave length than was the incident beam. After several such reflections, the radiation may be comprised almost entirely of the wave lengths which are selectively reflected, the remainder having been almost completely absorbed. The radiation remaining after several such reflections is called "residual rays" or "reststrahlen."

By this method, Rubens and his collaborators succeeded in isolating the residual rays from a number of substances. A few of their results are shown in Table I, in comparison with values of ν calculated by means of Einstein's formula from the curves for the specific heats of the same substances. The two sets of frequencies are in fair agreement with each other.

(b) *From Compressibilities*.—The frequency of vibration of the atoms, in Einstein's theory, will depend upon the rate at which a restoring force comes into existence as the atoms are displaced from their equilibrium positions. Now the atoms can also be displaced relatively to one other by *corn-pressing* the substance; and, when this is done, the ratio of the stress thereby called into existence to the displacement determines the compressibility of the substance. A relation might be expected to exist, therefore, between the characteristic frequency of Einstein's theory and the compressibility. Einstein developed a tentative theory of this relation¹ and used it to calculate values of ν from compressibilities. Two values of ν thus calculated are given in Table I.

(c) *From Melting Points*.—This method of computing frequencies is due to Lindemann.² It is based on the assumption that the melting point T_m of a solid is the temperature at which the amplitude of vibration of the atoms is equal to ("of the same order of magnitude as" would be preferable) the average distance apart d of the atoms. When an atom is executing a simple harmonic motion with amplitude d and frequency ν , we can write for its displacement

¹ *Awn. d. Physik*, vol. 34, pp. 170, 590 (1911).

² *Phys. Zeits.*, vol. 11, p. 609 (1910).

$$s = d \sin 2\pi \nu t.$$

Its energy is entirely kinetic at the end of a swing, when its velocity is

$$v = \left(\frac{ds}{dt} \right)_{\max} = 2\pi \nu d (\cos 2\pi \nu t)_{\max} = 2\pi \nu d;$$

hence, its energy of vibration is

$$\frac{1}{2} m v^2 = 2\pi^2 m \nu^2 d^2,$$

m being its mass. But by the principle of the equipartition of energy, cited in Sec. 160, combined with Lindemann's assumptions, the energy of the atom also equals $3kT_m$, where T_m is the melting point. Hence,

$$2\pi^2 m \nu^2 d^2 = 3kT_m.$$

Multiply this equation through by N_0 , the number of atoms in a gram-atom, and then substitute $N_0 m = A$, the atomic weight, and $kN_0 = R$, the gas constant for a gram-atom, and we obtain

$$2\pi^2 \nu^2 d^2 A = 3RT_m.$$

We can estimate d as follows. If V_A is the volume occupied by a gram-atom of the substance at temperature T_m , V_A/N_0 is the average volume per atom; supposing this volume to be cubical in form and hence equal, according to Lindemann's assumption, to d^3 , we have

$$d = \left(\frac{V_A}{N_0} \right)^{1/3}.$$

Substituting this value of d in the last equation and solving for ν , we find

$$\nu = C \sqrt{\frac{T_m}{A V_A^{1/3}}}, \quad C = \frac{1}{\pi} N_0^{1/4} \sqrt{\frac{3R}{2}}.$$

TABLE I.—COMPARISON OF CHARACTERISTIC FREQUENCIES

Substance	From residual rays		From compressibilities	From melting points	From Einstein's formula
	λ	ν^*	ν^*		
NaCl.	52.0 μ	5.77	...	...	4.6
KCl.	63.4	4.73	...	...	3.7
KBr.	82.6	3.64	...	...	3.2
Aluminum.	...	...	6.6	7.6	6.42
Silicon.	...	...	...	9.6	11.9
Lead.	...	...	2.2	1.8	1.61
Diamond (extrapolated).	...	...	..	32.5(?)	30.8

* All frequencies are to be multiplied by 10^{13} sec.⁻¹.

Some values of ν computed from observed melting points by means of this formula are shown in Table I.

The fact that values of ν computed by these diverse methods agree roughly with one another and with those calculated by means of an application of Einstein's formula to the specific heats indicates that there must be a substantial measure of truth in the assumptions underlying all of these theories. Yet it can scarcely be said that any one of these methods of determining the value of ν to be substituted in Einstein's equation leads to a very satisfactory theory of the specific heat.

163. Debye's Theory of Atomic Heats.—An empirical improvement was effected in Einstein's formula by Nernst and Lindemann, who represented the specific heat as the sum of two terms of the Einstein type, each containing a different frequency. This was a step in the right direction. A much more successful theory was developed, however, by Debye, who attacked the problem from a different angle.¹

Let us postulate, for the moment, a solid composed of atoms that have no thermal vibrations but are at rest in their respective positions of equilibrium. Let a system of standing waves, say longitudinal, be set up in the solid. Each atom is then vibrating with an amplitude that depends on the position of the atom with respect to the nodes and loops of the wave system. If we superpose more and more standing-wave systems of both the same frequency as the original and of different frequencies, the vibrations of any particular atom become more and more complex, until, finally, we may approach a condition of atomic agitation similar to the temperature vibrations. Conversely, we may think of the temperature vibrations of the atoms of a solid as being equivalent to a vast complex of standing waves of a great range of frequencies.

The set of standing waves thus imagined is similar to the standing waves which we used in Sec. 80 to represent the radiation field inside an enclosure, but it differs in two respects. In the first place, the range of frequencies here certainly does not extend to infinity. For each set of standing waves constitutes a degree of freedom of the solid, and the total number of degrees of freedom is only $3N$, where N is the number of atoms in the solid. There exists, therefore, an upper limit to the frequency of the standing waves. The wave length corresponding to the highest frequency should be of the order of magnitude of the distance between atoms.

¹ *Ann. d. Physik*, vol. 39, p. 789 (1912).

In the second place, two different types of waves are possible in a solid: (1) longitudinal waves, velocity v_L , and (2) transverse waves, velocity v_T . These velocities are determined by the elastic constants of the material and its density.

For the number of degrees of freedom per unit volume dn_L of longitudinal waves in the wave-length range $d\lambda$, we obtained in Sec. 80, Eq. (107):

$$dn_L = \frac{4\pi d\lambda}{\lambda^4}.$$

The corresponding number dn_T for the transverse waves in the solid will be the same as for transverse electromagnetic waves as given by Eq. (108) in Sec. 80;

$$dn_T = \frac{8\pi d\lambda}{\lambda^4}.$$

Changing wave length λ to frequency ν by the relations $\lambda = v/\nu$ and $|d\lambda| = (v/\nu^2)d\nu$, these equations become

$$dn_L = 4\pi \frac{\nu^2 d\nu}{v_L^3}, \quad dn_T = 8\pi \frac{\nu^2 d\nu}{v_T^3}.$$

The total number of degrees of freedom dn in unit volume of the solid in the frequency range ν to $\nu + d\nu$ is the sum, $dn_L + dn_T$, since both systems of waves are simultaneously present. Therefore,

$$dn = dn_L + dn_T = 4\pi \left(\frac{1}{v_L^3} + \frac{2}{v_T^3} \right) \nu^2 d\nu$$

Now let us assume that with each of these degrees of freedom of frequency ν there is associated an average amount of energy ϵ equal to

$$\epsilon = \frac{h\nu}{e^{h\nu/kT} - 1}$$

as stated in Eq. (116) in Sec. 83. The total energy dW_A in this frequency range ν to $\nu + d\nu$ in a volume V_A equal to the volume of a gram-atom is then

$$dW_A = V_A \epsilon dn = 4\pi V_A \left(\frac{1}{v_L^3} + \frac{2}{v_T^3} \right) \frac{h\nu}{e^{h\nu/kT} - 1} \nu^2 d\nu,$$

and the total energy W_A of the gram-atom, taking into account all frequencies, is

$$W_A = \int dW_A = 4\pi V_A \left(\frac{1}{v_L^3} + \frac{2}{v_T^3} \right) \int_0^{\nu_m} \frac{h\nu}{e^{h\nu/kT} - 1} \nu^2 d\nu, \quad (250)$$

where ν_m is the maximum frequency of the standing waves. We may take the lower limit to be 0 because the fundamental frequency of any solid body of experimental dimensions is very small relative to the frequencies of most of the standing waves.

An approximate value of ν_m was obtained by Debye from the condition that the total number of degrees per gram-atom must be $3N_0$, where N_0 is the number of atoms in a gram-atom. The number of the degrees of freedom can also be computed by integrating the expressions for W_A in Eq. (255) from $\nu=0$ to ν_m . Hence,

$$3N_0 = 4\pi V_A \left(\frac{1}{v_L^3} + \frac{2}{v_T^3} \right) \int_0^{\nu_m} \nu^2 d\nu = \frac{4\pi V_A}{3} \left(\frac{1}{v_L^3} + \frac{2}{v_T^3} \right) \nu_m^3, \quad (257)$$

Since the velocities are computable from the elastic constants, ν_m can be computed from this equation. By means of Eq. (257), we can also write for W_A , from Eq. (256),

$$W_A = \frac{9N_0}{\nu_m^3} \int_0^{\nu_m} \frac{h\nu}{e^{h\nu/kT} - 1} \nu^2 d\nu. \quad (258)$$

This equation may be compared with Eq. (253) in Einstein's theory.

To bring out the qualitative features of the result, and for purposes of computation, it is convenient to change the variable of integration in Eq. (258) from ν to

$$x = \frac{h\nu}{kT}; \quad \therefore \nu = \frac{kT}{h}x, \quad d\nu = \frac{kT}{h}dx.$$

The upper limit for x is then $h\nu_m/kT$. Since $h\nu$ and kT both have the dimensions of energy, x is dimensionless; hence, $h\nu_m/k$ must have the dimensions of a temperature. It is convenient to call a temperature T_c equal to $h\nu_m/k$ the *characteristic temperature* for the substance under discussion, so that

$$T_c = \frac{h\nu_m}{k}, \quad \therefore \nu_m = \frac{kT_c}{h}. \quad (259a,b)$$

Let us also introduce $R = N_0k$, the gas constant for a gram-atom. With these substitutions, Eq. (258) takes the form

$$W_A = 9R \frac{T_c^3}{T^3} \int_0^{T_c/T} \frac{x^3}{e^x - 1} dx.$$

To obtain C_v , we differentiate W_A with respect to T , according to the rules for the derivative of a product, remembering that in the second factor, the integral, T appears only in the limit of the integra-

tion. Therefore,

$$C_v = \frac{dW_A}{dT} = 9R \left[4 \frac{T_c^3}{T_c^3} \int_0^{T_c/T} \frac{x^3}{e^x - 1} dx - \frac{T_c^4}{T_c^3 e^{T_c/T} - 1} \left(\frac{T_c}{T} \right) \right]$$

or

$$C_v = 9R \left[4 \left(\frac{T}{T_c} \right)^3 \int_0^{T_c/T} \frac{x^3}{e^x - 1} dx - \frac{T_c}{T} \frac{1}{e^{T_c/T} - 1} \right]. \quad (261)$$

This is Debye's equation for the specific heat of a solid at constant volume. Like Einstein's equation, (254), it contains, besides the universal constant R , just one additional constant, here T_c , referring to the particular substance under consideration.

164. Experimental Test of Debye's Equation.—1. At *high temperatures* Debye's equation should yield the classical value of C_v , viz., $3R$. For large values of T , T_c/T becomes small, and we can write, since $e^x = 1 + x + \dots$, dropping all higher terms in the series,

$$e^{T_c/T} - 1 = \frac{T_c}{T}.$$

Similarly, in the integral in (261), x remains small throughout the range of integration and we can write, replacing $e^x - 1$ by x ,

$$\int_0^{T_c/T} \frac{x^3}{e^x - 1} dx = \int_0^{T_c/T} x^2 dx = \frac{1}{3} \left(\frac{T_c}{T} \right)^3.$$

Hence, by (261), approximately,

$$C_v = 9R \left(\frac{4}{3} - 1 \right) = 3R,$$

as we had expected.

2. At *low temperatures* T_c/T is large. Hence, the second term in the square bracket in Eq. (261) becomes negligible and we may replace T_c/T in the upper limit of the integral by infinity. As stated in Sec. 85,

$$\int_0^\infty \frac{x^3}{e^x - 1} dx = \frac{\pi^4}{15}.$$

Hence Eq. (261) becomes, approximately,

$$C_v = 9R \left(4 \frac{T_c^3 \pi^4}{T_c^3 15} \right) = \frac{12}{5} \pi^4 R \frac{T_c^3}{T_c^3} = \frac{234R}{T_c^3} T_c^3. \quad (262)$$

At temperatures much below its characteristic temperature, therefore, the atomic heat of a solid should be proportional to the *cube* of the absolute temperature. This is **confirmed** by experiment over a

considerable range of temperatures. Schrodinger¹ quotes extensive data bearing on this point; curves for several of the substances given

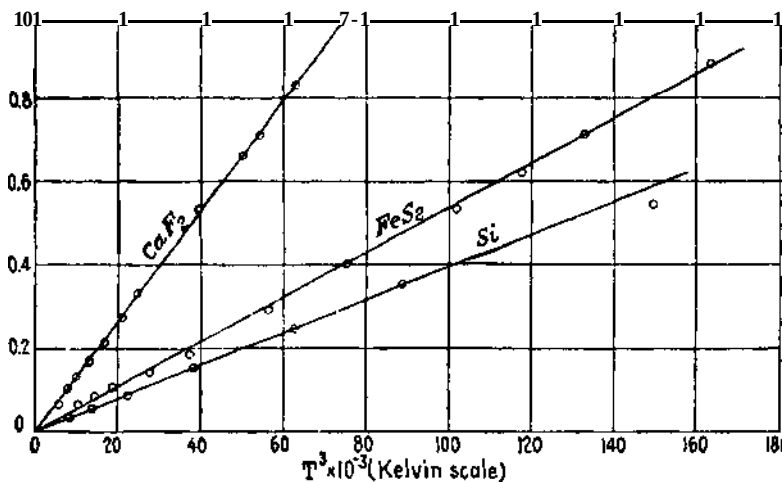


FIG. 4.31.—At low temperatures, the molecular heat at constant volume is proportional to the cube of the absolute temperature.

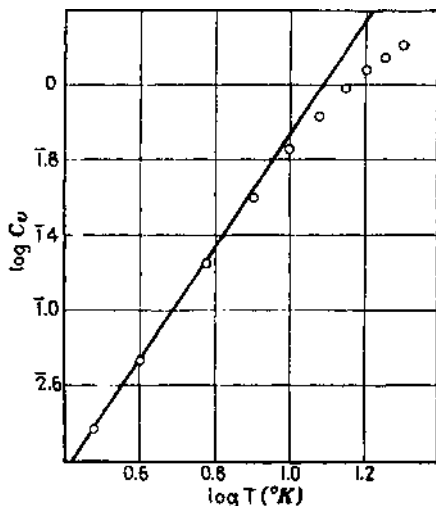


FIG. 132.—Variation of C_v of lead with temperature from 3°K. to 20°K. from measurements of Keesom and Van den Ende.

in his table are shown graphically in Fig. 131, where C_{vf} denoting the specific heat of a gram-molecule, is plotted against the cube of the absolute temperature for CaF_2 (17 to 40°K), FeS_2 (22 to 57°K), and Si (20 to 53°K.). A plot of C_v against $\log T$ for lead, at much

¹Phys. Zeits., vol. 20, p. 497 (1919).

lower temperatures, is shown in Fig. 132.¹ The predicted proportionality is seen to hold from almost vanishingly small values of C_v well up toward $C_v = 1$. Beyond this point, as theory requires, the curves become increasingly concave toward the axis for T^3 or $\log T$.

3. For *intermediate temperatures*, it is necessary to evaluate Debye's formula by the summation of series. The results of such computation

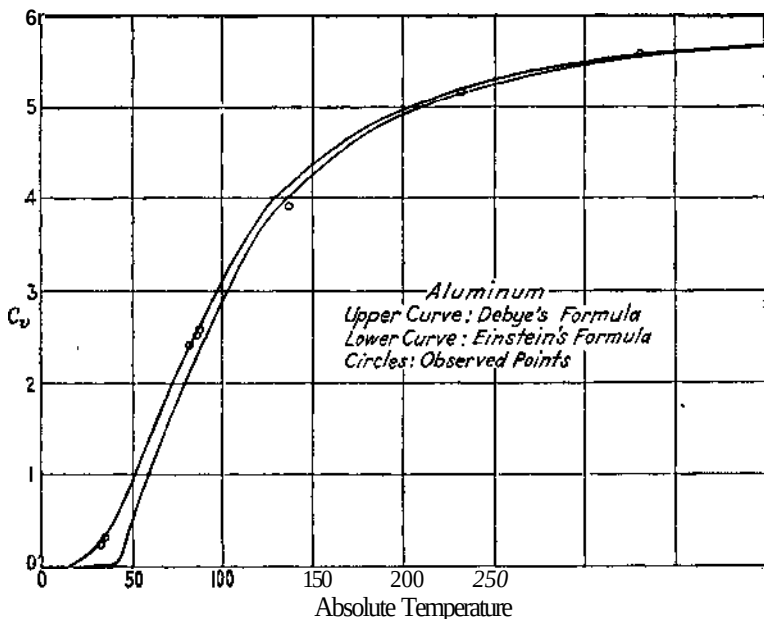


FIG. 133.—Comparison of specific-heat formulas with experiment.

are now available in the form of tables² giving C_v as a function of the quantity T_c/T .

At a temperature $T = T_c$, it is found that

$$C_v = 2.856 R = 5.670 \text{ cal. mole}^{-1} \text{ deg.}^{-1}$$

instead of the classical value, $C_v = 3R$. Thus the critical temperature of a solid, in Debye's theory, might be defined as that temperature at which the atomic heat equals $2.856 R$.

The upper curve of Fig. 133 is the graph of Debye's formula for aluminum, T_c being taken equal to 398° in order to secure the best fit with the data. The agreement between the graph and the observa-

¹ KEESOM and VAN DEN ENDE, *K. Akad. Amsterdam, Proc.*, vol. 33, p. 243 (1930).

² NBBNST, "The New Heat Theorem/" pp. 246-254, 1924; BEATTIE, *J. Math. Phys. (Mass. Inst. Tech.)*, vol. 6, p. 1 (1926).

tions is seen to be excellent. Similar agreement is found for many other substances.¹

Strictly speaking, however, the critical temperature is not an adjustable constant in the Debye theory but is supposed to be given by Eqs. (259a) and (257) in terms of universal constants and the elastic constants of the material. Values of the limiting frequency ν_m for several metals, calculated from the equivalent of Eq. (257) by Allen,² are given in Table II together with values of T_c , labeled "calculated," calculated from these values of ν_m by means of Eq. (259a). For comparison, the values of T_c that give the best fit between Debye's theory and the data³ are added as " T_c , observed." The agreement between the two sets of values of T_c is very good, in view of the approximations made in the theory.

TABLE II.—LIMITING FREQUENCIES ν_m AND CHARACTERISTIC TEMPERATURES T_c

Element	ν_m	T_c , calculated, °K.	T_c , observed, °K.
Al	$8.26 \times 10^{12} \text{ sec.}^{-1}$	395	398
Fe	9.67	463	453
Cu	6.81	326	315
Ag	4.39	210	215
Pb	1.49	71	88

It is an interesting fact that values of ν_m calculated in this manner from the elastic constants are also of the same order of magnitude as the "characteristic frequencies" obtained in other ways (Sec. 162).

165. Molecular Heats of Mixed Solids.—Excellent as it often is in so far as agreement with experiment is concerned, Debye's theory does not give a complete account of the specific heat of solids. A more complete theory for *crystalline* solids was developed by Born.⁴

A *crystal* can be regarded as composed of many similar groups or cells of atoms, the groups being spaced uniformly in any given direction. As a result of the mathematical analysis it may be stated that the theory of Debye can be used as a good representation of the motion of the groups as wholes. In addition to this motion, however, there are a finite number of other modes of vibration which can be regarded as modes of vibration of the atoms within each group. For each of

¹Cf. FOWLEK, R. H., "Statistical Mechanics," 2d ed., p. 125, 1936; LEWIS, W. M. C., "Quantum Theory."

²Roy. Soc. Proc, vol. 94, p. 100 (1918).

³SCHRSDINGER, Phys. Zeits., vol. 20, p. 450 (1919).

⁴"Atomtheorie des festen Zustandes," Encyclopadie der mathematischen Wissenschaften, vol. V, part 3, no. 25, 1923; cf. FOWLER, *op. cit.*, p. 119, 1936.

these modes a term in the energy of the Einstein form, as in Eq. (253) above, can be used. The result is a formula combining the expressions given for C_v by the theories of Einstein and Debye, as represented by Eq. (254) and by the equations from which Eq. (261) was obtained:

$$C_v = 9R \frac{d}{dT} \left(\frac{T^4}{T_c^3} \int_0^{T/T_c} \frac{x^3 dx}{e^x - 1} \right) + 3R \sum_{i=1}^n \left(\frac{h\nu_i}{kT} \right)^2 \frac{e^{h\nu_i/kT}}{(e^{h\nu_i/kT} - 1)^2}, \quad (263)$$

n being the number of internal modes of vibration of the atoms within a cell. A formula of this type had been suggested empirically by Nernst. At low temperatures the Einstein terms disappear and (263) becomes equivalent to Eq. (261).

By a judicious choice of T_c and of the ν_i 's, this last formula can be made to fit the experimental curve very closely in many cases where the simple Debye theory by itself does not succeed well. It might be expected that the frequencies of the residual rays would occur among the ν_i 's. Data confirming this conclusion are shown below; X_c denotes the wave length corresponding to ν_2 as determined by fitting formula (263) to the observed values of C_v , and X_∞ is the wave length of the residual rays for the same substance, as found by Rubens:¹

	NaCl	KCl	CaF ₂
λ_c	50 μ	61.5 μ	34 μ
λ_∞	52	63.4	31.6

For NaCl and KCl, only one ν_1 was used in Eq. (263); for CaF₂, two were used, of which one appears to be inactive in optical phenomena and is not considered here.

That characteristic frequencies calculated from *compressibilities* or *melting points* should be of the same order of magnitude as *residual* frequencies, as we have seen them to be (Sec. 162), is understandable if the forces between the atoms of a given cell, which determine the residual frequencies, are of about the same order of magnitude as those between atoms in different cells, which also play a role in the determination of compressibilities and melting points. It is not so easy, however, to see just why all of these frequencies should agree roughly, as they actually do, with the limiting frequency ν_m of the Debye theory.

466. The Molecular Heat of Gases: Classical Theory.—The specific heat of a gas of low density presents quite a different problem from that of a solid. The molecules are relatively far apart during

¹FORSTBRING, *Zeto. f. Physih*, vol. 3, p. 9 (1920).

most of the time, so that whatever mutual potential energy they may possess, associated with the forces that they exert on one other, forms only a small part of the total energy. The energy of the gas can thus be regarded as the sum of the energies of the individual molecules.

Each molecule will have, in general, at least three types of motion—translation, rotation, and, for polyatomic gases, vibration of the atoms of the molecule with respect to each other. In addition, we have, of course, electronic vibrations, which, for the present, we shall not consider. Associated with each one of these types of motion we should expect to find a quantity of energy depending on the number of degrees of freedom of each type. The total energy W_M of a gram-molecule at temperature T should, therefore, be

$$W_M = W_t + W_r + W_v, \quad (264)$$

where W_t , W_r , and W_v stand, respectively, for the energy per gram-molecule associated with translation, rotation, and vibration.

Regarding translation, there should be 3 degrees of freedom per molecule, whether the molecule is monatomic or polyatomic. Assume $\frac{1}{2}kT$ of kinetic energy per degree of freedom, we should have

$$W_t = N_0 \times 3 \times \frac{1}{2}kT = \frac{3}{2}RT. \quad (265)$$

We should also expect to find 3 degrees of freedom of rotation, for the molecule should be capable of rotation about each of three mutually perpendicular axes. Hence, there should be rotational energy per gram-molecule of magnitude $W_r = \frac{3}{2}RT$. Thus, a gas composed of monatomic molecules should have, since obviously $W_v = 0$, total energy per gram-molecule of magnitude

$$W_M = W_t + W_r = \frac{3}{2}RT + \frac{3}{2}RT = 3RT,$$

and its specific heat should be

$$C_v = 3R = 5.96 \text{ calories per mole per degree.}$$

Now the values of C_v for the monatomic gases helium and argon at -180°C . are observed to be 3.01 cal. per mole;¹ for all of the inert gases at all temperatures, C_v is close to 3. This value is *almost exactly half the value just predicted from classical theory*. To explain this fact, since we know that the molecules of a gas have *translation*, it must be assumed that the *rotation* of such atoms, if rotation exists, does not contribute to the molecular energy. Before the advent of quantum theory, it was supposed that such atoms were spherically symmetrical, so that the forces exerted on them during collisions

¹ "International Critical Tables," vol. V, p. 80.

would act along lines through their centers and hence would never set them into rotation, or would not alter any rotatory motion they might already possess. For a gas composed of such molecules, therefore, we should have

$$W_M = W_t = \frac{3}{2}RT = 2.98 T, \\ C_v = 2.98 \text{ calories per mole per degree,}$$

in good agreement with the experimental values.

A *diatomic* molecule might then be supposed to consist of 2 atoms *bound rigidly* together, forming a "dumbbell" molecule. If the line through the atomic centers is assumed to constitute an axis of symmetry, any rotational motion about this axis that might exist would never change as a result of collisions and therefore would contribute nothing to the specific heat. As was stated in Sec. 78, however, there would be 2 degrees of freedom corresponding to rotation about each of two mutually perpendicular axes passing through the center of gravity of the molecule at right angles to its axis. These 2 degrees of freedom should contribute kinetic energy of rotation of amount $W_r = 2 \times \frac{1}{2}kT = kT$. Thus, for a gram-atom of a gas composed of rigid symmetrical dumbbell molecules, we should have

$$W_M = W_t + W_r = \frac{3}{2}RT + RT = \frac{5}{2}RT$$

and

$$C_v = \frac{5}{2}R = 4.96 \text{ calories per mole per degree.}$$

The measured molecular heats for six diatomic gases are given below:¹

Gas.	H ₂	N ₂	O ₂	CO	HCl	Cl ₂
$\frac{1}{2}p$	4.84	4.94	4.98	4.94	5.02	6.02

Five of these values of C_v agree well with the value just deduced theoretically for rigid dumbbell molecules. The experimental value for chlorine, however, is much too high. Some explanation for this discrepancy must be found.

An obvious way in which higher values of C_v might arise is by the occurrence of energy of vibration of the atoms relative to each other. As we have seen in Sees. 78 and 79, the average energy for a vibrational degree of freedom is $2Ei = 2 \times \frac{1}{2}kT = kT$. Thus, for a gram-atom of *vibrating* diatomic molecules, we have $W_v = RT$; and, for its total thermal energy and specific heat, we have

¹ C_v is calculated as $C_v = C_p/\gamma$ from data in the "International Critical Tables," *loc. cit.*

$$W_M = W_t + W_r + W_v = (\frac{3}{2} + 1 + 1)RT = \frac{5}{2}RT,$$

$$C_v = \frac{5}{2}R = 6.95 \text{ calories per mole per degree.}$$

This, however, is *too large* to agree with the experimental value for chlorine! The specific heat of chlorine thus presents an outstanding difficulty for the classical theory.

A *rigid triatomic* molecule should certainly have 3 degrees of freedom in respect of rotation, if we assume that all 3 atoms do not lie in a straight line. Hence, for a gram-atom of such molecules,

$$W_M = W_t + W_r = (\frac{3}{2} + \frac{3}{2})RT = 3RT,$$

$$C_v = 3R = 5.96 \text{ calories per mole per degree.}$$

Actually, the molecular heat of water vapor is 5.96; that for CO_2 at 20°C . is about 6.2. The specific heat of gases the molecules of which contain more than 2 atoms, however, is usually well above $3R$. The excess could easily be accounted for by supposing that in such molecules the atoms are capable of vibration. But then it is hard to understand why atomic vibration should not occur in H_2O and CO_2 as well, the energy of which is fully accounted for by translation and rotation alone.

Thus classical theory, by means of special assumptions as to molecular structure, could account for the specific heats of many gases but not for all. Another respect in which the theory failed was in regard to the *variation of specific heat with temperature*. There is no reason, according to classical theory, to expect any variation at all. Yet experiment showed that the specific heats of apparently all polyatomic gases exhibit marked variation with temperature when observations are extended over a sufficiently wide range of temperature.

167. Quantum Theory of the Specific Heat of Gases.—The key to the solution of the difficulties just discussed, as to so many others, was eventually furnished by the quantum theory. It turned out that the *translational* energy of the molecules can be treated adequately by classical methods, but this is not true of their *internal energy*, in which is included the energies of rotation and of atomic vibration.

The quantum theory of the specific heat of gases can be developed in either of two slightly different forms. In the first form, the internal energy belonging to each quantum state of the molecule is treated as a whole, and no approximations are made. In the second form, the molecular energy is divided approximately into parts corresponding to the classical subdivision, as it is in the theory of band spectra.

(a) *{Relatively} Exact Theory of Molecular Specific Heats.*—Let us assume, as in the explanation of atomic and molecular spectra, that

each molecule, in addition to its motion of translation, is capable of existing in any one of a large number of internal quantum states. Let its energy when in state number j be ϵ_j . Then, when a gram-molecule of gas is in thermal equilibrium at temperature T , the N_0 molecules composing it will be distributed among their internal quantum states according to Boltzmann's formula. According to Eq. (148a) in Sec. 101, the number of molecules in state number j will be

$$N_j = C e^{-\epsilon_j/kT},$$

C being a constant. Or, if states having the same energy are grouped together, the number in composite state number r will be, as in Eq. (148c),

$$N_r = C w_r e^{-\epsilon_r/kT}, \quad (266)$$

w_r being the statistical weight of that state and ϵ_r the energy of a molecule when in that state. In the last equation, the value of the constant C is fixed by the condition that the total number of the molecules is N_0 . Hence, $\sum_r$ denoting a sum over all quantum states,

$$\sum_r N_r = C \sum_r w_r e^{-\epsilon_r/kT} = N_0.$$

Thus, we can also write, in place of (266),

$$N_r = \frac{N_0 w_r e^{-\epsilon_r/kT}}{\sum_r w_r e^{-\epsilon_r/kT}}, \quad (267)$$

the index of summation being changed here from T to K in the sum in order to avoid confusion with r elsewhere.

An expression for the internal energy of the molecules can now be written down. The energy of N_T molecules in state number T is $N_T \epsilon_T$; hence, the internal energy W_i of the molecules in a gram-molecule of gas is

$$W_i = \sum_r N_r \epsilon_r = \frac{N_0 \sum_r \epsilon_r w_r e^{-\epsilon_r/kT}}{\sum_r w_r e^{-\epsilon_r/kT}},$$

by (267). In order to find the total energy of the gas, we have to add to W_i the classical value of the translational energy, which, as given in Eq. (265), is $W_t = \frac{3}{2}RT$. The specific heat is then

$$C_v = d(W_t + W_i)/dT$$

or

$$C_v = \frac{3}{2}R + N_0 \frac{d}{dT} \frac{\sum_{\tau} \epsilon_{\tau} w_{\tau} e^{-\epsilon_{\tau}/kT}}{\sum_{\tau} w_{\tau} e^{-\epsilon_{\tau}/kT}}. \quad (268)$$

Now the molecular energies e_{τ} or e , that occur here are the energies of the same molecular states that are involved in the production of *molecular spectra*. It should be possible, therefore, to take values of e derived from a study of band spectra and to calculate C_v from them, by means of Eq. (268). Thus *a specific heat would be calculated from spectroscopic data*. This has actually been done in a number of cases, and with complete success. Such calculations constitute a striking confirmation of the correctness of the general basis underlying the quantum theory of molecular structure.

(6) *Approximate Theory of the Molecular Heats of Gases*.—In order to arrive at a general explanation of the observed features of gaseous specific heats, it is necessary now to introduce the fact, that, as explained in Sees. 155 and 156, the internal energy of a molecule can commonly be divided approximately into separate parts, in analogy with the classical picture of molecular energy as described above. We can write, approximately,

$$\epsilon_{\tau} = \epsilon_j + \epsilon_v + \epsilon_n, \quad (269)$$

where ϵ_j denotes *energy of rotation*; ϵ_v *energy of atomic vibration*; and ϵ_n , *electronic energy* associated with the electrons in the molecule.

Of these three parts, the electronic energy remains constant as long as the molecules are in thermal equilibrium at any temperature that can be realized in the laboratory; molecular collisions due to thermal agitation are not sufficiently violent to excite in appreciable number electronic states above the one of lowest energy. Hence, the existence of ϵ_n has no effect upon the specific heat. The other two parts of the energy, summed for all molecules, constitute, respectively, rotational energy of magnitude $W_r = \mathfrak{E} \ll J$ and vibrational energy of magnitude $W_v = 2e_v$. Thus, that portion of the energy of a gram-molecule of gas which varies with temperature can be written as the sum of only three parts, as in Eq. (264) above, **or** as

$$W_M = W_t + W_r + W_v,$$

Wt being as before the kinetic energy of the translatory **motion**. It is convenient **to** divide the **molecular heat** as well into three corresponding parts:

$$C_v = C_{vt} + C_{vr} + C_{vr}; \quad C_{vt} = \frac{dW_t}{dT}, \quad C_{vr} = \frac{dW_r}{dT}, \quad C_{vv} = \frac{dW_v}{dT}.$$

These three parts of WM and of C_v may then be discussed separately.

The *translational* energy and heat are, as in classical theory,

$$W_t = \frac{3}{2}RT, \quad C_{vt} = \frac{3}{2}R. \quad (270a,b)$$

A monatomic gas has no other kind of energy; hence for it $C_v = \frac{3}{2}R$, in agreement with experiment. No gas at low density can have (or has) a smaller specific heat than this.

The *rotational* energy, also, can be handled independently. To do this we need first to study the distribution law of the molecules from the standpoint of the rotational motion considered separately.

The three parts into which the energy is divided in Eq. (209) correspond, so to speak, to three sets of "partial" molecular quantum states; there are rotational, vibrational, and electronic quantum states, and a true quantum state for the molecule is a combination of one of each of these three. We may ignore the electronic state, however, since that never changes. Let the statistical weights (or number of degenerate fundamental quantum states) for the (composite) rotational and vibrational states be W_j and w_v , respectively. Then $w_T = W_j W_v$. Thus, the number of molecules in molecular quantum state number r is, from (266) and (269),

the factor $e^{-\epsilon_v/kT}$ being included here in the constant C . The total number of molecules in rotational state number J , symbol N_j , can then be found by adding up molecules in all possible vibrational states, which gives

$$N_J = \left(C' \sum_v w_v e^{-\epsilon_v/kT} \right) w_J e^{-\epsilon_J/kT}.$$

Here the quantity in parentheses represents a new constant of proportionality. Its value can be found from the condition that

$$\sum_J N_J = N_0,$$

which gives

$$N_0 = \left(C' \sum_v w_v e^{-\epsilon_v/kT} \right) \sum_J w_J e^{-\epsilon_J/kT}.$$

Solving this last equation for the quantity in parentheses and substituting in the preceding equation, we obtain

$$N_J = \frac{w_J N_0 e^{-\epsilon_J/kT}}{\sum_J w_J e^{-\epsilon_J/kT}} \quad (271)$$

for the number of molecules in rotational state number J . This is exactly the same distribution law that would hold if the molecules differed in no other respect than in their rotational states.

The rotational energy of a gram-molecule is then $W_r = \sum_J \epsilon_J N_J$; and the rotational part of the molecular heat is

$$C_{vr} = \frac{dW_r}{dT} = N_0 \frac{d \sum_J \epsilon_J w_J e^{-\epsilon_J/kT}}{\sum_J w_J e^{-\epsilon_J/kT}}.$$

This formula for C_{vr} predicts an interesting variation of C_{vr} with temperature. At very low temperatures C_{vr} should become vanishingly small. For, if T is small enough, the Boltzmann factor $e^{-\epsilon_J/kT}$ for the state of lowest ϵ_J will be immensely larger than it is for any higher state, even for the next higher. Both sums then reduce effectively to their respective first terms, in which $J = 0$. Thus, the last fraction in (272) becomes

$$\epsilon_0 w_0 e^{-\epsilon_0/kT} / w_0 e^{-\epsilon_0/kT} = \epsilon_0$$

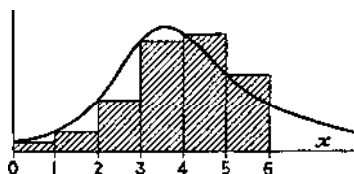


FIG. 134.

approximates to the classical value. For a molecule of the rigid diatomic or dumbbell type, the proof is so simple that we may give it here. For such a molecule, as was stated in Eq. (241) in Sec. 154,

$$\epsilon_J = J(J+1)B, \quad B = \frac{h^2}{8\pi^2 I},$$

I being the moment of inertia of the molecule about an axis perpendicular to its axis of symmetry. In this case $w_J = 2J+1$ (i.e., the number of different values of the quantum number M). If T is large enough, many terms will contribute appreciably to the sum, the variation from one term to the next being quite gradual except for the first few terms; hence, the sums can be replaced with sufficient approximation by integrals. Thus

$$\sum_J w_J e^{-\epsilon_J/kT} = \sum_{J=0}^{\infty} (2J+1) e^{-BJ(J+1)/kT} = \int_0^{\infty} (2x+1) e^{-Bx(x+1)/kT} dx$$

approximately. The equivalence of the integral to the sum is evident from Fig. 134, in which the shaded area composed of rectangles represents the sum, whereas the area under the curve represents the

integral. (The figure is merely qualitative, and is drawn for such a low temperature that the approximation would actually be poor.) In a similar way we find, integrating once by parts, that

$$\begin{aligned}\sum_j \epsilon_j w_j e^{-\epsilon_j/kT} &= B \sum_{j=0}^{\infty} J(J+1)(2J+1) e^{-BJ(J+1)/kT} = \\ B \int_0^{\infty} (x^2+x)(2x+1) e^{-B(x^2+x)/kT} dx &= -kT(x^2+x) e^{-B(x^2+x)/kT} \Big|_{x=0}^{x=\infty} \\ &+ kT \int_0^{\infty} (2x+1) e^{-B(x^2+x)/kT} dx = kT \int_0^{\infty} (2x+1) e^{-Bx(x+1)/kT} dx.\end{aligned}$$

Upon substituting these values of the sums in (272) we find that the integral cancels out and

$$C_{vr} = N_0 \frac{d}{dT}(kT) = N_0 k = R.$$

This is the classical value as found above (total $C_v = C_{vt} + C_{vr} = 5/2 R$, $C_{vt} = 3/2 R$, $C_{vr} = R$).

At intermediate temperatures C_{vr} has a value between 0 and R which can be calculated by evaluating the sums in Eq. (272).

The *vibrational* energy can be treated in a similar way, provided an expression can be found, or is assumed, for W_v . At low temperatures, all molecules will remain in their lowest vibrational states, and $C_{vV} = 0$. Since the spacing of the vibrational levels is commonly many times that of the rotational levels, C_{vV} will begin to increase from 0 only at a much higher temperature than C_{vr} . Eventually, however, C_{vV} likewise approaches the classical value, which, for a vibrating diatomic molecule, we have found to be equal to R .

Thus the quantum theory predicts the following qualitative mode of variation of C_v with temperature for a polyatomic gas of low density. At very low temperatures, $C_v = 3/2 R$ for all gases. As T increases, C_v rises (as molecular rotation develops) toward the value of C_v as deduced from classical theory with the inclusion of rotational motion. Thereafter, C_v will remain practically constant again, until molecular vibrations begin to set in, whereupon a further rise in C_v will occur. At very much higher temperatures, a third rise may be expected for any gas, even the monatomic ones, as electronic excitation begins to occur; but temperatures of this order are not at present available in the laboratory.

168. Comparison of the Theory with Observed Specific Heats.—So far as is known, the specific heats of gases vary with temperature in the manner just deduced from theory. The most extensive data exist for hydrogen¹; the observed curve for this gas is shown in Fig. 135,

¹ Cf. "International Critical Tables," vol. V, pp. 79, 84.

the absolute temperature being shown on a logarithmic scale. The shape of the curve agrees with the theoretical prediction.

From *b* to *c* (30 to 60°K.), only the translational motion of the molecules varies as the temperature rises; if there is rotational energy present, it is constant. Presumably this part of the curve continues in a horizontal direction, as indicated by *b* • .. *a*, to 0°K. In this region, hydrogen behaves like a monatomic gas.

At *c* (about 60°), a rise begins, owing to an increasing amount of rotational energy. At *d* (270°K.), the rotational energy has attained

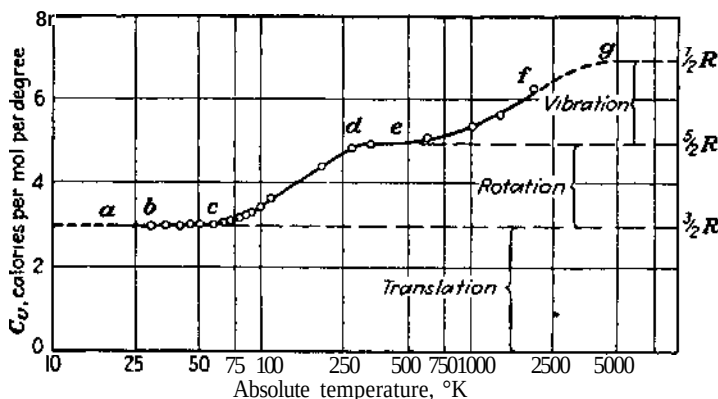


FIG. 135.—Specific heat at constant volume of ordinary hydrogen plotted against temperature on a logarithmic scale.

its full classical value equal to R ; and, from here to *e* (500°K.), C_v remains constant again, at very nearly the classical dumbbell value, $C_v = \frac{5}{2}R$. Beyond *e*, vibration begins, an appreciable fraction of the molecules beginning to exist in their next higher vibrational state.

The approximate constancy of C_v for hydrogen at "ordinary" temperatures is thus seen to be of the nature of an accident, resulting from the fact that the earth's atmosphere has a temperature of some 290°K.

The complete theory of the hydrogen molecule is, however, more complicated than as just described. There are *two forms* of the molecule; these represent alternative quantum states for the pair of protons in it, in analogy with the singlet and triplet states for 2 electrons in an atom (Sec. 138, 140). Hydrogen composed entirely of molecules of one of these two kinds is called parahydrogen, that composed of the other, orthohydrogen. Only 25 percent of ordinary hydrogen is parahydrogen; but, if ordinary hydrogen is adsorbed on charcoal cooled to 20°K. (by means of liquid hydrogen), it becomes converted in the course of a few hours into practically pure parahydrogen.

In most respects, the two forms of hydrogen possess similar physical and chemical properties, but they differ markedly in their band spectra and in the mode of variation of their specific heats with temperature. These differences arise from the fact that the rotational states of parahydrogen are restricted to even values of J , those of orthohydrogen to odd values. As a consequence, below 50°K . the molecules of orthohydrogen possess rotational energy which does not vary with temperature, J being effectively limited to the fixed value $J = 1$, whereas the molecules of parahydrogen are not rotating ($J = 0$, effectively). We have not space here, however, to pursue this interesting subject further.¹

As a concluding illustration of the application of quantum theory to the specific heats of gases, we may return to the problem of chlorine, which was left unsolved in Sec. 166. Whereas the second vibrational state ($\nu = 1$) for gases such as N_2 , H_2 lies far above the first ($\nu = 0$), for chlorine it lies only 560 cm^{-1} above the first, as may be inferred from a study of the electronic band system that extends in the absorption spectrum of chlorine from $4,800$ to $5,800\text{ \AA}$.² The difference in energy between these two states is thus

$$\epsilon_1 - \epsilon_0 = h\nu = 6.61 \times 10^{-27} \times 560 \times 3 \times 10^{10} = 1.11 \times 10^{-13} \text{ erg.}$$

Since $k = 1.38 \times 10^{-16}$, at $T = 288^\circ\text{K}$. the ratio of the Boltzmann factors for the two states is $e^{-(\epsilon_1 - \epsilon_0)/kT} = e^{-2.80} = 0.061$. Thus, 0.061 times as many molecules will be in the second vibrational state as in the first. For a rough estimate of the effect upon the specific heat, we may suppose that the fraction $e^{-(\epsilon_1 - \epsilon_0)/kT}$ of all of the molecules are in the second state. Then the excess vibrational energy due to this cause, in a gram-molecule of gas containing N_0 molecules, will be

$$N_0 (\epsilon_1 - \epsilon_0) e^{-(\epsilon_1 - \epsilon_0)/kT};$$

and the derivative of this expression with respect to T gives for the contribution of the excess energy to the molecular heat

$$N_0 \frac{(\epsilon_1 - \epsilon_0)^2}{kT^2} e^{-(\epsilon_1 - \epsilon_0)/kT} = R \left(\frac{\epsilon_1 - \epsilon_0}{kT} \right)^2 e^{-(\epsilon_1 - \epsilon_0)/kT} = 0.47 R,$$

where $R = N_0 k$. Adding the usual $\frac{5}{2}R$ for a diatomic molecule, we have then $C_v = 2.97R = 2.97 \times 1.98 = 5.88$ calories per mole per degree. This agrees fairly well with the observed value of 6.02 as given in Sec. 166. A more exact theoretical calculation, by the

¹Cf. KENNARD, E. H., "Kinetic Theory of Gases," p. 262, 1938; FARKAS, "Orthohydrogen, Parahydrogen and Heavy Hydrogen," 1935.

²KUHN, *Zeitsif. Physik*, vol. 39, p. 77 (1926).

method described in Sec. 167(a) above,¹ gives $C_v = 6.06$. Thus quantum theory succeeds where classical theory failed.

When a similar calculation is made for HC1, whose vibration-rotation band at 3.4u or 2,886 cm.¹ [Sec. 155(6)] indicates that $\epsilon_1 - \epsilon_0 = 2,886 \text{ cm.}^1 = 5.72 \times 10^{-13} \text{ erg}$, the ratio of the Boltzmann factors is found to be only $e^{-(\epsilon_1 - \epsilon_0)/kT} = 5 \times 10^{-7}$. Thus, molecular vibration can contribute nothing appreciable to the specific heat of HC1, in agreement with the observed fact that for it $C_v = \frac{5}{2}R$ very nearly. For O_2 , $\epsilon_1 - \epsilon_0 = 1,556 \text{ cm.}^1$, and vibration may contribute about 0.026R to C_v .

¹ TRAUTZ and ADER, *Zeits. f. Physik*, vol. 89, p. 15 (1934).

CHAPTER X

X-RAYS

There is probably no subject in all science which illustrates better than X-rays the importance to the entire world of research in pure science. Within 3 months after Roentgen's fortuitous discovery, X-rays were being put to practical use in a hospital in Vienna in connection with surgical operations. The use of this new aid to surgery soon spread rapidly. Since Roentgen's time, X-rays have completely revolutionized certain phases of medical practice. Had Roentgen deliberately set about to discover some means of assisting surgeons in reducing fractures, *it is almost certain that he would never have been working with the evacuated tubes, induction coils, and the like, which led to his famous discovery.*

In other fields of applied science, both biological and physical, uses have been found for X-rays, which approximate in importance their use in surgery. One may mention, for example, the study of the crystal structure of materials; "industrial diagnosis," such as the search for defects in the materials of engineering; the detection of artificial gems or of overripe fruit; the study of old paintings; the study of genetics by the biologist; the use of X-rays by the doctor in radiography and in radiotherapy; and many other uses.

But transcending these uses in applied science are the applications of X-rays made to such problems as the atomic and the molecular structure of matter and the mechanism **of** the interaction of radiation with matter. X-rays provide us with **a kind** of supermicroscope, by means of which we can "see" not **only atoms** and their arrangement in crystals but also even the interior **of the** atom itself. Roentgen's discovery must be ranked with **the most** important scientific discoveries of all time.

In this chapter, we shall give **a** brief account of the development and the present status of X-rays, with particular reference to their application to some of the fundamental problems of physics.

EARLY, MOSTLY QUALITATIVE DEVELOPMENTS IN X-RAYS (1896-1912)

169. Roentgen's Discovery.—In the autumn of 1895, **Wilhelm Konrad** Roentgen, professor of physics at **Wurzburg**, was studying that fascinating phenomenon, the discharge of electricity through

rarefied gases. A large induction coil was connected to a rather highly evacuated tube (Fig. 136), the cathode C being at one end and the anode A at the side. The tube was covered "with a somewhat closely fitting mantle of thin black cardboard."¹ With the apparatus in a completely darkened room, he made the accidental observation that "a paper screen washed with barium-platino-cyanide lights up brilliantly and fluoresces equally well whether the treated side or the other be turned toward the discharge tube."¹ The fluorescence was observable 2 meters away from the apparatus. Roentgen soon convinced himself that the agency which caused the fluorescence originated

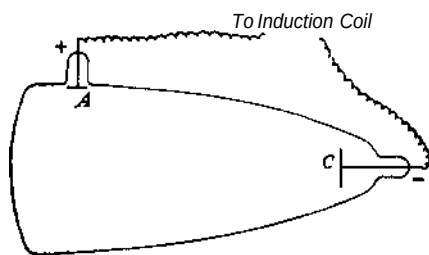


FIG. 136.—Diagram of the tube with which Roentgen discovered X-rays.

at that point in the discharge tube where the glass walls were struck by the cathode stream in the tube.

Realizing the importance of his discovery, Roentgen at once proceeded to study the properties of these new rays—the unknown nature of which he indicated by calling them "X-rays." In his first communications he recorded, among others, the following observations:

1. All substances are more or less transparent to X-rays. For example, wood 2 to 3 cm. thick is very transparent. Aluminum 15 mm. thick "weakens the effect considerably, though it does not entirely destroy the fluorescence." Lead glass is quite opaque, but other glass of the same thickness is much more transparent. "If the hand is held between the discharge tube and the screen the dark shadow of the bones is visible within the slightly dark shadow of the hand."

2. Many other substances besides barium-platino-cyanide fluoresce—calcium compounds, uranium glass, rock salt, etc.

3. Photographic plates and films "show themselves susceptible to X-rays." Hence, photography provides a valuable method of studying the effects of X-rays.

¹ Quotations from papers by Roentgen, *Electrician*, vol. 36, pp. 415, 850 (1896).

4. X-rays are neither reflected nor refracted. Hence, "X-rays cannot be concentrated by lenses." (Subsequent research, however, has demonstrated that both reflection and refraction can be, observed under special conditions).

5. Unlike cathode rays, X-rays are not deflected by a magnetic field. They travel in straight lines, as Roentgen showed by means of "pinhole" photographs.

6. X-rays discharge electrified bodies (1) whether the electrification is positive" or negative.

7. X-rays are generated when the cathode rays of the discharge tube strike any solid body. A heavier element, such as platinum, however, is much more efficient as a generator of X-rays than is a lighter element, such as aluminum.

It is a stirring tribute to Roentgen's masterly" thoroughness that most of the basic properties of X-rays were described in the paper in which the discovery was first announced.

Roentgen's discovery excited intense interest throughout the entire scientific world. His experiments were repeated and extended in very many laboratories in both America and Europe.² This early work is beautifully illustrative of the qualitative phase of development of a typical field of physics; it is with special interest, therefore, that we review in the next four sections a few of the outstanding experiments and theories of the period from 1895 to about 1912.

170. Production and Measurement of X-Rays. (a) *Tubes*.—Early tubes for the production of X-rays very soon became more or less standardised along the lines suggested by Roentgen in his second paper. Figure 137(a) shows such a tube. A residual gas pressure of the order of 10^{-3} mm. Hg provides, when voltage is applied, a few electrons and positive ions. These positive ions, bombarding the cathode C, release electrons—or "cathode rays," as they were called then—and the electrons, hurled against the anode A, give rise to X-rays. A curved cathode converges the electrons into a focal spot on A of desired shape and size. In this type of tube, known as the "gas" tube, the anode current, applied voltage, and gas pressure are more or less interdependent, and it is essential that the gas pressure be maintained at the desired value. Various ingenious devices were

¹ This property of X-rays was independently discovered by Sir J. J. Thomson. [see letter to *Electrician* (Feb. 4, 1896)]. Thomson pointed out that this phenomenon provides a method of studying X-rays much more delicate and expeditious than the photographic plate or the fluorescent screen and, further, that it yields quantitative measurements.

² The *Beiblätter zu den Ann. d. Physik* for 1896 contain 400 titles on X-rays.

introduced for accomplishing this. These gas tubes were practically the only source of X-rays until the introduction of the Coolidge tube¹ in 1913 and are still extensively used.

The Coolidge type of X-ray tube, illustrated in Figure 137(6), is evacuated to the highest attainable vacuum and incorporates as the essential part of the cathode a hot spiral filament *F* of tungsten of which the temperature, or electron emission, can be separately controlled. This distinguishing feature of the Coolidge tube, viz., control of the tube current independently of the applied voltage, has greatly facilitated certain types of X-ray research.

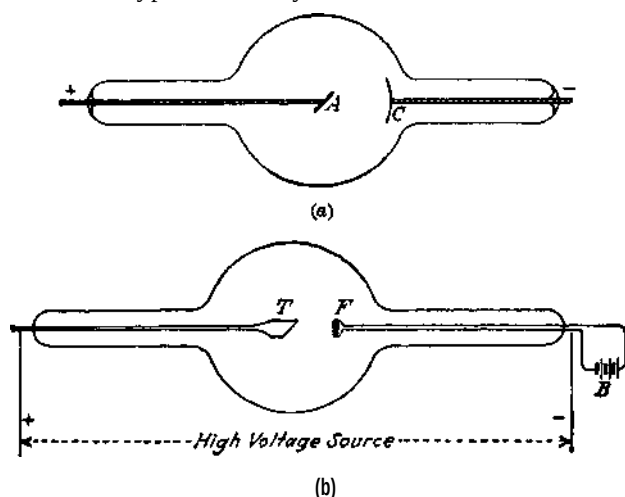


FIG. 137.—(a) An early form of X-ray tube; (b) the Coolidge X-ray tube.

(b) *Measurement of X-ray Intensity.*—Methods for detecting X-rays and for studying qualitatively their relative intensities were based upon the fluorescent and photographic effect of X-rays as reported by Roentgen. The photographic method was especially convenient in medical radiographic work. The application of intensifying screens to increase the effective speed of photographic emulsions was an improvement made independently by numerous investigators.² But for quantitative measurements of intensity the photographic method is rather complicated, and it soon gave way to the ionization method, which is still used today.

The discharging effect of X-rays upon charged bodies, noted above, was soon traced to ionization of the air molecules. Villari³ showed that the discharging action in a gas at a given pressure depends on the

¹ COOLIDGE, *Phys. Rev.*, vol. 2, p. 409 (1913).

² *Comptes Rendus*, vol. 122, pp. 312, 702, 720 (1896); *Electrician*, vol. 36, p. 702 (1896).

³ VILLARI, *Comptes Rendus*, vol. 123, pp. 418, 446 (1896).

nature of the gas. The following were found increasingly active in the order given: H, CO, air, CO₂, ether vapor, CS₂. Benoist and Hurmuzescu¹ showed that, for a given gas, the discharging action increases rapidly with density.

At first, the rate of discharge of a charged electroscope was used in measuring the intensity of an X-ray beam, the motion of the leaves being observed under a low-power microscope. Later, an auxiliary device known as an "ionization chamber" was introduced. This is shown schematically in Fig. 138. *C* is a metal tube several centimeters in diameter, from about 20 to 100 cm. long, and closed at both ends except for an opening or "window" *W*, over which may be placed a thin sheet of cellophane or aluminum for admitting the

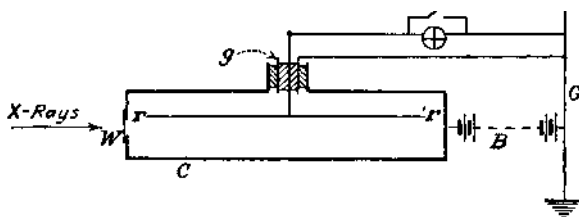


FIG. 138- An ionization chamber used for measuring electrically the intensity of beam of X-rays.

X-rays. A rod *rr* suitably supported by *good* insulating material, such as amber or quartz, is connected to an electrometer. An electric field is maintained between the rod *rr* and the cylinder *C* by a battery *B*, of about 100 volts, one end of which is connected to the ground wire (7). An earthed guard ring *g* prevents leakage from the cylinder to the rod *rr*. The cylinder may be filled with a heavy gas to make the arrangement more sensitive; argon or methyl bromide is often used. When X-rays enter the window *W*, the gas within the cylinder is made conducting, and, on account of the electric field between the cylinder and the rod, the latter acquires a charge at a rate which can be measured by the electrometer. This rate is a measure of the intensity of the X-ray beam.

Nowadays the electrometer is often replaced by a vacuum-tube amplifier and a galvanometer. For very low X-ray intensities, a Geiger counter with appropriate vacuum-tube circuit [Sec. 198(6)] may supplant the ionization chamber.

171. Classical Pulse Theory of X-rays.—We have seen that X-rays are produced whenever electrons are suddenly brought to rest by colliding with a solid obstacle. This fact early suggested a simple theory of the mode of their production. While being brought to rest, the electrons must experience, for a brief interval, a very large

¹ *Comptes Rendus*, vol. 122, p. 926 (1896).

negative acceleration. According to classical electromagnetic theory, such an accelerated body must radiate energy. The sudden stopping of each electron as it collides with the target, therefore, must result in the emission of an electromagnetic disturbance or pulse. On this theory, X-rays consist of a very rapid succession of such pulses, coming at random intervals. It was shown by Stoney that, if such a stream of pulses is analyzed into wave trains, the components of shorter wave length are the more intense the greater the velocity of the electrons which are brought to rest by the target. Stoney showed also that matter should, in general, be more transparent to the shorter waves than to the longer waves. The hard, or penetrating, X-rays should, therefore, be produced by high voltages applied to the X-ray tube. Qualitatively, this picture of the mechanism of the production of X-rays seemed to be in agreement with the experimental facts.

(a) *The Production of X-rays.*—According to Eq. (31) in Sec. 35, based on classical electromagnetic theory, if an electron moving at a velocity v_0 small relative to the velocity of light, is brought to rest with a uniform acceleration a , the total energy radiated is

$$W = \frac{2}{3} \frac{e^2 v_0^2 |a|}{c^3},$$

e being the electronic charge. The time taken in stopping is $t = v_0/|a|$; during this time the electron travels a distance $s = |a|t^2/2 = v_0^2/(2|a|)$. In terms of s , we can also write

$$W = \frac{e^2 v_0^4}{3c^3 s}.$$

This formula predicts that the energy radiated will increase very rapidly with the velocity of the electron and will be inversely proportional to the distance within which the electron is stopped. Qualitatively, these same features might be expected to hold for stoppage of an electron by an atom, although in that case the acceleration would not, of course, be uniform. Experiment shows that the intensity of the radiation from a given X-ray target increases roughly as v_0^4 , i.e., as the square of the energy of the cathode ray, which is at least not grossly inconsistent with our theoretical expectation. Furthermore, it is reasonable to assume that an electron moving with a given velocity will be stopped more suddenly when colliding with a heavy atom than with a lighter one. A target made of a heavy metal like platinum should, therefore, give out more X-ray energy, other things being equal, than one made of a lighter metal such as aluminum. This is in agreement with Roentgen's observations.

Qualitatively, therefore, the classical theory of X-ray production seemed to be well confirmed. Quantitatively, however, this theory

was destined to meet with serious difficulties. It still retains an interest, although we now know that it has reference only to one part of the X-ray spectrum (known as the "continuous" spectrum) and, for accuracy, must be replaced even in dealing with that part by a more abstract theory, wave-mechanical in nature.

The remark may be added here that other charged particles than electrons were later found to be effective for the production of X-rays, such as α -rays,¹ protons, and deuterons.²

(6) *Refraction and Diffraction of X-rays.*—Tests of the classical pulse theory were by no means confined to predictions concerning the production of X-rays. Convincing demonstrations of such classical phenomena as reflection, refraction, and diffraction were eagerly sought. Roentgen's original experiments to this end had been negative.

Gouy,³ using a narrow line source of X-rays, obtained a shadow of a fine platinum wire on a photographic plate. Then he placed a prism in the lower half of the beam in such a way that if the rays were *refracted* by the prism the shadow of the wire would be "broken." The shadow was sharp and continuous, showing no measurable refraction. Gouy estimated that the index of refraction could not be greater than 1.000005. It was pointed out by Maltezos⁴ that the Helmholtz dispersion formula

$$n^2 = 1 + C\lambda^2 + \dots,$$

where n is the refractive index and C is a constant, predicts that $n \rightarrow 1$ as $\lambda \rightarrow 0$. Thus the absence of measurable refraction of X-rays would be accounted for if they had *extremely short wave lengths*.

Gouy⁵ also searched for evidences of diffraction, with negative results. A little later, however, Haga and Wind,⁶ using wedge-shaped slits only a few thousandths of a millimeter wide, observed evidences of a slight widening of the image on the photographic plate, from which they deduced that the wave length of the rays must be of the order of 10^{-8} cm. Walter and Pohl,⁷ in a similar diffraction experiment, confirmed this order of magnitude of the wave length.

172. The Scattering of X-rays.—In an attempt to discover the reflection of X-rays, Imbert and Bertin-Sans⁸ arranged apparatus

¹ CHADWICK, *Phil. Mag.*, vol. 25, p. 193 (1913).

² CORK, *Phys. Rev.*, vol. 59, p. 957 (1941).

³ GOUY, *Comptes Rendu**, vol. 122, p. 1197 (1896).

⁴ *Comptes Rendus*, vol. 122, pp. 1115, 1474, 1533 (1896).

⁵ *Comptes Rendus*, vol. 123, p. 43 (1896).

⁶ *Ann. d. Physik*, vol. 68, p. 884 (1899).

⁷ *Ann. d. Physik*, vol. 29, p. 331 (1909). See also SOMMERFELD, *Ann. d. Physik*, vol. 38, p. 473 (1912).

⁸ *Comptes Rendus*, vol. 122, p. 524 (1896).

as shown diagrammatically in Fig. 139. Between the source S of the rays and the photographic plate P was placed a thick copper screen A . A plane mirror M was so placed that a beam of rays, if reflected, would pass to the photographic plate P , on which would appear an image or shadow of an obstacle B . Such a shadow was obtained *irrespective of the angular position* of the mirror M . Indeed, a plate of paraffin was just as effective as the mirror. From these facts Imbert and Bertin-Sans concluded that, instead of being reflected, the rays were diffused or scattered by M , somewhat as light is scattered by fog particles.

This phenomenon of the scattering of X-rays has played a very important part in the theories of modern physics and has been the object of many researches both theoretical and experimental.

(a) *Thomson's Theory*.—The first theory of X-ray scattering, an application of the classical pulse theory described in the last section, was proposed by Sir J. J. Thomson.¹ Let a beam of X-radiation be incident upon an electron of charge $-e$ and of mass m . The radiation, being supposed electromagnetic, contains an electric vector at right angles to the direction of propagation. As the wave passes over the electron, the latter will experience a force $F = -Ee$ and will, therefore, have an acceleration in the direction opposite to F of magnitude

$$a = \frac{eE}{m}. \quad (274)$$

Consequently, the electron will radiate energy at the instantaneous rate given by Eq. (28) in Sec. 35, or

$$\Omega = \frac{2e^2a^2}{3c^3} = \frac{2e^4E^2}{3m^2c^3} \text{ ergs sec.}^{-1}$$

if e and E are in electrostatic units.

This energy is abstracted from the primary beam and is reradiated or *scattered* as a *secondary* beam. Its magnitude is most usefully expressed in terms of the intensity I' of the primary beam. This intensity, defined as the amount of energy crossing unit area per

¹ THOMSON, J. J., "Conduction of Electricity through Gases," 2d ed., p. 321.

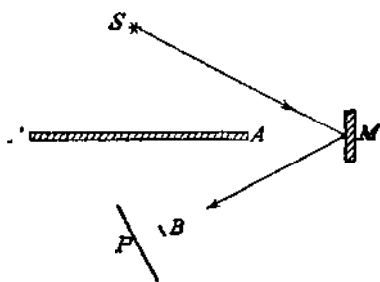


FIG. 139.—The arrangement of Imbert and Bertin-Sans showing pseudo-reflection (scattering) of X-rays from the mirror M .

second, is

$$I = \frac{cE^2}{4\pi} \quad (275)$$

[cf. Eq. (12) in Sec. 32]. Hence,

$$\Omega = \frac{8\pi}{3} \frac{c^4}{m^2 c^4} I \quad (270)$$

The ratio Ω/I , which we shall denote by σ , is called the *classical scattering coefficient for a free electron*. Of the incident radiation that falls on unit area of a surface drawn perpendicular to the beam, a fraction σ is scattered. In other words, the electron scatters as much radiation as falls on an area equal to a_e . For this reason, a_e is also called the *classical cross section for scattering by a free electron*. For its value, inserting $e = 4.80 \times 10^{-10}$, $m = 0.911 \times 10^{-27}$ gram, $c = 3 \times 10^{10}$ we find

$$\sigma = \frac{\Omega}{I} = \frac{8\pi}{3} \frac{e^4}{m^2 c^4} = 6.66 \times 10^{-25} \text{ cm.}^2 \quad (277)$$

(b) *The Number of Electrons in a Carbon Atom.*—The result just obtained provides a basis for an estimate of the number of electrons in an atom. The energy radiated by an electron must represent energy abstracted from the incident beam. If there are n electrons per unit volume in the scattering material, and if we assume that all electrons scatter independently (i.e., that no interference takes place between wavelets scattered from neighboring electrons), we may write for the fractional diminution $\Delta I/I$ of the incident beam in going a distance Δx

$$-\frac{\Delta I}{I} = n\sigma \Delta x.$$

Since a_e is independent of x , we may integrate both members of this equation, obtaining for the intensity I of the beam after traversing a thickness x of the material,

$$I = I_0 e^{-n\sigma x},$$

I_0 being the incident intensity. It is assumed here that the only loss of intensity is that due to scattering.

Actually, processes other than scattering are present and contribute to the diminution of the incident beam. Barkla and his collaborators,¹ however, were able to correct for the excess absorption, for absorbers of low atomic weight, and so were able to deduce, from measurements of J , I_0 , and x , a value of the product na_e . Using carbon of density ρ as the absorber, Barkla found

¹ BARKLA and SADLER, *Phil. Mag.*, vol. 17, p. 739 (1909); BARKLA, *Phil. Mag.*, vol. 21, p. 648 (1911).

$$n\sigma_s/\rho = 0.2,$$

from which and Eq. (277)

$$n = 3.00 \times 10^{23} \rho.$$

The number of atoms in a gram of carbon is No/A where No is Avogadro's number or 6.02×10^{23} (Sec. 79), and A is the atomic weight (12); hence the number of atoms in 1 cm.^3 is $No\rho/A$. Dividing this number into the value just found for n , we find for the number of electrons per atom of carbon

$$\frac{nA}{No\rho} = \frac{3.00 \times 10^{23} \times 12}{6.02 \times 10^{23}} = 6$$

(approximately). This was the first reasonably accurate determination of the number of electrons in an atom and was of great historical importance. However, it is only fair to say that, had Barkla performed his experiment under considerably different conditions of wave length and atomic number, the result of the computation would not have been so satisfactory.

(c) *Angular Distribution of Scattered X-rays.*—The distribution in direction of the radiation scattered by an electron is easily obtained from a formula deduced in Sec. 35. Consider the radiation that is scattered in a direction making an angle θ with the direction of the electric vector E in the primary beam. Let I_θ denote the intensity of the scattered radiation at a point P distant r cm. from the scattering electron in this direction. Then I_θ is the same as dW/dt as given by Eq. (20) in Sec. 35, except that here $q = -e$; hence

$$I_\theta = \frac{1}{4\pi} \frac{e^2 a^2}{c^3 r^2} \sin^2 \theta,$$

or, after substituting again for the acceleration a in terms of the intensity I of the primary pulse by means of (274) and (275),

$$I_\theta = \frac{I}{r^2} \frac{e^4}{m^2 c^4} \sin^2 \theta \text{ ergs cm.}^{-2} \text{ sec.}^{-1} \quad (278)$$

From this expression we see that if the primary X-ray beam is plane-polarized, with the electric vector in a fixed direction, then the radiation scattered by an electron in a direction making an angle θ with the direction of the primary electric vector E is proportional to $\sin^2 \theta$.

Let us consider now the more general case of a primary beam in which the direction of the electric vector varies at random. Such a

beam is unpolarized. Let us draw axes so that the xy plane contains the direction OP in which the scattered radiation is observed (Fig. 140), and let OP , which is perpendicular to the z -axis, make an angle ϕ with the z -axis or direction of propagation of the primary beam. Resolve the electric vector E in the primary beam into its y - and z -components, E_y and E_z . Then these components, acting upon the electron, will produce accelerations of magnitude $a_y = eE_y/m$ and $a_z = eE_z/m$, respectively. Each component of the acceleration, in turn, will produce a corresponding scattered electric vector at P . For these two scattered vectors we find, from Eq. (22) in Sec. 34, in which we insert $q = -e$ and, first $a = a_y$, $\theta = \pi/2 - \phi$, then $a = a_z$, $\theta = \pi/2$:

$$E'_1 = \frac{e^2 E_y}{c^2 m r} \cos \phi, \quad E'_2 = \frac{e^2 E_z}{c^2 m r}.$$

Since E'_1 and E'_2 are clearly perpendicular to each other at P , we have $E'^2 = E'^2_1 + E'^2_2$, where E' is the resultant electric vector in the scattered beam. If we insert the value of E'^2 thus found for E^2 in Eq. (12) in Sec. 32, we obtain for the intensity of the scattered beam at P :

$$I'_\phi = \frac{c}{4\pi} E'^2 = \frac{1}{4\pi} \frac{e^4}{c^3 m^2 r^2} (E_y^2 \cos^2 \phi + E_z^2).$$

The average value of I' can then be found by replacing E^2 and E in this expression by their average values $\overline{E^2}$ and $\overline{E}$, the bar over a symbol denoting its time average.

The average intensity of the primary beam is, similarly,

$$I = \frac{c}{4\pi} \overline{E^2} = \frac{c}{4\pi} (\overline{E_y^2} + \overline{E_z^2}).$$

Because of the random variation of the direction of E , however, $\overline{E_y^2} = \overline{E_z^2}$. Hence,

$$\overline{E_y^2} = \overline{E_z^2} = \frac{2\pi I}{c}.$$

Thus, we can write for the average intensity of the scattered beam,

$$\overline{I}'_\phi = \frac{I}{r^2} \frac{e^4}{m^2 c^4} \frac{1 + \cos^2 \phi}{2}. \quad (279)$$

Assuming independent scattering by the individual electrons, we may multiply Eq. (279) by the number of electrons per unit volume, and thus we obtain an expression suitable for experimental test.

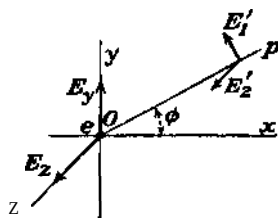


FIG. 140.

According to this result, the intensity of the scattered X-ray beam should be (1) symmetrical about the direction of propagation of the primary beam and (2) symmetrical forward and backward, relative to a plane (the yz plane in Fig. 140) passing through the scatterer and perpendicular to the direction of propagation. Early experiments seemed to confirm these predictions. More exact measurements, however, as will be pointed out in a later section [185(6)], reveal radical departures from the second condition of symmetry.

(d) *Polarization of X-rays.*—Among the early significant experiments on scattering is Barkla's classic work¹ on polarization. These

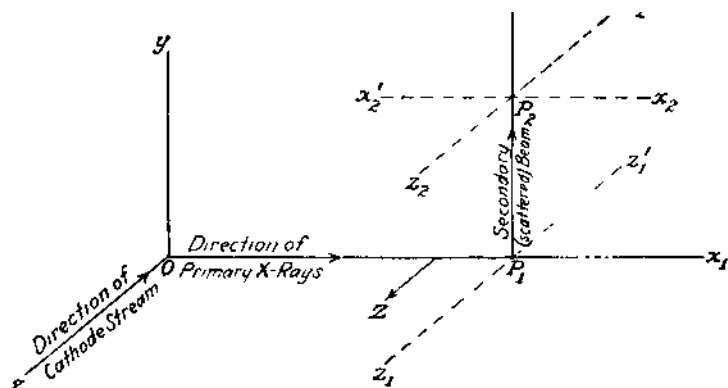


FIG. 141.

experiments test one of the conclusions drawn from the classical theory of the production of X-rays, and also the factor $(1 + \cos^2\theta)$ in Thomson's scattering equation, (279) above.

Let a stream of cathode rays, proceeding in the direction zO (Fig. 141) impinge on a target at O . Consider the beam of X-rays proceeding from this target in the direction Ox_1 , perpendicular to Oz . We shall call this the "primary beam." If we assume that the electrons of the cathode stream are all brought to rest by accelerations in the direction Oz , this primary beam should be plane-polarized with the electric vector in the z direction, as indicated by the vector Z . Let a scattering material—a small piece of carbon, say—be located at P_1 . As explained in the preceding section, according to classical theory the electrons of this scatterer should experience accelerations in the direction $P_1z'_1$ as a result of the passage of the primary beam, and the scatterer should emit a secondary or scattered beam. The intensity of this scattered

¹ BARKLA, *Roy. Soc. Proc.*, vol. 77, p. 247 (1906).

beam should be a maximum in the plane yOx_1 and should be zero in the direction $Z_1z'_1$. If this analysis is correct, the intensity of the X-rays scattered by a substance at P_1 should vary from zero in a direction P_1Z_1 to a maximum in the direction P_1P_2 .

In his experiments Barkla found that the intensity of the scattered radiation in the direction of P_1Z_1 , although not zero, was considerably less than the intensity in the direction P_1P_2 . This indicates that the primary beam is at least *partially* polarized, though not completely so. A little further consideration indicates that this is just what we should expect. We assumed, at the beginning of the discussion, that the electrons of the cathode stream are brought to rest by accelerations in the direction Oz . If we attempt to picture the sequence of events by means of which a swiftly moving electron is brought to rest by collision with the atoms of the target at O , we should conclude that rarely will an electron be stopped by a single "head on" collision with an atom. In the general case, it will pursue a zigzag course and will collide with many atoms before being brought to rest. Although the preponderance of accelerations may be in the general direction Oz , accelerations in quite different directions are to be expected. In the primary beam, therefore, although the z -components of the electric vector should predominate, y -components are also to be expected. This means that the intensity of the X-rays scattered from P_1 in the direction of P_1P_2 should merely be greater than in the direction P_1Z_1 —which is what Barkla observed.

With the *secondary* beam of X-rays proceeding in the direction P_1P_2 , however, the situation is different. This secondary beam is produced by the acceleration of electrons at P_1 due to the passage of the primary beam. On account of the transverse nature of electromagnetic radiation, this primary beam, regardless of its state of polarization, can accelerate electrons at P_1 only in directions lying in the plane $P_2P_1Z_1$, *i.e.*, in directions at right angles to P_1x_1 . Consequently, the electric vector of the *secondary* beam proceeding in the direction P_1P_2 must lie *entirely* in the plane $P_2P_1Z_1$; this secondary beam must be *completely* plane-polarized. If, then, this secondary beam be allowed to pass over a second scatterer placed at P_2 , the intensity of the *tertiary* radiation sent out from P_2 should vary from zero in the direction P_2Z_2 to a maximum in the direction P_2x_2 .

Barkla's experiments testing this conclusion indicated that the secondary rays were 70 percent polarized, instead of the predicted 100 percent. A similar experiment was performed much later by Compton and Hagenow¹ in which a more intense primary beam was

¹ COMPTON and HAGENOW, *J.O.S.A., Rev. Sci. Instruments*, vol. 8, p. 487 (1924).

employed, thereby allowing (1) better collimation, i.e., greater definition of the scattering angle, and (2) the use of smaller scatterers so as to decrease multiple scattering, i.e., successive scatterings at angles other than 90° . It was found that within the limit of error of the measurements the intensity of the tertiary radiation scattered in the direction P_2Z_2 is zero. We may conclude, therefore, that, as classical theory predicts, the secondary beam itself is *completely polarized*.

The results of these early experiments on the scattering of X-rays confirmed without exception the classical theory in its simple form as described above. But we must bear in mind, that, as viewed today, this theory of scattering is imperfect, principally for the following reasons:

1. We have dealt here with free electrons only, whereas most of the electrons involved in scattering are bound in atoms.

2. We have neglected possible interference effects between wavelets scattered from different electrons.

3. Another type of scattering process, the Compton scattering, also occurs. In a later section we shall return to the subject of scattering.

173. Absorption and Fluorescence.—Although the classical electromagnetic theory was remarkably successful in explaining most of the early observed phenomena of X-rays, there gradually accumulated a mass of data which this theory could not explain. One such group of data developed on the subject of scattering, already mentioned (Compton scattering), one in absorption (photoelectric absorption), and a third in regard to spectral distribution of energy of X-rays. The first work on the latter two groups of phenomena will be discussed in this last section dealing with early developments in X-rays.

(a) *Absorption*.—Roentgen's initial observation that the relative opacity of materials to X-rays depends not alone upon the density was early confirmed. Many investigators¹ independently concluded in 1896 that opacity of a material varied in a regular manner with its atomic weight. Not for some 15 years did the atomic number enter the picture.

In 1897, Buguet² made the significant observation that the opacity of a given material to X-rays varied with the thickness of the material previously traversed by the beam. This effect is explained by assuming that his X-ray beams were composed of several component beams

¹ See, for example, *Comptes Rendus*, vol. 122, pp. 146, 723 (1896).

² *Comptes Rendus*, vol. 125, p. 398 (1897).

having different absorbabilities. The general terms *quality* and *hardness* refer to the relative absorbability of a beam, a hard beam having low absorbability. Any beam whose quality is altered by transmission through an absorber is called a *heterogeneous* beam. In 1909, Barkla and Sadler,¹ studying these absorption questions, developed the first method of producing a *homogeneous* beam. This method is discussed in part (b) of this section.

The general formula for the absorption of radiation was found to apply to X-rays. That is, if an X-ray beam of intensity i is incident on a thin slab dx of absorbing material (Fig. 142), the change in intensity di , actually a diminution, is given by the relation

$$\frac{di}{i} = -\mu dx,$$

where μ is called the *linear coefficient of absorption*. For a homogeneous beam of X-rays, i.e., a beam for which μ is independent of x , the last equation can at once be integrated; we can write for the transmitted intensity I at any distance x

$$I = I_0 e^{-\mu x}, \quad (280)$$

I_0 being the intensity at $x = 0$. By measuring I , I_0 , and x , one may compute μ . In a later section, we shall discuss what is known concerning the dependence of μ upon atomic number and upon the quality of the rays.

In deriving Eq. (280), we are not concerned with what ultimately becomes of the radiation which is removed from the incident beam. From our discussion of scattered X-rays, however, we know that scattering constitutes one process by which an incident beam is weakened. Usually, however, the amount of radiant energy that is scattered from an X-ray beam is much smaller than the total amount of energy that is removed from the beam, as represented by the coefficient μ . This fact was not easy to explain on the classical theory.

(6) *Characteristic Secondary X-rays*.—A new type of phenomenon was revealed in 1896 by an important observation made by Winkelmänn and Straubel.² A beam of X-rays was passed through a photographic plate P (Fig. 143), the emulsion being on the rear side. Behind a part of the plate was placed a piece of fluor spar F . On developing the plate, it was found that the film was much denser in

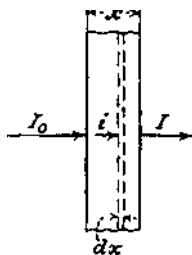


FIG. 142.

¹*Phil Mag.*, vol. 17, p. 739 (1909).

²*Jenaisch. Zeits. f. Naturwua.*, vol. 30 (1896).

the neighborhood of F , as if F had reflected the rays. An observation of this kind had been made by Roentgen. But Winkelmann and Straubel showed that the phenomenon was not one of true reflection, for they repeated the experiment with a thin sheet of paper $A A$ between F and P and found that the intensifying action of F was almost but not entirely destroyed, although the paper was very transparent to the *incident* beam. From this, they concluded that the *quality* of the rays had been altered by the spar in such a way as to make the beam of rays returned by F more absorbable in paper than was the original beam. In other words, the primary rays, incident on the spar, had been transformed into *characteristic* "spar" rays.

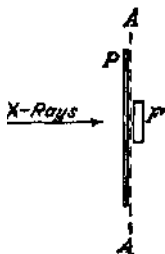


FIG. 143

Barkla and his collaborators made a thorough and systematic study of this phenomenon.¹ Let a primary beam of hard X-rays from a target T (Fig. 144), after passing through holes in lead screens $S_1 S_1$, fall upon the secondary emitter E . Let the secondary beam, taken off at right angles to the primary beam, after passing screens $S_2 S_2$, enter the ionization chamber C , by means of which the intensity of the secondary beam can be measured either with or without slabs of absorbing material placed at A or at B .

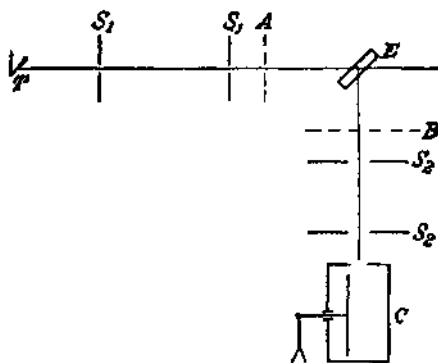


FIG. 144 —Schematic arrangement for studying secondary radiation.

When the secondary emitter E is of some light material, such as carbon, an aluminum absorbing screen placed at B absorbs nearly the same fraction of the secondary beam as it does of the primary beam when placed at A . This shows that the quality of the secondary beam as measured by its absorption in aluminum is nearly the same

¹ Cf. *Phil. Mag.*, vol. 16, p. 550; vol. 22, p. 396; vol. 23, p. 987; *Nature*, vol. 80, p. 37; *Cambridge Phil. Soc., Proc.*, vol. 15, p. 257 (1908-1912).

as that of the primary beam. We may conclude that the primary beam has merely been *scattered* by the secondary emitter.

If, however, a heavier material, such as silver, is substituted for the carbon at *E*, the absorption coefficient of the secondary beam in aluminum is found to be *greater* than that of the primary beam—which indicates that the *quality* of the secondary beam is no longer the same as that of the primary. Barkla found that the quality or hardness of the secondary beam, as measured by its coefficient of absorption in aluminum, is *characteristic of the material* used as secondary emitter. Almost without exception, *the hardness of the secondary beam*, as measured by its absorbability in aluminum, *increases with increasing atomic weight of the secondary emitter*.

When absorption coefficients in different absorbing materials were compared, a striking feature appeared. If the absorption coefficients in, say, iron, for a series of secondary radiations obtained from a series of emitters of different atomic weights, are plotted against the absorption coefficients for the same radiations in aluminum, a curve is obtained of the type shown in Fig. 145.

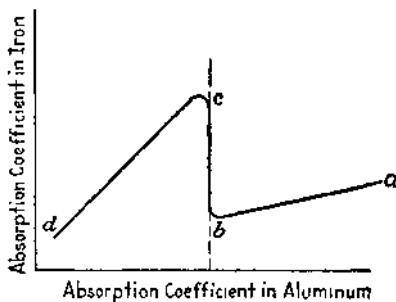


FIG. 145.

Point *a* corresponds to a soft radiation easily absorbed in aluminum. As the penetrating power of the radiation is increased, the coefficients of absorption in both aluminum and iron decrease along the line *ab*. At *b*, however, the absorption coefficient of the radiation in the iron absorber suddenly increases to point *c*. This increase occurs for radiation of which the absorbability in aluminum is about *the same* as that for the radiation from an iron secondary emitter. Thereafter, for more penetrating radiations, the absorption in iron decreases again toward point *d*.

Similar results were obtained with other absorbing materials in place of iron. The position of the discontinuity was found to depend on the atomic weight of the absorber: the higher its atomic weight, the greater the penetration in aluminum of the radiation for which the discontinuity occurred.

These facts point unambiguously to the conclusion that a secondary emitter besides scattering the primary beam also emits a fluorescent radiation characteristic of the emitter, whenever the primary beam is of greater hardness than this fluorescent radiation. Barkla called the latter the "*K fluorescent radiation*" of the pmittpr. For the

of Fig. 145. This indicated that along with the *K* fluorescent radiation another and *softer* fluorescent radiation was emitted, which Barkla designated by *L*. Barkla recognized that these *K* and *L* radiations constitute *lines* (or groups of lines) in the fluorescent X-ray spectra of the several secondary emitters.

A further conclusion from these observations is that there exists, for each of many absorbers, a *K* absorption discontinuity which is not like the resonance absorption line to be expected from the classical electromagnetic theory. In some cases, a similar *L* discontinuity was found.

The graph of Fig. 145, with the abscissas corresponding, as we shall see, to an X-ray wave-length scale, foreshadows a new era in X-ray developments, an era distinguished by studies of spectral properties of X-rays and by experimentation with homogeneous beams.

X-RAY SPECTRA

174. The Crystal Diffraction Grating.—The researches of Barkla in isolating the *K* and *L* fluorescent radiations pointed unmistakably toward the existence of *definite, discrete* wave lengths in X-rays. An experiment was greatly needed by which a heterogeneous beam could actually be separated into its spectral components. The only method then available for such spectral analysis involved diffraction of X-rays from a slit, and the dispersion obtainable by this method was insufficient for the purpose at hand.

(a) *A Crystal as a Diffraction Grating.*—A new and practical method for resolving X-ray beams developed out of a brilliant suggestion by Laue. The order of magnitude of X-ray wave lengths, as revealed by the diffraction experiments described above, is the same as the order of magnitude of the spacing of the atoms in crystals. Laue suggested, therefore, that a crystal, with its regular three-dimensional array of atoms, might behave toward a beam of X-rays in somewhat the same way as does a ruled diffraction grating toward a beam of ordinary light.

Let it be assumed that plane electromagnetic waves traveling in a given direction fall upon a crystal. Then each atom will scatter some of the incident radiation. If the crystal is perfectly regular, the wavelets scattered by different atoms will combine, in general, in all sorts of phases and so will destroy each other by interference. Laue argued, however, that for certain wave lengths and in certain directions the wavelets should combine in phase and so produce a

strong diffracted beam.¹ It would be expected, therefore, that such diffracted beams might be observed upon passing a heterogeneous X-ray beam through a crystal. Such an experiment was performed by Friedrich and Knipping in 1913.

(6) *The Experiment of Friedrich, Knipping, and Laue.*²—By means of suitable screens, S_1S_2 (Fig. 146), a narrow pencil of X-rays from the target T was allowed to pass through a crystal C beyond which was a photographic plate PP . After an exposure of many hours, it was found on developing the plate that, in addition to the interior central image at O , where the direct beam struck the plate, there were present on the plate many fainter but regularly arranged spots, indicating that the incident X-ray beam had been diffracted by the crystal in certain special directions, just as Laue had predicted. Figure 147 shows such a photograph, taken by Dr. George L. Clark,³ of an iron crystal. In their original paper, Friedrich, Knipping, and Laue, from an

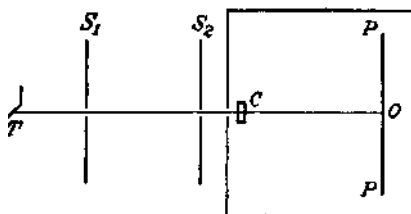


FIG. 146.—The arrangement by which Friedrich and Knipping discovered the action of a crystal on a beam of X-rays.

analysis of a series of photographs of a crystal of zinc blende oriented at various angles with respect to the incident pencil, concluded that there were present in the X-ray beam wave lengths varying between 1.27×10^{-9} cm. and 4.83×10^{-9} cm. This positive result seemed to prove the correctness of the two postulates underlying the experiment: (1) that X-rays are electromagnetic waves of definite wave lengths and (2) that the atoms of a crystal are arranged in regular three-dimensional order, as suggested by the external symmetry of crystals.

This experiment marked the beginning of a new era in the technique of X-ray measurement and in X-ray theory. Two new and very important fields of investigation were at once opened up: (1) in X-rays, the study of spectra and the use of homogeneous beams in experiments on scattering, absorption, etc.; (2) the study of the arrangements of atoms or molecules in crystals. In the following sections, we shall confine our discussion to some of the more important aspects of the former field. For presentations of the allied field of

¹See, for example, COMPTON and ALLISON, "X-rays in Theory and Experiment," pp. 331-340, 1935.

²FRIEDRICH, KNIPPING, and LAUE, *Bayer. Akad. d. Wiss.*, 1912; *Le Radium*, vol. 10, p. 47 (1913).

³This photograph is used by permission of Dr. Clark.

molecular and crystal structure, the student is referred to standard treatises on the subject.¹

(c) *Bragg's Analysis of the Three-Dimensional Grating.*—A very simple and convenient way of looking at the process of diffraction by a crystal grating was proposed by Bragg.² He pointed out that through any crystal a set of equidistant parallel planes can be drawn which, among them, pass through *all* the atoms (or similar groups of

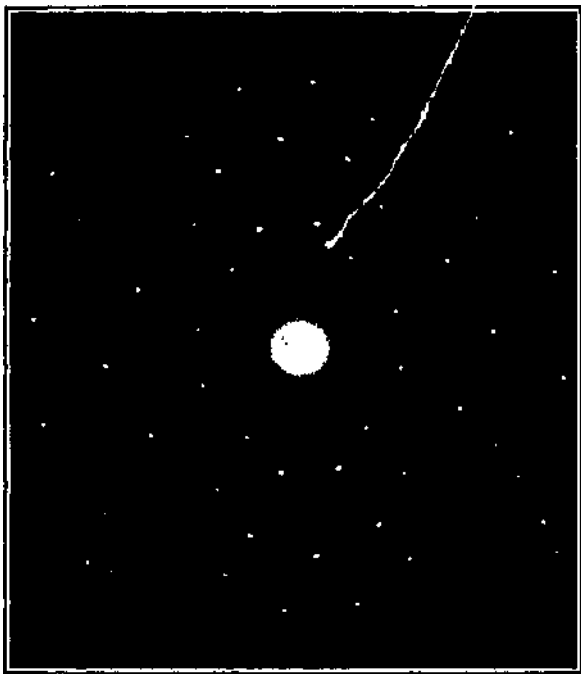


FIG. 147.—Laue photograph of an iron crystal. (Photograph by Dr. George L. Clark.)

atoms) which compose the crystal. Indeed, a great many such families of planes may be drawn, the planes of each family being separated from each other by a characteristic distance. Such planes are called *Bragg planes*, and their separations, *Bragg spacings*. Traces of five families of Bragg planes are shown in Fig. 148.

If plane monochromatic waves fall upon the atoms in a Bragg plane, a wavelet of scattered radiation spreads out from each atom in all directions. There is just one direction in which, irrespective of the atomic distribution in the plane, the scattered wavelets will meet

¹ *E.g.*, BRAGG, W. H., and W. L., BRAGG, "The Crystalline State," vol. I, 1933; CLARK, G. L., "Applied X-Rays," 3d ed., 1940.

² BRAGG, W. L., *Cambridge Phil. Soc, Proc*, vol. 17, p. 43 (1912).

in the same phase and will constructively interfere with each other, viz., the direction of *specular reflection* from the plane. This follows from the ordinary Huygens construction as used for the reflection from a mirror. The beam scattered in this direction may be thought of as reflected from the Bragg plane. But, now, we note that each Bragg plane is one of many regularly spaced parallel planes. The beams reflected from these various parallel planes will combine, in general,

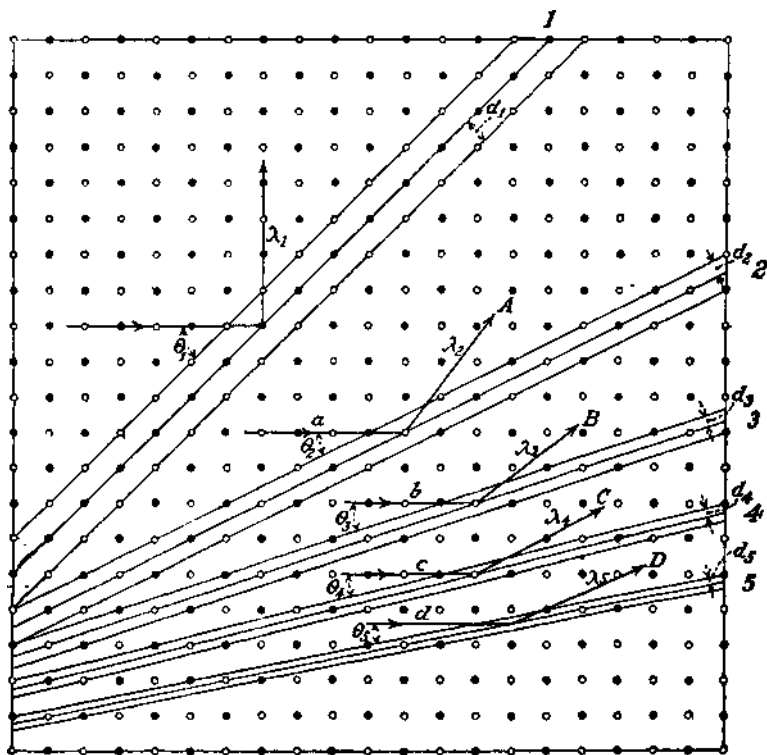


FIG. 148.—Schematic representation of the reflection of monochromatic beams of X-rays by a crystal of NaCl when a heterochromatic beam is incident upon it.

in different phases and so will destroy each other by interference. Only if certain conditions as to wave length and angle of incidence of the beam on the planes are satisfied will the waves from different planes combine in the same phase and reinforce each other. The necessary conditions are easily found.

In Fig. 149, let the horizontal lines represent the traces of two successive Bragg planes spaced d apart. Denote by θ_i the angle between the direction of propagation of the incident beam and the planes; this angle is called the *glancing angle* of the beam on these planes. Let a

ray meet the two planes at O_1 and O_2 , respectively, and let a line drawn from O_2 perpendicular to the planes cut the other plane at O . Draw O_1O and O_2O representing rays specularly reflected from the two planes, and draw aOb perpendicular to O_1O and O_2O to represent a wave front of the reflected beam. Then constructive interference will occur if the path O_1O_2, b taken by waves scattered at O_2 , exceeds the path O_1a for waves scattered at O_1 by an integral number of wave lengths. Draw Oc perpendicular to O_1O_2 . Then $O_1a = O_1c$; hence the differ-

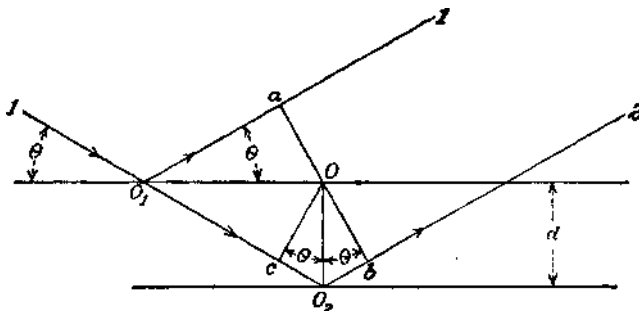


FIG. 149.

ence in path is $cO_2b = 2d \sin \theta$. The conditions that there should be a reflected beam are, therefore,

$$\theta = \theta' \quad (281)$$

and

$$n\lambda = 2d \sin \theta, \quad (282)$$

where n is an integer, called the "order of the reflection," and θ' is the angle between the Bragg planes and the direction of the diffracted rays. These two conditions are known as *Bragg's law* for X-ray reflection. The first condition is often omitted in the statement of Bragg's law with the understanding that it is implied by the term "reflection."

Suppose, now, that a parallel wave train, containing a *continuous spectrum* of wave lengths, is incident upon a crystal, as represented by the parallel arrows a, b, c, d in Fig. 148. In the figure, traces of five families of Bragg planes are shown, numbered 1, 2, 3, 4, 5, with their characteristic spacings $d_1, d_2, \dots$. Many other families of planes might be imagined, some perpendicular and some not perpendicular to the plane of the paper. Suppose that in the incident beam there is a wave length λ_2 such that

$$n\lambda_2 = 2d_2 \sin \theta_2,$$

where n is an integer, d_2 is the distance between the set of planes numbered "2," and θ_2 is the glancing angle between the direction of

the incident radiation and these planes. Then there will be reflected from this group of planes a beam A , of wave length X_2 , which will proceed in the direction of the arrow A . Similarly, we may have reflected beams $B, C, D, \dots$ in different directions in the plane of the paper, and also many other beams reflected from other families of planes in directions not in the plane of the paper. Each "Laue spot" in the experiment of Friedrich and Knipping may be interpreted as produced by such a reflected beam. In general, the most intense spots correspond to reflections from Bragg planes containing the greatest number of atoms on each plane.

The crystal represented in Fig. 148 is one of a very simple type, but the conception of Bragg planes is applicable to all types of crystals. It should be pointed out, however, that it is not necessary to draw the Bragg planes actually *through* the atoms; instead of a given family of these planes, any other set of planes parallel to them and spaced the same distance apart could be employed and would lead to the same conditions for strong reflection.

It should be pointed out also that the Bragg equation (282) does not give a complete solution to the interference problems of X-rays scattered from a crystal. The equation predicts only the position of the center of the expected diffraction pattern for a given wave length and family of Bragg planes; nothing is said about the intensity distribution in this diffraction pattern.

175. The X-ray Spectrometer.—Immediately following the announcement by Friedrich, Knipping, and Laue of their successful experiment, many investigators took up the study of the new phenomenon. Among these were W. H. and W. L. Bragg,¹ to whom we are chiefly indebted for the early development of the X-ray spectrometer.

A spectrometer of the Bragg type is shown diagrammatically in Fig. 150(a). X-rays from the target T of an X-ray tube pass through two narrow slits S_1 and S_2 , a few hundredths or tenths of a millimeter wide, the edges of which are made of some material, such as lead or gold, which is very opaque to X-rays. This ribbon-shaped incident beam of X-rays I falls at a glancing angle θ on the cleavage face of a crystal K —rock salt, calcite, mica, gypsum, quartz, etc.—which is mounted on a table D , the angular position of which can be accurately read by verniers or micrometer microscopes. The reflected beam of X-rays, which makes an angle 2θ with the incident beam, enters, through the "window" w , an ionization chamber C by means of which the intensity of the reflected beam *msky* be measured. By suitably

¹ BRAGG, W. L., *Nature*, vol. 90, p. 410 (1912); BRAGG, W. H., and W. L. BRAGG, *Roy. Soc, Proc*, vol. 88, p. 428 (1913); BRAGG, W. H., *Nature*, vol. 91, p. 477 (1913).

turning the table D about the axis A , the incident beam may be made to strike the face of the crystal at any glancing angle θ . The ionization chamber C is mounted on an arm (not shown) by means of which the chamber can be rotated about the axis A so as to admit the reflected beam through the window w . For protection against stray scattered radiation, a third slit S_3 is attached to the chamber.

For photographic registration, the ionization chamber may be replaced by a photographic plate PP [Fig. 150(b)]. With the crystal

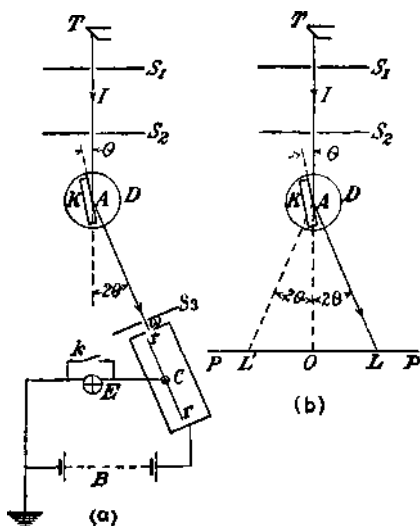


FIG. 150—The X-ray spectrometer, using (a) the ionization method, (b) the photographic method.

being similar to that shown in Fig. 148, which represents one plane of atoms. Taking the atomic weight of chlorine as 35.46 and of sodium as 23.00, we find the molecular weight of NaCl to be 58.46. Therefore, 58.46 grams of the NaCl contain $2N_0$ atoms, viz., N_0 atoms of Na and N_0 atoms of Cl, where N_0 is Avogadro's number. Thus, if we use the older value, $N_0 = 6.064 \times 10^{23}$, we find for the number of atoms n in 1 cm.-cube of rock salt

$$n = 2 \times 6.064 \times 10^{23} \times \frac{\rho}{58.46},$$

where $\rho = 2.163$, the density of crystalline NaCl. If d is the distance between the center of one atom and the next along the edge of the cube, $1/d$ is the number of atoms in a row of atoms 1 cm. long, and the

¹ See BRAGG, W. H., and W. L. BRAGG, "The Crystalline State," vol. I, 1933.

set at a glancing angle θ , the reflected beam will strike the plate at L (or at L' , if the crystal is "reversed"). From the position O at which the direct beam strikes the plate, the distances OL and OA and hence the angles 2θ and θ may be determined. The wave length λ is then obtained from the Bragg formula, $n\lambda = 2d \sin \theta$.

The distance d between the reflecting planes of a crystal such as NaCl is determined as follows. From his investigations, Bragg showed¹ that, in the rock-salt crystal, the Na and the Cl atoms or ions occupy alternate positions at the corners of elementary cubes in the cubic lattice characteristic of the crystal, the arrangement

number of atoms in the centimeter-cube is

$$n = \frac{1}{d^3}.$$

Solving these two equations simultaneously for d , which is the desired distance between the (cleavage) planes in NaCl, we find

$$d = 2.814 \times 10^{-8} \text{ cm. (Old value.)}$$

The value of d is seen to be dependent on M , the molecular weight of rock salt; on p , the density of rock salt; and on Avogadro's number, N_0 . The values of N_0 and p are not known to much better than 0.1 percent. There is, hence, a corresponding uncertainty in the value of d . At present, measurements of X-ray wave lengths can be made with a precision manyfold greater than the best measurements of M , N_0 , or p .

It was, therefore, supposed to be expedient to adopt a new unit of length, called the "X unit" or "X. U.," which was *very nearly* 1×10^{-11} cm. but which was accurately defined by taking the grating space of NaCl at 18°C. as *exactly* 2,814.00 X.U. The grating space of any other crystal could then be determined from this arbitrarily chosen standard as follows. Let θ_{NaCl} be the angle at which a line of given wave length is reflected from a rock-salt crystal, and θ_c the angle at which the same line is reflected from some other crystal of grating space d_c . Then, from Bragg's law [Eq. (282)], we have

$$n\lambda = 2d_{\text{NaCl}} \sin \theta_{\text{NaCl}} = 2d_c \sin \theta_c,$$

from which d_c may be determined. In later work, calcite was adopted as the best crystal for practical use, but the standard unit for wave lengths was not changed.

Recently, however, it has been found that there was an error in the old value of e (Sec. 42) and in the value of N_0 deduced from it. If the present value, $N_0 = 6.023 \times 10^{23}$, is used in the preceding calculation, we find

$$d = 2.820 \times 10^{-8} \text{ cm. (New value.)}$$

Furthermore, it has been found possible to measure X-ray wave lengths by means of ruled gratings, the grating space of which can be measured directly, so that no assumptions as to structure of a crystal or as to the value of N_0 are involved. It seems certain that in the future the absolute values of X-ray wave lengths will be based upon measurements with ruled gratings, crystals being used only to compare one wave length with another.

For this reason it has been proposed to redefine the X unit as being *exactly* 10^{-11} cm. Objection to this procedure has been raised recently by Siegbahn¹ on the ground that the accuracy attainable in the comparison of X-ray wave lengths by means of crystals will undoubtedly greatly exceed the accuracy attainable with a grating for a considerable time to come. Hence, if all wave lengths are now converted into absolute values on the basis of existing grating measurements, they will probably require further correction in the future, and the corrections are likely to exceed considerably the errors of the crystal measurements. It is considered preferable by Siegbahn and others to retain an artificial X unit based on the old assumed value of the grating space of calcite, and to express all wave lengths that are measured with crystals in terms of this unit. The true wave length in terms of 10^{-11} cm. as the unit can then always be found by increasing all the crystal wave lengths by a certain percentage, by 0.20 percent according to present data. Confusion can be avoided by expressing all true wave lengths in units of the metric system or in angstroms. This procedure will be followed in the present edition of this book. For many purposes the difference between the two units is immaterial.

There are two important corrections which must be made when, by use of a crystal grating, X-ray wave lengths are compared with a precision of the order of a few parts in 100,000. First, because of the thermal expansion of crystals, the grating space d varies with temperature, and correction must be made if the temperature differs from 18°C, which is the temperature to which grating spaces are usually referred. Second, it has been found that a beam of X-rays is slightly deviated by refraction as it enters or leaves a crystal (see Sec. 187 below). Accordingly the *observed* value of the glancing angle θ of the Bragg formula is not quite the same as the angle, *within* the crystal, at which the beam of rays strikes the Bragg planes.

In Table I are listed some of the crystals commonly used in X-ray spectroscopy together with their respective grating spaces. In the second column, the spaces are listed as given by Siegbahn,² on the basis of $d = 2,814$ X.U. for rock salt; in the third column, headed "corrected/" the spaces are given as found by multiplying Siegbahn's values by $3.03560/3.02945$ so as to bring them into harmony with the grating spacing of calcite ($d = 3,035.60$ X.U.) as calculated by Bearden from observations of the diffraction by a calcite crystal of X-rays whose wave length had been measured with a ruled grating.³

¹ SIEGBAHN, *Nature*, vol. 151, p. 502 (1943).

² SIEGBAHN, "Spektroskopie der Röntgenstrahlen," 2d ed., 1931.

³ BEARDEN, *Phys. Rev.*, vol. 48, p. 385 (1935).

It is quite beyond the scope of this book to discuss the various (ionization) spectrometers or (photographic) spectrographs which have been developed for the measurement of X-rays. Some, *e.g.*, the Bragg spectrometer, use a single plane crystal and "external" reflection, some use transmission and "internal" reflection, some use two plane crystals in succession, and some use a bent crystal in order to focus the reflected rays. For details of the various designs and of their operation, the reader is referred to other books.¹

TABLE I.—GRATING SPACES OF SOME CRYSTALS USED IN X-RAY SPECTROSCOPY

Crystal	Grating space d at 18°C.		Change in d per degree centigrade, X.U. or 10^{-11} cm.
	Siegbahn, X.U.	Corrected, $\times 10^{-11}$ cm.	
Rock salt (NaCl)	2,814.00	2,819.71	0.11
Calcite (CaCO_3)	3,029.45	3,035.60	.03
Quartz (SiO_2)	4,246.02	4,254.65	.04
Gypsum ($\text{CaSO}_4 \cdot 2\text{H}_2\text{O}$)	7,584.70	7,600.1	.29
Mica	9,942.72	9,962.9	.15

It is obvious from Bragg's law,

$$n\lambda = 2d \sin \theta,$$

that the maximum wave length λ_m measurable by use of a given crystal is $2d$. Actually, λ_m is somewhat less than this, since the glancing angle θ cannot usefully exceed about 70° . By use of a quartz crystal, for example, wave lengths up to about 7,500 X.U. can be measured. For measurement of extremely long wave lengths, certain organic crystals of large grating space were originally used (*e.g.*, sugar, $d = 10.57$ Å; lead melissate, $d = 87.5$ Å). In recent years, however, it has been found more expedient to use a ruled grating for this purpose. There is no such upper limit to the wave lengths which may be measured by means of a ruled grating. In general, crystals are now used from very short wave lengths up to about 50 Å, and gratings from very long wave lengths to somewhat below 50 Å.

176. Bragg's Discovery of Monochromatic Characteristic Radiations.—With a beam of X-rays from a platinum target incident on the cleavage face of a rock-salt crystal, as shown in Fig. 150(a), W. H.

¹ SIEGBAHN, *op. cit.* See particularly the two-crystal spectrometer, pp. 128-134, an instrument of very high resolving power. Cf. also, Compton and Allison, *op. cit.*, pp. 750-756, for focusing spectrographs.

Bragg¹ rotated the crystal in steps of A0 and the ionization chamber in steps of 2A0. He plotted the curve of ionization current against glancing angle θ and found that the current or X-ray intensity did not vary uniformly with angle but rose at certain angles to a sharp maximum. A curve similar to that shown in Fig. 151 was obtained. A group of three maxima, a_1 b_1 and c_1 was observed at the respective angles θ of 9.9, 11.6, and 13.6°. A second group of three maxima, a_2 , b_2 , and C_2 , was observed at approximately double these angles.

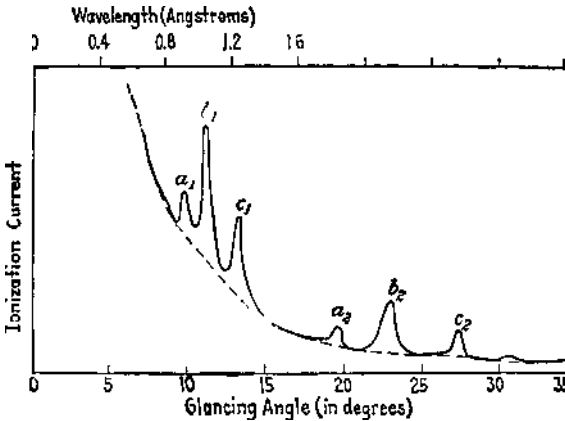


FIG. 151.-- Bragg's curve for the energy distribution in an X-ray spectrum, showing the characteristic lines a , b , c

This second group is similar, as to relative intensities of the maxima, to the first group. Bragg interpreted the maxima a_1 b_1 , and C_1 as three monochromatic lines; and the second group of maxima a_2 , b_2 , and C_2 as second-order reflections of the lines a_1 b_1 , and c_1 . He computed their wave lengths by the formula $n\lambda = 2d \sin \theta$, taking $n = 1$ for the "lines" a_1 b_1 and C_1 and $n = 2$ for the second-order lines a_2 , b_2 , and c_2 . A third-order peak b_3 was observed at 36.6°. Assuming $d = 2.814$ angstroms for rock salt, he found the following wave lengths:

TABLE II.—FIRST MEASUREMENTS OF X-RAY WAVE LENGTHS BY BRAGG
(PLATINUM TARGET)

Line	θ , degrees	$\sin \theta$	n	λ , angstroms
a_1	9.9	0.172	1	0.97
b_1	11.6	.200	1	1.13
b_2	23.6	.400	2	1.13
b_3	36.6	.597	3	1.12
c_1	13.6	.235	1	1.32

¹ *Nature*, Jan. 23, 1913.

Curves similar to Fig. 151 were obtained with other crystals, the only difference being that the maxima occurred at different glancing angles, indicating that each crystal had a characteristic grating space d . Bragg convinced himself, however, that these respective maxima for different crystals always represented the same monochromatic radiation, since, for example, the absorption in aluminum of peak b_1 was always the same, whatever the crystal used. In short, the peaks of the curve in Fig. 151 represent *spectral lines* the wave lengths of which are *characteristic of the target emitting the rays*. These monochromatic lines are superimposed on a *continuous* spectrum represented by the partially dotted line in the figure. Curves of the type shown in Fig. 151, therefore, represent (subject to certain corrections to be mentioned later) the distribution of *energy* in the X-ray spectrum, continuous and characteristic combined, of an element.

177. Moseley's Law.—In two classic papers,¹ Moseley presented a systematic study of the characteristic radiations emitted by various

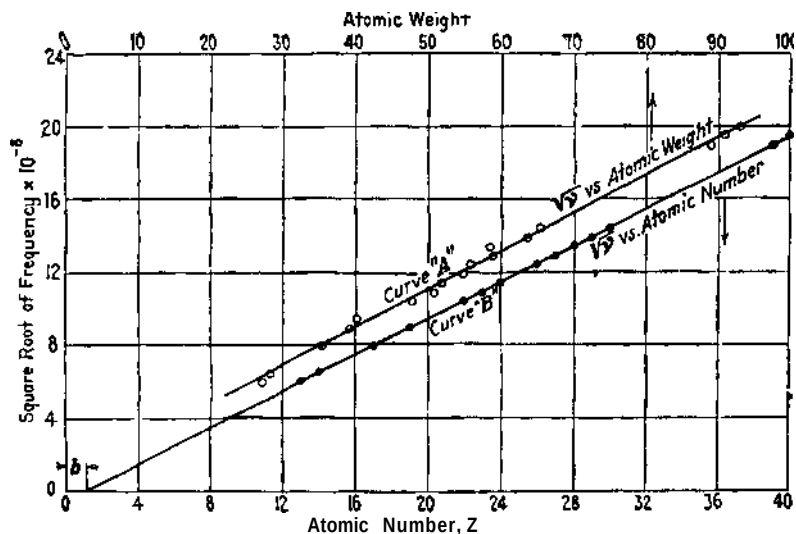


FIG. 152 —Moseley's curve showing the relation between the frequencies of X-ray lines and atomic number.

targets, using a photographic method similar in principle to that shown in Fig. 150(6). He found a larger number of characteristic lines than Bragg and, also, that these lines could, in general, be classified into two groups: (1) a group of shorter wave lengths, which, by means of the value of absorption coefficients in aluminum, he identified with Barkla's K characteristic secondary radiations and (2)

¹ MOSELEY, *Phil Mag.*, vol. 26, p. 1024 (1913); vol. 27, p. 703 (1914).

a group of lines of longer wave length, similarly identified with Barkla's L radiation.

Unlike the optical spectra, however, the X-ray characteristic spectra of the elements were found to be similar from element to element, homologous lines occurring, in general, at shorter wave lengths the greater the atomic weight of the element in which the lines originate. In searching for a relation between the frequency of a given line, say the $K\alpha$ line, and some property of the atom in which the line originated, Moseley first observed that the frequency did not vary uniformly with the atomic weight, as is shown by curve A (Fig. 152) in which the square root of the frequency is plotted against the atomic weight. In Bohr's theory of the origin of spectra, however, which at that time had recently been proposed, the charge on the nucleus played a fundamental role. According to Bohr's theory, the frequency ν of a spectral line is given by

$$\nu = Z^2 \frac{2\pi^2 mc^4}{h^3} \left(\frac{1}{n_2^2} - \frac{1}{n_1^2} \right)$$

[Sec. 95, Eq. (140a)], from which

$$\sqrt{\nu} \propto Z,$$

Z being the atomic number. Rutherford had shown, from his experiments on the scattering of α particles, that the value of the nuclear charge, for a given atom, is very approximately one-half the atomic weight; and Barkla had shown, from experiments on the scattering of X-rays, that the number of electrons surrounding the nucleus is also approximately one-half the atomic weight. Guided by these considerations, Moseley assigned atomic numbers Z to the elements he had investigated and then plotted a curve between $\sqrt{\nu}$ and Z . Such a plot of Moseley's data for the $K\alpha$ line is shown in Fig. 152, curve B. The graph is seen to be a straight line, with a small intercept $b = 1$ on the Z -axis. It is obvious from a comparison of the two curves A and B of Fig. 152 that, as far as the determination of the frequency of characteristic lines is concerned, atomic number is a much more fundamental quantity than atomic weight.

Empirically, the relation between the frequency ν of the $K\alpha$ line and Z , as determined by Moseley from Fig. 152, curve B, is

$$\nu = 0.248 \times 10^{16} (Z - 1)^2. \quad (283)$$

In Bohr's equation for the frequency ν of a spectral line, as just

written, if we set $n_2 = 1$ and $n_1 = 2$, we obtain for ν

$$\nu = 0.328 \times 10^{16} \times Z^2 \left(\frac{1}{1^2} - \frac{1}{2^2} \right) \quad (284a)$$

$$= 0.246 \times 10^{16} \times Z^2 \quad (284b)$$

by inserting the numerical values of m , e , and h . Except for the slight correction to Z , Eqs. (283) and (284b) are seen to be almost identical.

This agreement suggests rather naturally the hypothesis described in the next section as to origin of the *Ka* line.

178. The Origin of X-ray Lines, (a) *Emission Spectra.*—In Sec. 119, we have seen that in an atom there are 2, and only 2, electrons corresponding to the electronic quantum number, $n = 1$. These electrons, which constitute the *K* shell and may be called the *K* electrons, have (in the zero-order stage of perturbation theory) 1-electron wave functions whose values are very small except close to the nucleus. The electrons in question are, therefore, very unlikely to be found at any considerable distance from the nucleus and may be regarded as the innermost electrons in the atom. In any neutral atom of atomic number $Z > 2$, there exist also 1 or more, up to a maximum of 8, electrons in quantum states with $n = 2$, called "*L* electrons." Their wave functions extend effectively to much greater distances from the nucleus than those of the *K* electrons, so that, although any one of the *L* electrons might be found closer to the nucleus than the *K* electrons, they are most likely to be found at much greater distances. Then, if Z is large enough, there may occur, also, *M* electrons, with $n = 3$, characterized by wave functions extending still farther out, and so on, until all of the circumnuclear electrons are accounted for.

Now suppose an atom in the target of an X-ray tube is bombarded by an energetic cathode ray and that 1 of the two *K* electrons is "knocked out" of the atom. The atom is thereby converted into an ion, and the ion is left in a quantum state characterized by the absence of one *K* electron. This state of the ion may be called a *K quantum state of the ion*. Subsequently an *L* electron may "drop" into the *K* vacancy, i.e., change from an electronic state with $n = 2$ to one with $w = 1$. The ion thereby undergoes a transition from a *K* state to an *L* state, its excess energy being emitted as a quantum of radiant energy. Atomic processes of this sort are assumed to give rise to the *Ka* line. According to this assumption, the quantum $h\nu$ of *Ka* radiant energy is equal to the difference in energy between the *K* and the *L* quantum states or levels of the ionized atom.

In a similar way, the *K β* line is assumed to originate when an *M*

electron left near the nucleus. This electron "screens" the nucleus and makes its *effective* nuclear charge about 1 unit less; hence, the factor ($Z - 1$). Further work has shown, however, that Moseley's law holds only as a first approximation.

After the work of Bragg and Moseley, the techniques of X-ray spectrometry developed rapidly. Many new characteristic lines were discovered, and at the present time it is believed that practically the entire spectrum of most elements has been mapped out. Following and extending the notation of Barkla, the lines are classified into series known as the *K*, *L*, *M*, *N*, etc. series, respectively; the letter designating a series refers to the initial atomic (ionic) state in the transitions giving rise to the lines in question. The principal lines in the *K* series and *L* series from tungsten are plotted in Fig. 153, as explained under the figure. Each series of lines contains several strong lines and numerous faint ones. To illustrate further the appearance of X-ray spectra, there is shown in Fig. 154 a section of the *L* series of silver, as recorded with a two-crystal ionization spectrometer of high resolving power and dispersion.

Many faint X-ray lines, called "satellite lines," originate in atomic transitions between states of double ionization and will be discussed in more detail in a later section [190(a)]. Except in that section, the discussion will be confined to "first-order lines," arising in the manner just described from transitions between states of *single* ionization.

In Table III(a) are listed the wave lengths of the four principal lines of the *K* series, the a_1 , a_2 , *P*, and γ lines, for a number of elements.

It is more convenient from the theoretical standpoint, however, to work with quantities proportional to the frequencies of the lines instead of with their wave lengths; for this purpose values of ν/R are commonly employed, ν being the frequency and *R* denoting the Rydberg constant for an atom of infinite mass (in the same units as ν). Values of ν/R for the same *K* lines are shown in Table 111(6). They may be obtained most quickly by taking the reciprocal of the wave length expressed in X.U., dividing this by the value of *R* in terms of cm.^{-1} , and multiplying by 10^{11} .

It is to be noted that the phrase "series of lines" as used in X-ray spectroscopy refers to a group of lines arising from a common *initial* atomic state, whereas in the part of spectroscopy that deals with the outermost electrons of the atom the term "series" is applied to a group of lines having a common *final* atomic state.

(b) *Absorption Spectra*.—The theoretical picture that we have described furnishes, furthermore, a simple explanation of the discontinuities that Barkla had observed in the absorption coefficient of

various materials for X-rays. Absorption, as we saw in studying atomic spectra, is the reverse of emission. If a photon of $K\alpha$ radiation is emitted when an electron drops from an L shell into a K shell, then

TABLE III(a).—WAVE LENGTHS IN X UNITS OF THE MORE PROMINENT K LINES OF SOME REPRESENTATIVE ELEMENTS*

Element	Z	$\gamma^\dagger$	$\beta^\dagger$	α_1	α_2	λ_K
S	16		5,021.1	5,361.3	5,363.7	5,008.8
Ca	20		3,083.4	3,351.69	3,354.95	3,064.3
Fe	26		1,753.01	1,932.08	1,936.01	1,739.4
Zn	30	1,281.07	1,292.55	1,432.17	1,436.03	1,280.5
Br	35	918.53	930.87	1,037.59	1,041.66	918.09
Mo	42	619.70	630.98	707.83	712.11	618.48
Ag	47	486.30	496.01	558.28	562.67	484.48
I	53	374.71	383.15	432.49	437.03	373.44
W	74	178.98	184.22	208.62	213.45	178.22
Pb	82	141.25	146.06	165.16	170.04	140.49
U	92	108.42	111.87	126.40	130.95	106.58

* See Note 1 below.

† See Note 2 below.

TABLE III (b).— ν/R VALUES OF THE MORE PROMINENT K LINES OF SOME REPRESENTATIVE ELEMENTS*

ν = frequency; R = Rydberg constant for an atom of infinite weight

Element	Z	$\gamma^\dagger$	$\beta^\dagger$	α_1	α_2	$(\nu/R)_K$
S	16		181.49	169.97	169.90	181.93
Ca	20		295.54	271.88	271.62	297.38
Fe	26		519.83	471.65	470.69	523.90
Zn	30	711.34	705.02	636.29	634.58	711.67
Br	35	992.10	978.95	878.25	874.82	992.57
Mo	42	1,470.52	1,444.23	1,287.42	1,279.69	1,473.4
Ag	47	1,874.96	1,837.20	1,632.29	1,619.55	1,880.9
I	53	2,431.9	2,378.4	2,107.0	2,085.2	2,440.2
W	74	5,091.4	4,946.6	4,368.2	4,269.3	5,113
Pb	82	6,452	6,239	5,518	5,359	6,486
U	92	8,405	8,146	7,209	6,959	8,550

* See Note 1 below.

† See Note 2 below.

it should be possible for a photon of this frequency to be absorbed by an atom while one of the K electrons is raised into an L shell—provided there is a vacancy in the L shell into which it can go. As we have seen, however, in any atom with $Z > 10$, the L shell normally

TABLE III(r).— ν/R VALUES OF THE L ABSORPTION LIMITS OF CERTAIN ELEMENTS*

Element	Z	L_I	L_{II}	L_{III}
Mo	42	212 43	193 39	185 81
Ag	47	280 85	259 91	246 76
I	53	382 26	357 71	335 78
W	74	890 37	850 28	715 30
Pb	82	1,166.47	1,119 05	960.03
U	92	1,604.3	1,541 0	1,264.2

* See Note 1 below.

TABLE 111(d).— ν/R VALUES OF THE M ABSORPTION LIMITS OF CERTAIN ELEMENTS*

Element	Z	M_I	M_{II}	M_{III}	M_{IV}	M_V
W	74	208.77	189.95	167 9	140.5	136.0
Pb	82	283.7	262 1	225 9	192 0	184.3
U	92	408.9	382.1	316 7	273.9	261.0

* See Note 1.

Note 1.—The wave lengths and ν/R values in these tables are taken from Siegbahn,¹ and are expressed in conventional X units. To obtain values in terms of 10^{-11} cm. as a unit, the wave lengths must all be increased, or the ν/R values decreased, by 0.20 per cent.

Note 2.—The $A\gamma_3$ line is really a close doublet, experimentally resolved for Rb 37 and elements of higher atomic number. The two components are called (Siegbahn) β_1 and β_2 , β_1 being of shorter wave length. For Ag 47, for example, the wave lengths are $\lambda\beta_1 = 496.01 \text{ X.}$ $\lambda\beta_2 = 496.65 \text{ X. U.}$ In the table, the wavelength of β_1 is given for elements for which the line is resolved. The γ line is also a close doublet.²

contains as many L electrons as can get into it. Hence for such atoms the $K\alpha$ lines cannot actually be observed as absorption lines. The same happens to be true, for analogous reasons, of all of the X-ray emission lines that are commonly observed. A photon *can* be absorbed, however, if it has enough energy to remove an inner electron from the atom entirely. Let W_K denote the energy required to remove a K electron and to leave it at rest outside the atom. Then a photon of frequency ν can eject a K electron provided $\nu >: \nu_K$ where

$$h\nu_K = W_K.$$

The explanation of Barkla's absorption curves is now clear. As ν is progressively increased, at the frequency $\nu = \nu_K$ the absorption suddenly increases, since absorption in the K shell then begins; and thereafter this absorption continues for all larger values of ν . The curve representing the absorption coefficient plotted against frequency

¹ *loc. cit.*
² See HUDSON and VOGT, *Nat. Acad. Set., Proc.*, vol. 19, p. 447 (April, 1933).

or wave length will, therefore, show a sudden rise toward the side of shorter wave length at $\nu = \nu_K$. This is what Barkla observed, as is illustrated in Fig. 145, where the abscissa can also be regarded as representing wave lengths plotted on a certain scale, with shorter wave lengths toward the left. The critical wave length λ_K , or frequency ν_K , at which absorption in the K shell begins is called the *K absorption limit*. Values of XK and of ν_K/R are listed in the last columns of Tables III(a),(b). Since, in the theoretical picture, the energy W_K required to remove a K electron is somewhat greater than the energy change when the ionized atom undergoes a transition between its K state and another of its discrete quantum states, it is to be expected that $h\nu_K$ will be greater than the corresponding quantity for any of the K lines. Hence ν_K should exceed ν for any of these lines, but all of these frequencies should be of the same order of magnitude. These conclusions are in agreement with the observations of Barkla.

The intimate connection between the emission lines of the K series and the K absorption limit is further shown by the critical voltage which must be applied to the X-ray tube in order to generate these lines. The critical voltage V_K is found to be determined by the relation

$$eV_K = h\nu_K,$$

where ν_K is the frequency of the K absorption limit for the material of which the target is composed, h is Planck's constant, and e is the (numerical) electronic charge. Furthermore, *all of the lines of the K series are excited at the same critical voltage V_K* , and, as the voltage is increased above V_K , these lines all increase in intensity at exactly the same rate. In the case of tungsten, for example, for which $A_K = 178.22$ X. U., we find from the last equation, or from Eq. (1476) in Sec. 99, that $V_K = 69,300$ volts. When voltages higher than this are applied to a tungsten-target X-ray tube, all of the X-series lines of tungsten appear. The explanation of these facts, in terms of the picture described above, is obvious.

Similar facts and interpretations hold for each of the other series of characteristic X-ray lines, the L , M , N , $\cdot \cdot \cdot$ series. A significant difference, however, is that here more than one absorption limit exists in each case. Thus, *three L absorption limits* are found, denoted, in order of decreasing frequency, by L_I , L_{II} , and L_{III} . There are *five M limits*, *seven N limits*, etc. The theoretical reason for the existence of these multiple limits will be discussed in Sec. 180. The N and O limits have very long wave lengths, however, and can be observed only for the heaviest atoms. The ν/R values of some L and M limits are listed in Tables III(c),(d).

(c) *Fluorescent Spectra*.—Finally, the *fluorescent radiation* that Barkla observed to be emitted when X-rays fall upon material bodies is likewise easily explained in terms of the theoretical picture. A *K* electron may be ejected from the atom as the result of an absorption process, as well as by cathode-ray bombardment. Such an act of absorption leaves a vacancy in the atom which may be filled subsequently by an electron dropping into it from the *L* shell, or from the *M* shell, and so on. Thus, as with cathode-ray excitation, characteristic radiation belonging to the entire *K* series of lines appears in fluorescence whenever any member of the series appears. The frequency of this fluorescent radiation should be lower than the frequency of the original incident radiation, in agreement with observation. Fluorescence is a comparatively unusual phenomenon in the visible region, but in the case of X-rays it is very common.

179. *X-ray Energy Levels*.—As in the spectroscopy of longer wave lengths, X-ray spectra are conveniently interpreted in terms of a set of *energy levels*, corresponding to the various quantum states of the atom. In the emission of a *K α* line, for example, an (ionized) atom undergoes a transition or "jumps" from a *K* energy level to an *L* level, the difference in energy between these levels being emitted as a *K* photon. Although we often speak of an electron as making the jump, it is to be emphasized that the energy of each state or level involved in the transition really belongs to *the atom as a whole*. It would not be correct to think of this energy as belonging to a single electron, and it would be an unnecessary complication to endeavor to introduce a set of *electronic* energy levels.

When a *K* electron is removed from a neutral atom, the atom is both ionized and raised into an atomic *K* level of much higher energy. The difference in energy between the *K* level of the ion and the normal level of the neutral atom is equal to the work required to remove a *K* electron and leave it at rest at an infinite distance from the atom; thus it equals what we have called *W_K* or $h\nu_K$, where ν_K is the frequency of the *K* absorption limit. Similarly, there will be three slightly different *L* levels having energies greater than that of the normal level by amounts equal to the work of removing an *L* electron. Below the *L* levels, again, is a group of five *M* levels, associated with the removal of an *M* electron; and so on. Finally, just above the normal state of the neutral atom lies an extensive set of optical levels connected with the emission of the lines in the arc spectrum; and above this set of levels lies another set belonging, like the X-ray levels, to the ionized atom and associated with the *spark* spectrum in the visible and ultraviolet regions. Ordinary X-ray spectra constitute, in a sense, an extension of the spark spectrum to higher levels of energy, resulting

from an electronic vacancy in the interior of the ionized atom rather than from the excitation of one of its outer electrons.

The typical distribution of X-ray atomic levels is illustrated diagrammatically in Fig. 155. Lines indicating transitions between X-ray levels on such diagrams are often drawn with an arrowhead at each end, to indicate that, in a sense, an electron moves upward on the diagram as the atom moves downward. Thus, in the emission of a $K\alpha$ line, the atom drops from its K level to one of its L levels, an

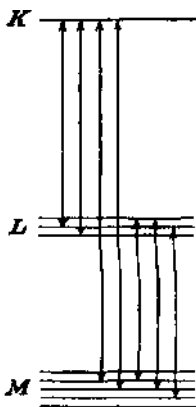


FIG. 155.—
Typical arrange-
ment of the higher
X-ray energy levels.

electron simultaneously changing from an L shell to a K shell; for an $L\alpha$ line, an electron changes from M to L ; and so on. The $K\beta$ line results from the transition of the atom from the K to an M level; the loss of energy being larger, the $K\beta$ line is of higher frequency than the $K\alpha$ line. The $K\gamma$ line results from a transition, K to N . The transition from K to the normal state, representing capture of a free electron of negligible initial kinetic energy into the vacancy in the K shell, would cause the emission of a frequency ν_K equal to that of the K absorption limit; this would represent the "series limit" for the K lines.

In a similar way, the lines of the L series are emitted by atoms in which a vacancy in the L shell has been caused by the impact of a cathode-ray electron or by the absorption of a photon. The atom then drops from its L energy level to the M level, or to some lower level, emitting a photon of radiation belonging to one of the L lines. Lines belonging to M , N , . . . series may similarly occur.

The energy-level diagram suggests at once a simple quantitative relation that should exist between *line frequencies* and *absorption limits*. It is obvious that the loss of energy by the atom when a $K\alpha$ line is emitted, i.e., the $h\nu$ value of this line, is equal to the *difference* in the energies, or in the $h\nu$ values, of the K and L *absorption limits*. The same equality will exist for the frequencies, or for the values of ν/R . This is a conclusion that can easily be checked from the data. In columns 2 and 3 of Table IV are shown the wave lengths and the values of ν/R for, respectively, the K , L_I , L_{II} , and L_{III} absorption limits of molybdenum [see Tables III(b),(c)]. In column 4 are shown the differences in ν/R values between the K limit and each of the three L limits. The second part of the table shows the directly observed values of λ and of ν/R for the $K\alpha_1$ and the $K\alpha_2$ lines of molybdenum. It is seen that the *observed value of ν/R for the $K\alpha_1$ line is almost exactly*

equal to the difference between the value of ν/R of the K limit and that of the L_{II} limit. Similarly, the ν/R value of the $K\alpha_2$ line is equal to the

TABLE IV.—RELATION OF THE FREQUENCIES OF THE $K\alpha_{1,2}$ LINES OF MOLYBDENUM TO THE FREQUENCIES OF THE K AND THE L ABSORPTION LIMITS

Absorption limits			$(\nu/R)_K - (\nu/R)_L$	Observed emission lines		
Limit	λ , X units	ν/R		Line	λ , X units	ν/R
K	618.48	1,473.4				
L_I	4,289.7	212.4	(1,261.0)		Missing	
L_{II}	4,712.0	193.4	1,280.0	$K\alpha_2$	712.10	1,279.7
L_{III}	4,904.2	185.8	1,287.6	$K\alpha_1$	707.83	1,287.4

difference between the value of ν/R of the A'' limit and that of the L_{II} limit. Putting this relation in symbols, we can write

$$\left(\frac{\nu}{R}\right)_{K\alpha_1} = \left(\frac{\nu}{R}\right)_K - \left(\frac{\nu}{R}\right)_{L_{III}},$$

$$\left(\frac{\nu}{R}\right)_{K\alpha_2} = \left(\frac{\nu}{R}\right)_K - \left(\frac{\nu}{R}\right)_{L_{II}};$$

or

$$\nu_{K\alpha_1} = \nu_K - \nu_{L_{III}}, \quad \nu_{K\alpha_2} = \nu_K - \nu_{L_{II}}.$$

This example illustrates a law of great importance: *The frequencies of X-ray lines are given by the differences between the frequencies of absorption limits.* By means of this law it is possible to infer, from data on the lines, the relative positions of certain absorption limits which cannot be directly measured.

When absorption limits can be observed, on the other hand, they enable us, as we have seen, to locate the various energy levels relative to the normal state of the atom. Thus, from data on X-ray lines and absorption limits, the whole scheme of X-ray atomic levels can be built up, and this scheme then serves to represent a wide range of facts in a compact and clear form.

180. The Quantum Theory of X-ray Terms and Lines.—The simple picture of the atomic mechanism for the production of characteristic X-rays that was described above affords no explanation of the occurrence of *several different* L , M , . . . levels lying close together. This feature was not correctly understood until wave mechanics was developed. The basis for its explanation is furnished by the theory of atomic structure described in Sees. 118 to 120 and 133. As was there stated, however, the scope of this book is such that only a brief description of the theoretical results can be given here.

As explained in Sec. 118 or 120, to a first approximation the electrons in an atom can be supposed to be distributed, in accordance with the Pauli exclusion principle, among a number of possible 1-electron quantum states, each corresponding in wave mechanics to a different possible form of the electronic wave function. Each of these states can be chosen so as to be characterized by four quantum numbers (Sec. 133), $n l j m$. Such states are appropriate when, as in the interior of atoms, the electronic spin-orbit interactions are relatively strong. The first quantum number, n , is always a positive integer; the values $n = 1, 2, \dots$ correspond to the various shells, which we have also denoted by $K, L, M, \dots$. The number l is 0 or a positive integer less than n ; it serves to designate a subshell. The number j can take on, for given $l > 0$, either of two values, $j = l + \frac{1}{2}$ or $l - \frac{1}{2}$; and m may have any one of the $2j + 1$ integrally spaced values from $m = j$ down to $m = -j$. If $l = 0$, $j = \frac{1}{2}$. For convenience, there is shown below

Shell, ...	<i>K</i>	<i>L</i>		<i>M</i>				<i>N</i>			
<i>n</i>	1	2		3				4			
<i>l</i>	0	0	1	0	1	2	0	1	2	3	
<i>j</i>	$\frac{1}{2}$	$\frac{1}{2}$	$\frac{1}{2}$ $\frac{3}{2}$	$\frac{1}{2}$	$\frac{1}{2}$ $\frac{3}{2}$	$\frac{3}{2}$ $\frac{5}{2}$	$\frac{1}{2}$	$\frac{1}{2}$ $\frac{3}{2}$	$\frac{3}{2}$ $\frac{5}{2}$	$\frac{5}{2}$ $\frac{7}{2}$	
Electrons . . .	2	2	2 4	2	2 4	4 6	2	2 4	4 6	6 8	

the constitution of the first four shells when completely filled, the X-ray designation of each shell being given opposite "Shell," and the total number of electrons that have the same set of values of n, l, j being given opposite "Electrons."

Let us fix our attention now on a single subshell of a normal, neutral atom. This subshell is specified by a particular choice of n and l (e.g., $n = 3, l = 1$, the subshell containing 6 electrons). We wish to consider the effect of removing 1 electron.

When the subshell is full of electrons, there is only one possible choice that can be made of 1-electron quantum states to contain the electrons; it is necessary to use all of the electronic quantum states that are available for the given values of n and l . Under such circumstances the total angular momentum of all of the electrons in the subshell is easily shown to be zero; this is true likewise of the total orbital momentum of the electrons and of their spin momenta, and also for the sum of the angular momenta of both kinds.

If, now, 1 electron is removed from the subshell, there is a variety of states in which the incomplete subshell can be left, corresponding to the different combinations of the electronic quantum states that

can be chosen **for** the remaining electrons. There are, in fact, just as many different possible states for the incomplete subshell as there are different electronic states represented in the complete subshell, each state of the incomplete subshell corresponding to the omission of one of these electronic states. The resultant angular momenta, orbital and spin, of the electrons remaining in the incomplete subshell can then be added into a resultant, which we may indicate, according to spectroscopic terminology, as in Sec. 130, by the quantum number J . In addition, wave-mechanical theory indicates that the orbital angular momenta can be added separately into a resultant, with a quantum number L , and their spin momenta into a separate resultant characterized by a quantum number S . Corresponding to what was just said about the number of possible states for the subshell, however, the sets of values of J , L , and S that can occur for the remaining electrons in the subshell are exactly the same as they are for a single electron existing all by itself in the subshell. For this reason, it is customary to use, instead of the capital letters J , L , and S , the small letters j , l , and s , which would be appropriate if only 1 electron were present. We can, in fact, think of the "hole" that is left by removing an electron as functioning like an electron, a conception that has far-reaching applications. The hole is sometimes thought of as formed by the omission of an electron which already possessed the particular values of l and j in question. This idea must not be taken too literally, however; it does not imply, for instance, that jj coupling necessarily holds in the inner subshells of atoms.

The removal of an electron thus leaves the subshell in a state characterized by certain values of n , l , j , and m . Theory indicates, however, that the energy should not vary with m , so that for many purposes we may ignore this latter quantum number and retain only n , l , and j . Furthermore, the energy, although varying greatly with the value of n , should vary only slightly with l and j . Thus, for an atom with a vacancy in a given shell (given n), there will exist a close group of energy levels whose number will be equal to the number of the possible sets of values of l and j . For $n = 1$ (K shell), the only values possible are $l = 0$, $j = \frac{1}{2}$; and there is only one level. For $n = 2$ (L shell), there is one level for $l = 0$, $j = \frac{1}{2}$, and also one each for $l = 1$, $j = \frac{1}{2}$ and $l = 1$, $j = \frac{3}{2}$, or three in all. For $n = 3$ (M shell), there are five possible levels, with $l = 0$, $j = \frac{1}{2}$; $l = 1$, $j = \frac{1}{2}$ or $\frac{3}{2}$; $l = 2$, $j = \frac{3}{2}$ or $\frac{5}{2}$ and so on. The number of the absorption limits as actually observed is thus explained.

Of course, the total **number** of levels that can occur **for** an **actual atom is limited also by the number of electrons it contains.** Among all

chemical types of atoms that occur naturally, the uranium atom will exhibit the maximum number of levels. The energy-level diagram for ionized uranium, with energy plotted as $\sqrt{v}/R$ on a logarithmic scale,

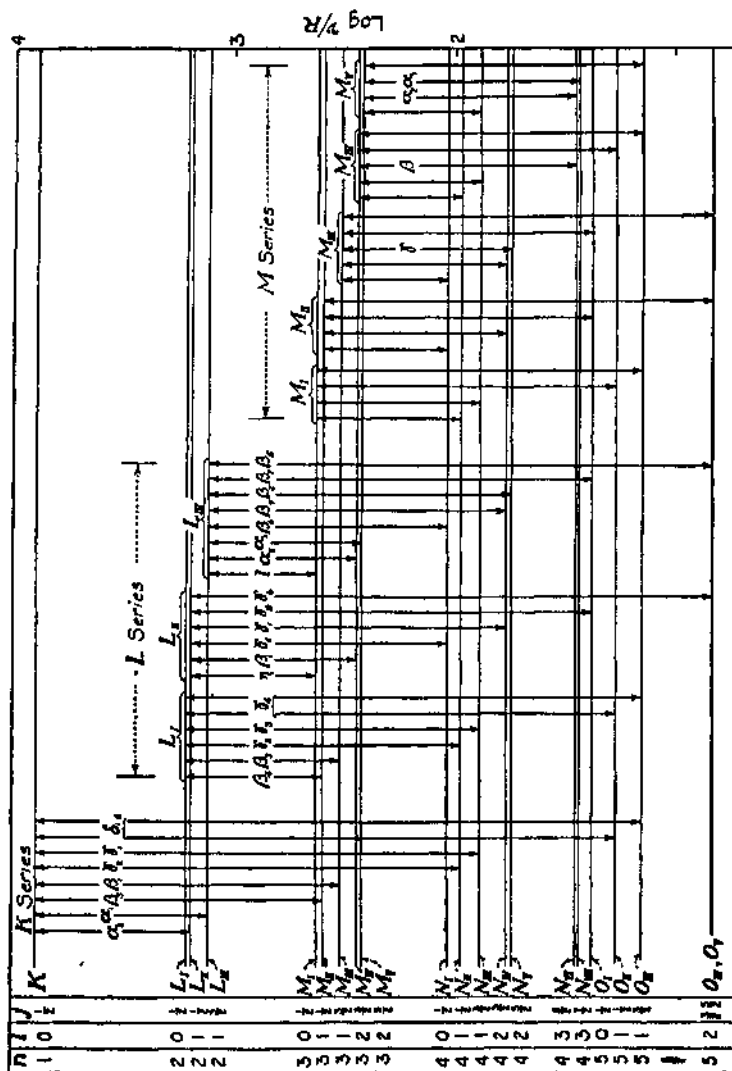


FIG. 156.—X-ray energy-level diagram for U(92), showing the transitions permitted by the selection rules $\Delta l = \pm 1$; $\Delta j = \pm 1, 0$.

is shown in Fig. 156. As one proceeds toward elements of lower atomic number, the electron shells of higher quantum number n progressively disappear (see Appendix III) and therewith, also, the corresponding energy levels, as well as the X-ray lines originating from these levels. In copper, for example, there are no O levels and only one N level.

One should not expect to find, therefore, in the X-ray spectrum of copper such lines (see Fig. 156 for their origin) as LB_5 , Lra , Ly_6 , etc.

In addition to accounting for the general features of the energy-level diagram, wave mechanics makes definite predictions concerning the *relative probabilities of transitions* between the various levels. For a transition accompanied by what is called radiation of the dipole type, we have the same selection rules as for 1 electron (Sec. 133):

$$\Delta l = \pm 1, \quad \Delta j = 0 \text{ or } \pm 1.$$

Experimentally, it is observed that between many pairs of levels transitions do not occur. For example, in Table IV, the emission line corresponding to the transition $K \rightarrow L_I$ is listed as "missing." In Fig. 156, also, many apparently possible transitions are not shown by arrows because the corresponding lines are not observed. For a long time, no reason was known for such apparent vagaries. Let us see whether or not we can account for them by means of the wave-mechanical selection rules.

We must first discover how values of l and j are to be assigned to such levels as L_I , L_{II} , L_{III} . This can be done by comparing the observed array of lines with those allowed by the selection rules and endeavoring to assign values of l and j in such a way as to bring the theoretical predictions into harmony with the facts.

Referring to the uranium diagram in Fig. 156, which is typical of that for all atoms except for the absence of certain outer levels as explained above, we note that transitions occur between the K level, with $l = 0$, and the L_{II} and L_{III} levels, giving rise to the $K\alpha_2$ and $K\alpha_1$ lines. Hence, it must be that $l = 1$ for L_{II} and L_{III} , in order to make $\Delta l = 1$. For L_I , there remains then only $l = 0$, as indicated at the left side of the figure. The $K\beta$ and $K\gamma$ "lines," also, are actually doublets like $K\alpha$, which can be resolved in heavy atoms; they fix $l = 1$, likewise, for M_{II} , M_{III} , N_{II} , N_{III} . The remaining M levels, M_I , M_{IV} , and M_V , must then consist, in some order, of $(0, \frac{1}{2})$, $(2, \frac{3}{2})$ and $(2, \frac{5}{2})$ (i.e., of levels with $l = 0$ and $j = \frac{1}{2}$, etc.). These latter three levels could all combine with L_{II} , L_{III} , i.e., with $(1, \frac{1}{2})$ and $(1, \frac{3}{2})$, without violating the rule that $\Delta l = \pm 1$. The rule that $\Delta j = 0 \text{ or } \pm 1$, however, limits $(2, \frac{5}{2})$ to combining with $(1, \frac{3}{2})$ only. It is found experimentally that M_V combines only with L , giving rise to the brightest of the L lines, called La_1 , whereas M_I and M_{IV} combine with both L_{II} and L_{III} (lines L_n , LI , $L\beta$, La_2 , the first two very weak). Hence, it must be that $M_V = (2, \frac{5}{2})$, and also that $L_n = (1, \frac{1}{2})$, so that, between L_{II} and M_V , $\Delta j = 2$ and the transition

is forbidden. Thus we have identified L_I and L_{II} ; L_{III} is then also identified, by a process of elimination, as $(1, \frac{3}{2})$.

If enough lines of the M group were known, we could continue in this manner and identify all of the M levels. In default of such knowledge, we may fall back on the conclusion drawn from wave-mechanical perturbation theory that the levels should lie in the same order in all shells, and with levels of larger j at lower energies. This principle agrees with all that can be learned from the observed lines; and it suffices to complete the arrangement of the levels, as shown in Fig. 156. Wave mechanics is presumably capable of predicting the spacing of all levels in full detail, but it has not yet been possible to carry out the necessary mathematical calculations.

The values of n , I , and j thus associated with the various X-ray levels are shown at the left in Fig. 156. It may be remarked that, instead of labels in terms of n , I , and j , the corresponding spectroscopic notation for atomic levels is also sometimes employed, the correspondence between this notation and the older X-ray symbols being as follows (Sec. 131; all the terms here are doublet terms):

K $1S_{\frac{1}{2}}$	L_I $2S_{\frac{1}{2}}$	$L_{II,III}$ $2P_{\frac{1}{2},\frac{3}{2}}$	M_I $3S_{\frac{1}{2}}$	$M_{II,III}$ $3P_{\frac{1}{2},\frac{3}{2}}$	$M_{IV,V}$ $3D_{\frac{3}{2},\frac{5}{2}}$
N_I $4S_{\frac{1}{2}}$	$N_{II,III}$ $4P_{\frac{1}{2},\frac{3}{2}}$	$N_{IV,V}$ $4D_{\frac{3}{2},\frac{5}{2}}$	$N_{VI,VII}$ $4F_{\frac{5}{2},\frac{7}{2}}$		

Certain predictions as to the *relative intensity* of X-ray lines may also be derived from the theory. Lines in which I and j change in the same sense should be stronger than those in which I and j change in opposite directions; and the lines for the largest values of I and j should be, as a rule, the strongest. These conclusions from the approximate theory agree, in a general way, with the data. The order of the subscripts commonly used in designating lines, $\alpha, \beta, \gamma, \delta, \dots$, indicates, in general, decreasing intensity; lines such as n and I are very weak. In the case of lines that start from a common initial level, furthermore, the theory leads to *quantitative* predictions concerning the relative intensities. When the initial level is the same for two lines, uncertainties in regard to the relative probabilities of excitation do not arise. Thus, theoretically, $K\alpha_1$ should be twice as strong as $K\alpha_2$; and experimentally this is at least very nearly true.

Experimental work on the relative intensities of X-ray lines is, however, usually very difficult, because of the numerous troublesome corrections that have to be made to the observed intensities.¹

¹ PARRATT, *Phys. Rev.*, vol. 54, p. 99 (1938); see also Sec. 182,

It should be remarked in conclusion that the dipole selection rules, employed above, do not hold in all cases, in agreement again with theoretical predictions. Many weak lines "forbidden" by these rules have been observed; especially in the spectra of the heavier elements, lines have been reported¹ corresponding to transitions $L_{III} \rightarrow N_{II}$ and $L_{III} \rightarrow N_{III}$, for each of which $\Delta l = 0$. Such lines are called "quadrupole" lines. They are subject to a different set of selection rules (e.g., $\Delta l = 0$ or ± 2).

181- The Continuous X-ray Spectrum.—The characteristic line spectrum emitted by an X-ray target as observed is superposed upon a *continuous spectrum*. This appears clearly, for example, in Fig. 151.

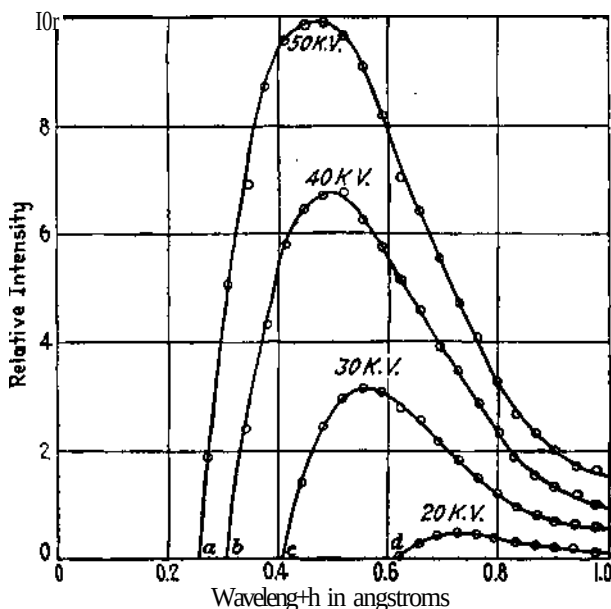


FIG. 157.—Ulrey's curves for the distribution of energy in the continuous X-ray spectrum of tungsten at various voltages.

The positions, i.e., wave lengths, of the lines are determined solely by the material of the target; their *intensity* is determined, for a given target material and tube current, by the voltage applied to the tube. On the other hand, the *wave-length* characteristics of the *continuous spectrum* are quite independent of the *material* of the target but are determined by the voltage applied to the tube. The *intensity* of the continuous spectrum, for a given tube current, is dependent both on the target material and on the applied voltage, as well as on the thick-

¹ IDEI, *Tdoku Univ., Set. & Technol. Reports*, vol. 19, p. 559 (1930); KAUFMAN, *Phys. Rev.*, vol. 40, p. 116 (1932).

ness of the target. In this section, we shall give a brief survey of some of the more important facts concerning the continuous spectrum.

A series of four spectral-energy distribution curves, recorded with an ionization spectrometer for the radiation from a tungsten target, and for applied voltages of 20,000, 30,000, 40,000, and 50,000 volts, respectively, is shown in Fig. 157.¹ The spectral region of this figure lies between the *K* and *L* series lines of tungsten. Starting at the long-wave-length side, the curves rise to a maximum and then drop rapidly toward zero; the position of the maximum depends on the applied

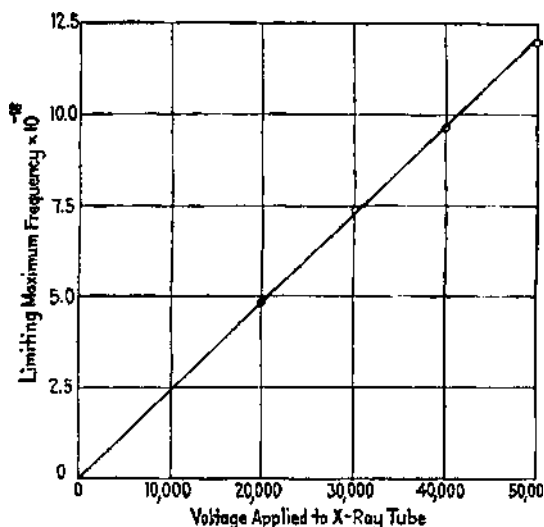


FIG. 158.—The relation between the limiting frequency and the applied voltage.

voltage. The curves are seen to meet the axis at finite angles, as is shown at the intersections *a*, *b*, *c*, and *d*, respectively. These intersections can be very accurately measured by making readings nearer the axis² than the readings shown in Fig. 157. It is seen that the intersections come at shorter wave lengths, the higher the voltage.

A very simple relation exists between these limiting frequencies and the applied voltage. This relation is shown graphically in Fig. 158, in which limiting frequencies are plotted as ordinates against applied voltages as abscissas. The graph is a straight line passing through the origin; the limiting frequency is strictly proportional to the applied voltage, the empirical equation of the curve being

$$v_0 = 2.43 \times 10^{14} V, \quad (285)$$

¹ These curves are from the measurements of Ulrey, *Phys. Rev.*, vol. 11, p. 401 (1918). They are subject to certain corrections to be mentioned later.

² See DUANE and HUNT, *Phys. Rev.*, vol. 6, p. 166 (1915).

where ν_0 is the *maximum* frequency which an applied voltage V (in volts) can generate.

The existence of this sharp "cutoff," at a limiting minimum wave length, is hard to explain on the basis of classical theory. The continuous spectrum is believed to be emitted as a consequence of the deflection of the cathode electrons by the strong fields surrounding the nuclei of the atoms of the target, in contrast to the line spectrum, which is emitted by electrons of the target itself in atoms which have been ionized. According to classical theory, the acceleration of an electron as it undergoes deflection in the field of a nucleus will cause the emission of a pulse of radiation. The energy-distribution curve for such a pulse can be found by making a Fourier analysis of the pulse into monochromatic wave trains. Now a form of pulse can be invented which produces a sharp cutoff of the distribution curve;¹ but no good reason can be assigned to explain why just this form of pulse should always be emitted in consequence of the deflection or stoppage of a cathode particle. Quantum theory, on the other hand, easily explains the existence of the sharp cutoff.

According to quantum theory, our conception of the process by which the negative acceleration of cathode electrons produces radiation requires the same sort of modification that has been found to be necessary in the case of a spectral line emitted by an atomic electron. According to *classical theory*, a spectral line represents radiation that is emitted continuously by a vibrating electron; its frequency is equal to the frequency of vibration of the electron, whose energy decreases continuously. According to *quantum theory*, on the other hand, the radiation composing a spectral line is emitted in the form of quanta or photons, each containing energy $h\nu$, and the frequency ν of the radiation is determined by the Einstein equation

$$h\nu = \Delta W$$

where ΔW is the total loss of energy by the electron. Similarly, a cathode electron passing through the strong electric field near the nucleus of an atom in an X-ray tube is not to be thought of as undergoing continuous acceleration, accompanied by the continuous emission of radiation, as is required by classical theory. The process is essentially one of alternative probabilities. The electron may actually pass through the field undeviated in direction; it is much more likely, however, to issue from the atom in a different direction from its direction of approach. It may suffer no change in its energy, being merely deflected or "scattered"; and in this case no radiation is emitted.

¹ KENNARD, *Phys. Zeits.*, vol. 24, p. 372 (1923).

Or, it may suffer a decrease ΔW in energy; in this latter case the energy lost is emitted as a photon of frequency ν given by Einstein's equation, $h\nu = \Delta W$. The amount of energy lost cannot exceed the total kinetic energy with which the electron enters the target, but it may have any value smaller than this.

It follows that the *maximum frequency* VQ of the X-rays will be the frequency of a photon emitted when the electron is brought entirely to rest as the result of a single elementary process. For the energy with which the electron approaches the atom, we can write eV , e being the (numerical) electronic charge and V the potential difference through which it falls in passing from the cathode of the tube to the target. The maximum frequency VQ present in the X-rays will thus be determined by the equation

$$h\nu_0 = eV. \quad (286)$$

In addition to this maximum frequency ν_0 , we should then expect also a whole spectrum of *lower* frequencies, emitted by electrons which lose only a part of their energy in a single encounter with a nucleus, in the form of a smaller quantum $h\nu$. A further reason for the emission

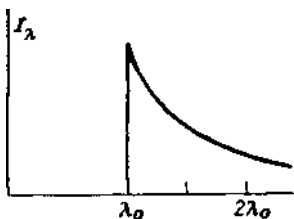


FIG. 159.—Approximate distribution of energy in the continuous X-ray spectrum from a very thin target, near the short-wave-length limit.

of frequencies lower than ν_0 lies in the fact that, in the vast majority of cases, an electron will experience many collisions with atoms of the target before being brought to rest, and at each of these collisions some of the initial energy eV is dissipated. Thus, a great deal of radiation is emitted by electrons which impinge upon atoms with incident energy less than eV . If the target is sufficiently thin, however—say of very thin gold foil—only a few of the electrons of the incident cathode stream will collide with

atoms in it, most of them passing through the target undeviated. Thus slowly moving electrons will not be present in a *thin* target to the same degree as in a *thick* one. Accordingly, we should expect that a greater "proportion of the energy in the continuous spectrum from thin targets should lie near the ν_0 limit than from thick targets. This is in agreement with experiment.

In the continuous spectrum from a *very thin* target, as a matter of fact, experiment indicates, in agreement with the wave-mechanical computation of Sommerfeld, that the maximum of the energy-distribution curve occurs *at the limiting wave length* VQ itself. On the short-wave side of λ_0 , the curve drops abruptly to the axis of abscissas,

whereas toward longer waves it falls nearly in proportion to $1/X^2$, as illustrated in Fig. 159. The curves for a thick target, as in Fig. 157, can be regarded as arising from the superposition of many elementary curves, such as that in Fig. 159, with various values of X_0 .¹ In such a target it might be expected that there would be a *most probable* type of collision which would correspond to the peak or maximum of the energy-distribution curve.

With the *shape* of the energy-distribution curve we are not, for the moment, concerned. Rather, let us return to the empirical equation (285) and to the curve (Fig. 158) which showed, in complete agreement with Eq. (286), that $V < ax \ll v_0$. Writing Eq. (286) in the form

$$v_0 = \frac{e}{h} V = \frac{e}{h} \frac{V'}{300},$$

where V is in electrostatic units and V' is in volts, we see that the numerical constant in Eq. (285) should be equal to the multiplier of V in this last equation. That is, we should have

$$\frac{v_0}{300h} = 2.43 \times 10^{14},$$

from which, if we use the old value, $e = 4.77 \times 10^{-10}$ e.s.u., we find

$$h = 6.55 \times 10^{-27},$$

At the present time the value of h as derived from precise measurements of the limiting X-ray frequency and a modern value of e of about 4.803×10^{-10} , is² in good agreement with values obtained in other ways. In this book the value $h = 6.61 \times 10^{-27}$ is retained, but perhaps 6.62 is better.³

182. Intensity of the Continuous Spectrum.—Determinations of the total intensity, and of the distribution of intensity in the continuous X-ray spectrum are by no means so satisfactory as the measurements of minimum wave lengths. This is due to the fact that numerous troublesome corrections have to be made to the *observed* energy-distribution curves, such as those shown in Fig. 157, before the true

¹ Cf. COMPTON and ALLISON, *op. cit.*, p. 90.

² Cf. Dunnington, *Rev. Modern Phys.*, vol. 11, p. 65 (1939), and for the best experimental work, DuMond and Bollman, *Phys. Rev.*, vol. 51, p. 400 (1937).

³ Cf. BIBGE, *Rev. Modern Phys.*, vol. 13, p. 233 (1941).

curve can be obtained.¹ Among these necessary corrections may be mentioned the following:

1. *The variation with wave length of the absorption of the radiation in its path from the atom of the target in which the radiation originates to the ionization chamber* where the intensity of the radiation is measured. This path includes the material of the target itself, since the X-rays originate at finite (though very small) depths below its surface, the walls of the X-ray tube, the window of the ionization chamber, and the air between the tube and the chamber.

2. *The reflecting power of the crystal grating*, which varies with wave length.

3. *Incomplete absorption in the ionization chamber.* Either an ionization chamber of sufficient length to absorb the entire X-ray beam must be used, or else corrections must be made for variation with wave length of the absorption within the chamber, since the *observed* ionization current is proportional to the energy *absorbed* in the chamber.

4. *"Second-order reflection."* According to Eq. (282) which may take the form

$$\lambda = \frac{2d \sin \theta}{n},$$

when the crystal in the spectrometer [Fig. 150(a)] is set at a given angle θ , a series of wave lengths λ will be reflected for different values 1, 2, 3, . . . of n , *provided* those several wave lengths are present in the incident radiation. Suppose, for example, that when the voltage applied to the X-ray tube is 50 kilovolts, the crystal is set at such an angle θ_0 as to give a *first-order* reflection {i.e., $n = 1$) of $\lambda = 0.6$ angstrom. According to Fig. 157, the minimum wave length present in the X-ray spectrum generated by 50 kilovolts is 0.250 Å. Under these conditions, there will be reflected into the ionization chamber the *second-order* wave length $\lambda = 0.3$ Å, in addition to the first-order wave length $\lambda = 0.6$ Å. The *observed intensity* when the spectrometer is set at $\lambda = 0.6$ Å is, therefore, the sum of the intensities due to both wave lengths. For an applied voltage of 30 kilovolts, however, only first-order reflection is present at λ_0 , and no correction is necessary.

5. *Secondary production of X-rays.* If one is interested primarily in the radiations produced by impacts of the incident cathode rays alone, corrections must be made for the radiations arising from impacts of the various types of *secondary electrons* produced in the target itself, e.g., electrons ejected from the atoms of the target by the impact

¹ These corrections, however, in no wise affect the determination of minimum wave lengths.

of cathode rays, or by photoelectric ejection, or by ejection as Compton electrons in connection with the scattering of the X-rays in the target [Sec. 186(a)].

6. *Characteristic lines.* There generally exist several characteristic X-ray lines superimposed on the continuous spectrum; the intensities of such lines must be subtracted from the intensity of the beam as observed.

Without discussing the methods of making these several corrections,¹ the final results may be summarized as follows:

- a. The total radiation, *i.e.*, the area under the (corrected) energy-distribution curves, is, for a given target material and tube current, very nearly proportional to the square of the applied voltage.
- b. At a given voltage and tube current, the total radiation varies roughly as the first power of the atomic number of the target.
- c. The wave length X_m at which a given energy-distribution curve for a thick target attains its maximum is roughly 1.5 times the wave length at which the same curve meets the wave-length axis.
- d. Kulenkampff² gives the following empirical formula for the intensity I_ν of the X-ray energy at frequency ν from a target of atomic number Z when bombarded by electrons which have fallen through a potential V :

$$I_\nu = CZ(\nu_0 - \nu) + BZ^2,$$

where ν_0 is the maximum frequency determined by the quantum relation $eV = h\nu_0$, and C and B are constants independent of voltage and target material. The formula holds only for $\nu > \nu_0$: at $\nu = \nu_0$ the intensity is assumed to sink discontinuously to zero.

The *efficiency* of an X-ray tube (with thick target) is very low. If we define the efficiency as the fraction of the energy carried by the cathode electrons which is converted into X-ray energy, it varies from about 0.01 to 0.1 percent.

INTERACTIONS OF X-RAYS WITH ATOMS

183. The Absorption of X-rays.—In contrast with the apparently chaotic state of affairs in regard to the absorption of light in the visible or near-visible portions of the spectrum, we find comparative simplicity in the empirical laws for the absorption of X-rays. Measure-

¹ See KIBKPATRICK, *Phys. Rev.*, vol. 22, p. 414 (1923).

² *Ann. d. Physik*, vol. 69, p. 548 (1922).

ments of the absorption coefficient for a monochromatic beam of parallel rays are readily made by use of the ionization spectrometer [Fig. 150(a)]. For a given crystal angle θ and, therefore, wave length λ , the ionization current is measured both with and without a sheet of absorbing material of known density p and thickness x placed in the path of the beam, say between the two slits S_1 and S_2 . These measurements

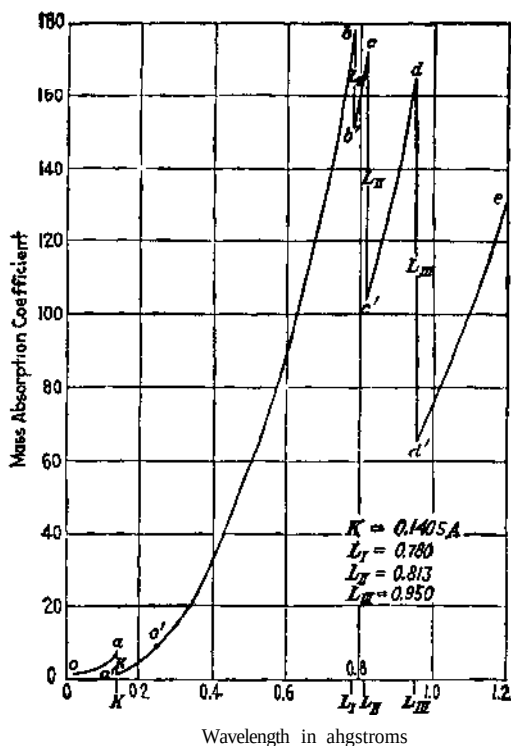


FIG. 160.—K and L absorption limits of lead.

give, respectively, I and J_0 in the equation [cf. Eq. (280) in Sec. 173]

$$I = I_0 e^{-\mu x} = I_0 e^{-\frac{\mu}{\rho} \rho x}, \quad (287)$$

from which the *linear absorption coefficient* μ or the *mass absorption coefficient* μ/ρ may be computed. In this way, being careful to use such a tube voltage as to eliminate second-order reflections, one may obtain values of μ/ρ for various wave lengths and various substances. The mass absorption coefficient μ/ρ is much more commonly used than the linear coefficient.

(a) *Properties of μ/ρ* —Figure 160 shows roughly the mass absorption coefficient of lead in the wave-length range $0.1 < X < 1.2 \text{ \AA}$.

Beginning at point 0, u/P rises rapidly with increasing wave length until point a, corresponding to $X = 0.1405 \text{ \AA}$ and $u/p = 8$ (about) is reached, at which the value of u/p suddenly drops to point a' . This is the K absorption limit (Sees. 173 and 178). Up to this point, absorption accompanied by ejection of K electrons has constituted the major part of the absorption. At longer wave lengths, this K absorption does not occur at all, hence the drop in u/p .

With further increase in wave length, however, the absorption again increases rapidly, being mostly due now to ionization of atoms in the L shell, until at point b, corresponding to $X = 0.780 \text{ \AA}$, the L_I absorption limit is reached, at which there occurs another drop in the absorption to point b' . From V to c, absorption still occurs owing

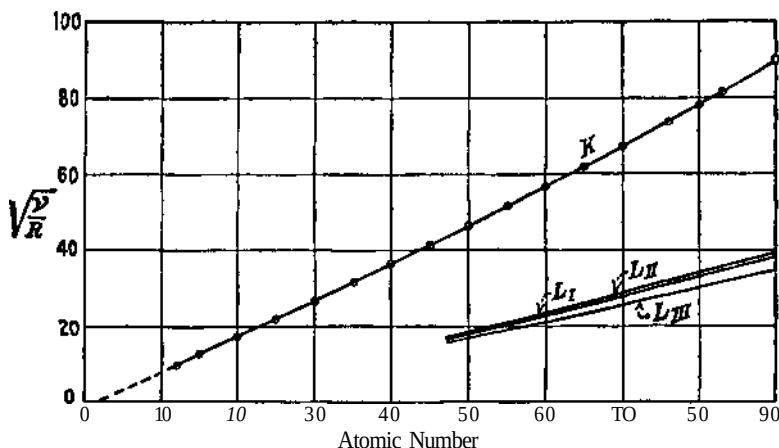


FIG. 161.—The Moseley diagram for the K and L absorption limits.

to processes which leave the ionized atoms in L_{III} or L_{low} states; i.e., the atoms are raised to their L_{II} or L_{III} energy levels. None, however, is raised to the L_I level. Similar drops or "breaks" occur at cc' ($X = 0.813 \text{ \AA}$) and at dd' ($X = 0.950 \text{ \AA}$), the L_{II} and L_{III} limits, respectively; from c' to d , absorption into the L_{III} state is still occurring, but beyond d' absorption with ionization in the L shell ceases altogether. Beyond point d' , however, the absorption again increases rapidly. If we could follow, by direct measurement, the absorption beyond point e, we should find that, in the region $3.2 < X < 5.0 \text{ \AA}$, a group of five "breaks" occurs, representing the five M absorption limits; beginning at about $14 \text{ \AA}$, there would come the group of seven N limits; etc.

Curves similar to Fig. 160 are obtained for the absorption of X-rays in other substances, the respective discontinuities or limits occurring

at longer wave lengths, the lower the atomic number of the absorber. In fact, a curve plotted between the square root of the frequency of a given limit and the atomic number is nearly a straight line, similar to Moseley's curve (Fig. 152) for line spectra. Such a curve for the K limit, from Mg ($Z = 12$) to Th ($Z = 90$) is shown in¹ Fig. 161, in which, as ordinates, are plotted $\sqrt{\nu/R}$, where R is the Rydberg constant for an atom of infinite weight. The graph is seen to be nearly, but not quite, a straight line. For comparison, there are also shown the corresponding graphs for the L_I , L_{II} , and L_{III} limits from Ag ($Z = 47$) to Th ($Z = 90$).

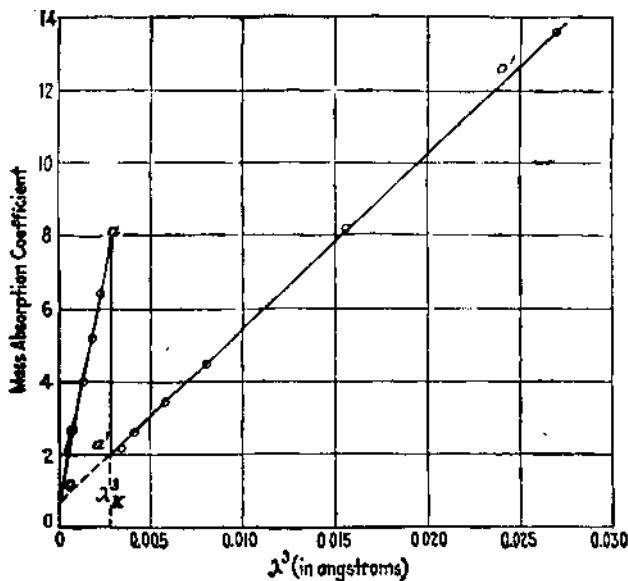


FIG. 162.-Mass absorption coefficients of lead as a function of wave length, showing the K discontinuity and the linear relation between μ/p and λ^3 .

Not only is there regularity in the wave lengths of the absorption limits from one element to another, but there is also a remarkable regularity in the *magnitude* of the mass absorption coefficient, from wave length to wave length in a given element, and from one element to another. Figure 162 shows, on a much larger scale, the mass absorption coefficients of lead in² the region o to o' of Fig. 160, plotted as a function of the *cube* of the wave length. The two parts of the graph on each side of the K absorption limit aa' are seen to be *straight lines*, with very nearly equal intercepts of about $\mu/p = 1$. We may,

¹ Data from SIBGBAHN, *loc. cit.*

² RICHTMYEB, *Phys. Rev.*, vol. 27, p. 1 (1925).

accordingly, write the empirical equation for the variation of μ/P with X in the regions oa and $a'o'$, respectively, as

$$\frac{\mu}{\rho} = k\lambda^3 + b, \quad \frac{\mu}{\rho} = k'\lambda^3 + b, \quad (288a,b)$$

where k and k' are the slopes of the respective lines and b is a constant. Thus, to a first approximation,¹ the mass absorption coefficient for a given absorber is a linear function of the cube of the wave length, due regard being given to the spectral region in which the equations apply.

A quantity of more fundamental significance than μ or μ/ρ is the *atomic absorption coefficient* μ_a . This quantity represents the absorption due to a single atom and is obtained by dividing the mass absorption coefficient μ/ρ by the number of atoms per gram, N_0/A , where N_0 is Avogadro's weight:

Equations (288a,b) may then be written

$$\mu_a = k_a\lambda^3 + b_a, \quad \mu_a = k'_a\lambda^3 + b_a. \quad (289a,b)$$

(More commonly the symbol r_a is written for b_a , for the reason stated below.) In terms of μ_a a second type of regularity has been found to exist in the absorption coefficient. The constants k_a and k'_a vary approximately as the fourth power of the atomic number.² Accordingly, Eq. (289a,b) may be written in the following form, λ_K and λ_L denoting, respectively, the K and L absorption limits:

$$\lambda < \lambda_K: \quad \mu_a = C_a Z^4 \lambda^3 + b_a \quad (290)$$

$$\lambda_K < \lambda < \lambda_L: \quad \mu_a = C'_a Z^4 \lambda^3 + b'_a. \quad (291)$$

Here, when X is in centimeters, $C_a = 2.25 \times 10^{-2}$, $C'_a = 0.33 \times 10^{-2}$.

It seems surprising that such simple relations should hold, even approximately, for the absorption coefficients. The theoretical treatment of this topic is not in a very satisfactory state.³

(b) *Physical Interpretation of the Absorption Formulas.*—As we have seen, the so-called "absorption" of X-rays is due partly to the photoelectric effect, to be described in the next section, and partly to

¹ There are slight systematic departures from the X^3 law. See COMPTON and ALLISON, *op. cit.*; CUYKENDALL, *Phys. Rev.*, vol. 50, p. 105 (1936); JONES, *Phys. Rev.*, vol. 50, p. 110 (1936); ANDREWS, *Phys. Rev.*, vol. 54, p. 944 (1938).

² BRAGG and PIERCE, *Phil. Mag.*, vol. 28, p. 626 (1914); RICHTMYER and WARBURTON, *Phys. Rev.*, vol. 22, p. 539 (1923).

³ See COMPTON and ALLISON, *op. cit.*

scattering (Sees. 172 and 173). Hence, we may conveniently write

$$\mu = T + a = T_a + (T_s + a_s)$$

where T or T_a represents the part of the absorption that is due to the photoelectric effect and a or a_s the part that is due to scattering. According to classical theory the scattering should be independent of frequency [Sec. 172(a)], and for this reason a or a_s has often been identified with the constant term b or b_s in the empirical formulas (289a,b) or (290, 291). The term in λ^3 is then identified with the photoelectric absorption. We shall see, however, that in reality the scattering coefficient varies considerably with wave length, so that such an interpretation of the empirical formulas can possess only approximate validity.

184, The Photoelectric Effect for X-rays.—If part of the absorption of X-rays is due, as we have assumed, to a photoelectric effect, it should be possible to observe the photoelectrons thus produced, and Einstein's photoelectric equation should be applicable to them. In the ordinary photoelectric effect, the electrons come from among the "free" electrons in the metal (Sec. 54), and their maximum kinetic energy, when produced by a beam of light of frequency ν , is $\frac{1}{2}mv_m^2$, as given by Eq. (48) in Sec. 46:

$$\frac{1}{2}mv_m^2 = h\nu - \omega_0,$$

where ω_0 is the work required to remove an electron from just under the surface of the metal ($h =$ Planck's constant). When photoelectrons are produced by X-rays, on the contrary, they will come (at least principally) from the deeper-lying shells of the atoms. If a photoelectron is ejected from the K shell by X-rays of frequency ν , the amount of energy absorbed from the radiation is $h\nu$; but work equal to $h\nu_K$ must be done by the electron in escaping from the atom, ν_K being the frequency of the K absorption limit. Hence the electron will emerge from the atom with kinetic energy equal to $h\nu - h\nu_K$. If the atom lies on the surface of the absorbing material, the electron may escape into the surrounding space with this amount of energy (more exactly with energy $h\nu - h\nu_K - \omega_0$, but ω_0 is usually negligibly small). Otherwise it may lose part of its energy in passing through a layer of matter. Thus the *maximum* kinetic energy with which photoelectrons produced by absorption in the K shell may emerge from the absorbing material will be

$$\frac{1}{2}mv_m^2 = h\nu - h\nu_K = h(\nu - \nu_K) \quad (292)$$

(provided w_0 may be neglected). If $\nu < \nu_K$, no photoelectrons at all can be ejected from the K shell. As ν increases above ν_K , the maximum energy of the photoelectrons increases linearly, just as in the ordinary photoelectric effect.

The number of photoelectrons ejected from the K shells of atoms of a given absorber decreases as ν increases, thereby causing the known decrease in the absorption coefficient as ν recedes from ν_K . The number of photoelectrons is strictly proportional to the intensity of the X-rays.

The photoelectrons ejected from the L shell consist of three slightly different groups, with maximum energies corresponding to the three L absorption limits, ν_{Lb} , ν_{LII} , ν_{LIII} , the maximum kinetic energies in the three groups are

$$\frac{1}{2} m v_{mI}^2 = h(\nu - \nu_{LI}), \quad \frac{1}{2} m v_{mII}^2 = h(\nu - \nu_{LII}), \\ \frac{1}{2} m v_{mIII}^2 = h(\nu - \nu_{LIII}).$$

The differences in these maximum energies arise from differences in the quantum state in which the remaining L electrons are left

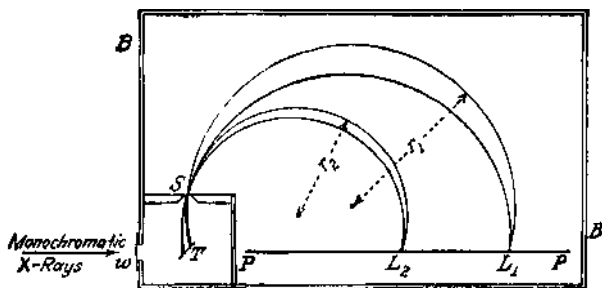


FIG. 163.—Robinson's magnetic spectrograph for studying the photoelectric action of X-rays.

upon the departure of the ejected one. Similarly, from the M shell there are five groups; and so on. If we start with X-rays of very high frequency, photoelectrons of all kinds are produced. If the frequency is then decreased, as it passes ν_K , the K photoelectrons disappear; as it passes the L limits, the three groups of L photoelectrons disappear in turn; and so on.

Among the experiments demonstrating these facts may be mentioned those of Robinson and his collaborators,¹ whose apparatus is

¹ ROBINSON and RAWLINSON, *Phil. Mag.*, vol. 28, p. 277 (1914); ROBINSON, *Roy. Soc. Proc.*, vol. 104, p. 455 (1923); *Phil. Mag.*, vol. 50, p. 241 (1925); ROBINSON and CASSIE, *Roy. Soc. Proc.*, vol. 113, p. 282 (1927); ROBINSON and YOUNG, *Roy. Soc. Proc.*, vol. 128, p. 92 (1930).

shown diagrammatically in Fig. 163. A beam of X-rays of frequency¹ ν enters through a thin window W a highly evacuated brass box BB and falls upon a target T of the material under investigation. Photoelectrons are expelled from the surface of T in all directions and with various velocities, as indicated by Eqs. (292), (293a,b,c). The whole apparatus is placed in a known magnetic field H , at right angles to the plane of the paper, which can be varied at will. The photoelectrons will describe circles in the field. Some of them will pass through the narrow slit S and eventually strike the photographic plate PP .

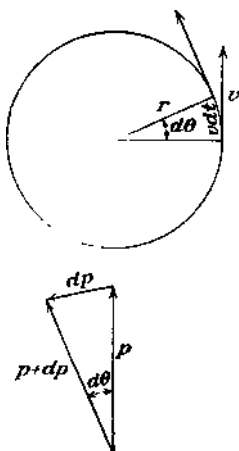


FIG. 164.

If the electrons leaving T have velocities $V_1, V_2, \dots$, they will move in circles of radii $r_1, r_2, \dots$ and will strike the plate at points $L_1, L_2, \dots$. As shown in the figure, the arrangement is such as to "focus" electrons leaving the different parts of the target with the same velocity onto the plate at such positions as L_1 and L_2 , the diameter of the circle being the distance between S and L_1 or L_2 .

In an experiment such as this, the velocity of the electrons is so great that variation of mass with velocity must be taken into account. This is easily done as follows. As the electron goes around in a circle under the influence of the magnetic field, its momentum p is constant in magnitude but continually changes in direction. We can imagine that the electron receives continual vector increments of momentum directed toward the center of the circle. As the direction of motion changes by $d\theta$ an increment dp of momentum must be added of magnitude (Fig. 164)

$$dp = p d\theta.$$

The force on the electron is $e\nu H$, e being the electronic charge in electromagnetic units and ν its velocity. Hence during a time dt the electron receives momentum

$$dp = e\nu H dt.$$

¹ It is extremely difficult by present experimental means to get a strictly monochromatic beam of X-rays of sufficient intensity for such an experiment as this. Accordingly, use is made of the fact that, with suitable exciting voltage, the $K\alpha$ lines from an X-ray target are much more intense than the accompanying radiation of the other wave lengths and, therefore, serve effectively as a "monochromatic" beam, particularly if the beam is first passed through a filter of suitable thickness, the wave length of whose K limit is just shorter than the wave length of the $K\alpha$ lines. [See HULL, *Phys. Rev.*, vol. 10, p. 661 (1917).]

Thus

$$dp = p d\theta = evH dt.$$

But (Fig. 164)

$$\begin{aligned} d\theta &= \frac{v dt}{r} \\ \therefore p &= erH. \end{aligned} \quad (294)$$

Now according to the relativistic formula [Eq. (60) in Sec. 63]

$$p = \frac{mv}{(1 - v^2/c^2)^{1/2}}, \quad (295)$$

where m stands for the rest mass of the electron (0.911×10^{-27} gram); and Eq. (61) in Sec. 64 gives for its kinetic energy

$$T = mc^2 \left[\frac{1}{(1 - v^2/c^2)^{1/2}} - 1 \right].$$

From (295)

$$\frac{p^2}{m^2 c^2} = \frac{v^2/c^2}{1 - v^2/c^2}, \quad \therefore \frac{p^2}{m^2 c^2} + 1 = \frac{1}{1 - v^2/c^2}.$$

Hence we can also write¹

$$T = mc^2 \left[\left(1 + \frac{p^2}{m^2 c^2} \right)^{1/2} - 1 \right]; \quad (296)$$

and for an electron revolving in a magnetic field, by (294),

$$T = mc^2 \left[\left(1 + \frac{e^2 r^2 H^2}{m^2 c^2} \right)^{1/2} - 1 \right]. \quad (297)$$

(Here e is in electromagnetic units.)

By means of this last formula, values of the kinetic energy T of the photoelectrons can be calculated from measured values of r , for comparison with $h(\nu - \nu_A)$, where ν_A is any absorption limit. If losses of energy by the electrons within the absorber are negligible, we can substitute values of T so found in the photoelectric equation

$$T = h(\nu - \nu_A)$$

and, knowing ν , can obtain values of the absorption limit ν_A , for comparison with values measured spectroscopically.

Robinson found on his plates a number of "lines" representing groups of photoelectrons. The corresponding absorption limits, calculated in the manner just described, are given in Table V; values

¹ Calculations from such formulas are most easily made thus:

$$T = mc^2 [(\cos \tan^{-1} p/mc)^{-1} - 1].$$

determined spectroscopically are added for comparison. We note that Robinson's results are *in complete agreement with the data of X-ray spectroscopy*. It is noteworthy that he found *one K level, three L levels, five M levels*, and, in the case of U 92, five of the seven N

TABLE V.—SOME X-RAY ABSORPTION LIMITS DETERMINED FROM THE MAGNETIC SPECTRUM OF PHOTOELECTRONS (AFTER ROBINSON); COMPARISON WITH SPECTROSCOPIC METHOD

Level	Uranium Z = 92		Tungsten Z = 74		Silver Z = 47		Copper Z = 29	
	Robin-son	Spec.	Robin-son	Spec.	Robin-son	Spec.	Robin-son	Spec.
K							660.1	661.6
L _I			891.3	890.8	279.4	282.0	79.9	81.0
L _{II}			850.2	849.9	258.7	261.3	68.3	70.3
L _{III}			750.8	751.3	245.8	248.6		68.9
M _I	408.7	408.5	208.1	207.3	47.1	54.7	5.8	8.9
M _{II}	380.1	381.5	189.1	189.3		46.1		5.7
M _{III}	316.0	316.8	165.9	167.5	39.3	43.7		
M _{IV}	274.4	274.2	139.0	137.5	23.8	29.2		0.4
M _V	260.7	261.2	133.2	132.9		28.8		
N _I	106.4	106.0	33.0	43.3	4.3	8.7		
N _{II}	92.5	93.5		36.0		5.9		
N _{III}	75.0	76.6		31.0				
N _{IV,V}	53.2	55.9*	17.2	18.1*		2.0*		
N _{VI, VII}	26.0	28.0*						

(Data are given in terms of V_0/R where v_0 is the frequency of the limit and R is the Rydberg constant.) The columns headed "Robinson" give Robinson's observations. Those headed "Spec." give the values of the levels determined from X-ray spectroscopy, as given by Siegbahn.

* Indicates mean value.

levels, the pairs of levels N_{IV,V} and N_{VI,VII} being too close together to be resolved in his apparatus.

The technique of measuring the velocities of photoelectrons ejected by X-rays has been developed to the point that this method is now one of the most reliable for the precise determination of fundamental constants.¹

185. The Scattering of X-rays.—In spite of the remarkable early success of the classical theory in accounting for many features of the

¹ See, e.g., KRETSCHMAR, *Phys. Rev.*, vol. 43, p. 417 (1933).

scattering of X-rays, developments in this subject since 1920 illustrate strikingly the inadequacy of the classical theory and the superiority of quantum theory.

(a) *The Total Scattering Coefficient.*—When a beam of monochromatic X-rays of intensity I traverses a thickness dx of scattering material, part of the incident energy is scattered, for which we may write $(\kappa I) dx$. The coefficient κ thus defined is called the *total linear scattering coefficient* for X-rays of that wave length. According to the classical theory, on the two assumptions (1) that the electrons scatter as if they were free and (2) interference between the wavelets scattered by different electrons may be ignored, the value of κ can be found by multiplying the scattering coefficient σ_e for a single electron, as given in Eq. (277) in Sec. 172, by n , the number of electrons per cubic centimeter of the material. If we also divide by ρ , the density of the material, we find thus for the *classical mass scattering coefficient*

$$\frac{\sigma_0}{\rho} = \frac{n\sigma_e}{\rho} = \frac{8\pi}{3} \cdot \frac{e^4}{m^2 c^4} \cdot \frac{n}{\rho}.$$

Thus, if we assume that $\sigma = \sigma_0$, σ/ρ should be independent of wave length, and it should not vary greatly with the material, since n is roughly proportional to ρ (the number of electrons per atom being about half of the atomic weight); it should, in fact, be nearly equal to 0.2, as may be verified by putting into Eq. (298) the numerical values of the constants e , m , and c and the values of n and ρ for some scattering material, such as carbon or aluminum. In Fig. 1G5 are shown data on scattering for a number of elementary substances; the ordinates represent the ratio of the scattering coefficient κ for the substance, denoted by κ_s in the figure, to the value κ_0 as given in Eq. (298). We note from the figure that for light scattering materials and for X-rays of $\lambda > 0.2$ angstrom the scattering coefficients approximate to the classical, free-electron value, κ_0 . Otherwise, there are wide departures.

For heavier materials, and especially at longer wave lengths, or greatly exceeds κ_0 . This excessive scattering is easily explained classically. When the wave length becomes comparable with the distances between the electrons in the atoms, they no longer scatter independently; the waves scattered by different electrons become superposed on each other more or less in the same phase, and constructive interference occurs. If the wave length is actually long as compared to the distances between the N electrons in a given atom, the relative phase differences of the waves are all small, consequently the amplitude of the resultant scattered wave is approximately

proportional to N , and the intensity of the scattered radiation is proportional to N^2 . For a heavy atom, the difference is enormous.

The decrease of a at short wave lengths which is evident from Fig. 165, on the other hand, is impossible of explanation by classical theory. For $X = 0.1$ Å, the observed scattering coefficient is only about 0.15, instead of the value of about 0.2 deduced above. For r-rays, the wave lengths of which are of the order of .002 Å or less, the total absorption coefficient of carbon, including photoelectric

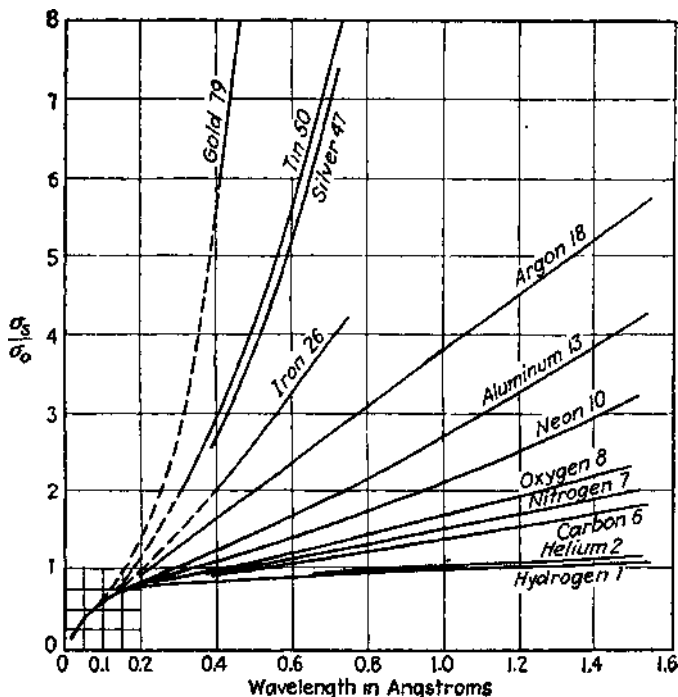


FIG. 165.—Scattering coefficients r_a of various elements for different wave lengths relative to the classical value r_0 . (Reproduced by courtesy of D. Van Nostrand Company, Inc., from Fig. III-3 in A. H. Compton and S. K. Allison, "X-rays in Theory and Experiment," 1935.)

absorption as well as that due to scattering, is only about 0.06. For such short waves, however, the assumption that the electrons scatter independently, and approximately at the same rate as if they were free, should certainly be justified; for, on classical theory, they will not move far enough from their equilibrium positions to call appreciable elastic forces into play. At this point classical theory fails badly. In the next section, a result will be quoted from wave mechanics which agrees very well with observations in this region of short wave lengths.

(b) *The Angular Distribution of Scattered X-ray Energy.*—The classical pulse theory of X-rays led to Eq. (279) in Sec. 172,

$$I_{\phi}' = \frac{I}{r^2} \frac{e^4}{m^2 c^4} \frac{1 + \cos^2 \phi}{2}$$

for the intensity I' of X-rays scattered per electron at an angle ϕ from the direction of the incident beam, the intensity of which is J ; r is the distance between the scatterer and the place where the scattered energy is measured. According to this equation, the scattering should be symmetrical forward and backward about a plane passing through the scatterer and perpendicular to the direction of the incident rays.

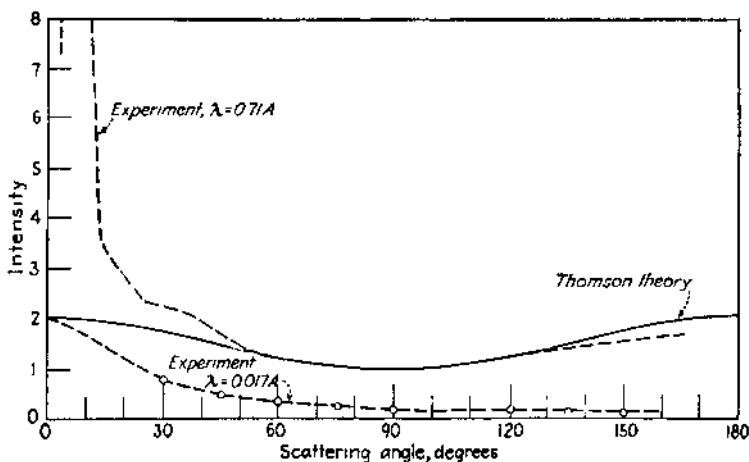


FIG. 166.—Relative scattered intensity at various angles of scattering, for X-rays of λ 0.71 Å scattered by mesitylene, and for X-rays of λ 0.017 Å scattered by iron. (Reproduced by courtesy of D. Van Nostrand Company, Inc., from Fig. III-4 in A. H. Compton and S. K. Allison, "X-rays in Theory and Experiment," 1935.)

If $\phi = 0$ or $\phi = \pi$, $1 + \cos^2 \phi = 2$; if $\phi = \pi/2$, $1 + \cos^2 \phi = 1$. Thus the energy scattered in the forward or backward directions should be twice that scattered at 90° from the incident beam.

In Fig. 166 are shown data on the variation with scattering angle ϕ of the intensity of X-rays of wave length $\lambda = 0.71$ Å scattered from liquid mesitylene $C_6H_3(CH_3)_3$. For comparison, the variation predicted by the classical theory of Thomson, as represented by the above formula, is shown by a solid line in the figure. We note that the intensity of scattering agrees well with the classical formula at scattering angles greater than 45° but shows a marked departure at smaller angles. Such a departure might perhaps be explained classically in the same way in which the excess scattering at long wave lengths was explained above. But in Fig. 166 data are also shown

on the scattering of γ -rays of wave length 0.017 Å from iron. These data show much smaller scattering than is required by the classical formula (except, presumably, at a zero scattering angle), and the scattered radiation exhibits great dissymmetry, being confined almost entirely to forward directions.

Such dissymmetry in the angular distribution presents difficulties for any classical theory, but it is fully accounted for by the wave-mechanical theory.¹

186. The Compton Scattering of X-rays, (*a*) *The Compton Effect.* It was early found that secondary radiation is usually less penetrating

than the primary radiation. This fact was later explained by the presence in the secondary beam, not only of scattered radiation, but also of the *characteristic fluorescent* radiation, which, as we have seen above [Sec. 178 (c)], is always of longer wave length and, therefore, is less penetrating than the exciting primary beam. It was found, however, that this explanation was not adequate in all cases.

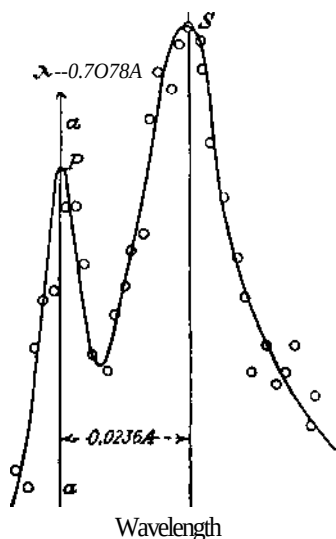


FIG. 167.—The spectrum of scattered X-rays, showing the unmodified line *P* and the modified or "shifted" line *S*. (Compton.)

A scattering material of very low atomic number, such as carbon, emits fluorescent radiation of such long wave length as to be absorbed by even a thin film of air. Yet, even with carbon immersed in air as the scatterer, the secondary beam was found to

be somewhat less penetrating and presumably therefore of somewhat longer mean wave length than the primary.

The explanation of such cases was discovered by A. H. Compton. He showed² that, when *monochromatic* primary radiation is used, the scattered beam is composed of *two lines*, one corresponding in wave length to that of the primary beam and the other being of definitely *longer* wave length. Figure 167 shows the spectrum of scattered radiation from a carbon scatterer when irradiated by the *K α* line of molybdenum, the scattered radiation being observed at an angle of

¹ Cf. HEITLER, W., "The Quantum Theory of Radiation," p. 154, 1936.

² COMPTON, A. H., *Phys. Rev.*, vol 21, p. 715; vol, 22, p. 409 (1923).

90° from the incident radiation. The vertical line aa indicates the wave length of the primary radiation. There are seen to be two maxima, or lines, one of which, P, corresponds *exactly* to the primary radiation; the other, S is "shifted" toward longer wave lengths from the position of the primary, the difference in wave length between the *shifted* and the *unshifted* components being 0.0236 angstrom.

In order to explain the occurrence of the shifted component, Compton boldly applied the extreme quantum picture of radiant energy, according to which a beam of radiation consists of a stream of localized quanta of energy, or photons. He assumed that the scattering process could be treated as an elastic collision between a photon and an electron and that this collision is governed by the two laws of mechanics: (1) the conversion of energy and (2) the conservation of momentum.

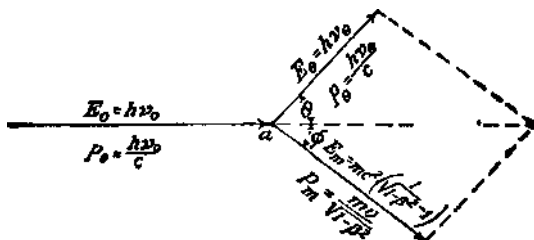


FIG. 168.—Vector diagram showing the conservation of momentum when a photon is scattered by an electron.

Let an incident photon containing energy $E_0 = h\nu_0$ rebound from an electron of rest mass m , initially at rest at the point a (Fig. 168). After the collision, the electron will move in a direction making an angle of ϕ , say, with the initial direction of motion of the photon, while the photon itself will move in a direction making an angle θ , say, with its initial direction. Qualitatively, one sees, at once, that the energy E_θ of the photon after collision and, therefore, also the frequency must be less than that of the photon before collision, since some of the energy of the incident photon must have been given to the electron.

If E_m is the kinetic energy given to the electron as the result of the collision, we have, from the law of the conservation of energy,

$$E_0 = E_\theta + E_m.$$

Using the relativistic formula, Eq. (61) in Sec. 64, we have for the kinetic energy of the electron

$$E_m = mc^2 \left(\frac{1}{\sqrt{1 - \beta^2}} - 1 \right), \quad \beta = \frac{v}{c} \quad (299)$$

where m is its rest mass, v is its velocity after the collision, and c is the speed of light. Inserting this value of E_m in the preceding equation, and also $E_a = h\nu_a$, $E_s = h\nu_s$, we have

$$h\nu_0 = h\nu_s + mc^2 \left(\frac{1}{\sqrt{1 - \beta^2}} - 1 \right). \quad (300)$$

Furthermore, as stated in Sec. 33, a beam of radiation in free space carries momentum equal to the energy carried by it divided by c . Hence a photon, containing energy $h\nu$, carries momentum $h\nu/c$. The momentum of an electron moving with speed v is given by Eq. (60) in Sec. 63. Thus for the momenta we have:

$\frac{h\nu_0}{c}$ = the momentum of the photon before the collision,

$\frac{h\nu_s}{c}$ = the momentum of the photon after the collision,

$mv/\sqrt{1 - \beta^2}$ = the momentum of the electron (after the collision).

From the law of the conservation of momentum, therefore, we can write for the x-component of momentum (in the direction of propagation of the incident photon)

$$\frac{h\nu_0}{c} = \frac{h\nu_s}{c} \cos \theta + \frac{m\beta c}{\sqrt{1 - \beta^2}} \cos \phi \quad (301)$$

and for the y-component, in the plane containing the directions of motion of both the scattered photon and the electron,

$$0 = \frac{h\nu_s}{c} \sin \theta - \frac{m\beta c}{\sqrt{1 - \beta^2}} \sin \phi. \quad (302)$$

In the three independent equations (300), (301), and (302), we have the four unknown quantities v , θ , ϕ , and β , assuming ν_0 , the frequency of the incident photon, to be known. By simultaneous solution, we can obtain three of these unknowns in terms of the fourth, say θ , which is the direction with respect to the incident beam in which the photon is "scattered" as a result of the particular collision.

If we introduce the wave lengths of the 2 photons (whatever wave length may mean in terms of this picture!), $\lambda_0 = c/\nu_0$ and $\lambda_s = c/\nu_s$, we can write the last two equations thus:

$$\frac{A}{\lambda_0} - \frac{h}{\lambda_s} \cos \theta = \frac{m\beta c}{\sqrt{1 - \beta^2}} \cos \phi, \quad \frac{h}{\lambda_s} \sin \theta = \frac{m\beta c}{\sqrt{1 - \beta^2}} \sin \phi.$$

Squaring these two equations and adding them, we have

$$\frac{h^2}{\lambda_0^2} + \frac{h^2}{\lambda_s^2} - \frac{2h^2}{\lambda_0 \lambda_s} \cos \theta = \frac{m^2 \beta^2 c^2}{1 - \beta^2} = \frac{m^2 c^2}{1 - \beta^2} - m^2 c^2. \quad (303)$$

Similarly Eq. (300) can be written, after dividing it through by c ,

$$\frac{h}{\lambda_0} - \frac{h}{\lambda_\theta} + mc = \frac{mc}{\sqrt{1 - \beta^2}}.$$

$$\therefore \frac{h^2}{\lambda_0^2} + \frac{h^2}{\lambda_\theta^2} - \frac{2h^2}{\lambda_0\lambda_\theta} + 2mch\left(\frac{1}{\lambda_0} + \frac{1}{\lambda_\theta}\right) + m^2c^2 = \frac{m^2c^2}{1 - \beta^2}.$$

Subtracting Eq. (303) from the last equation, we obtain

$$\frac{2h^2}{\lambda_0\lambda_\theta} (\cos \theta - 1) + 2mch\left(\frac{1}{\lambda_0} - \frac{1}{\lambda_\theta}\right) = 0.$$

$$\therefore \lambda_\theta - \lambda_0 = \frac{h}{mc}(1 - \cos \theta). \quad (304)$$

If we insert here the values $h = 6.61 \times 10^{-27}$, $m = 0.9107 \times 10^{-27}$, $c = 2.998 \times 10^{10}$, and multiply by 10^8 , we have

$$\lambda_\theta - \lambda_0 = 0.0242(1 - \cos \theta) \text{ angstroms.} \quad (305)$$

This important equation states that when incident radiation of wave length λ_0 is scattered by a free electron at an angle θ , the wave length λ_θ of the scattered radiation should be greater than that of the incident radiation by the quantity $0.0242(1 - \cos \theta)$ angstroms, which, for a given angle θ , is constant whatever the incident wave length. When the scattering angle is 90° , the "shift" in wave length between the primary and the scattered beam should be 0.0242 angstrom. This is very nearly in accord with Compton's measurements, quoted above. Further, the shift should be entirely independent of the material of the scatterer, and for various angles θ of scattering it should be proportional to $(1 - \cos \theta)$. To a first approximation these predictions were soon confirmed by the investigations of Compton and his collaborators,¹ Ross,² J. A. Becker,³ Allison and Duane,⁴ and others.

From Eqs. (299), (300), and (304) it is readily found that the energy E_m of the recoiling electron is

$$E_m = h\nu_0 \frac{\zeta(1 - \cos \theta)}{1 + \zeta(1 - \cos \theta)}, \quad \zeta = \frac{h\nu_0}{mc^2}.$$

From (301), in which $v_e < v_0$, it is also clear that, necessarily, $\cos \phi > 0$ and hence $\phi < 90^\circ$, so that the electron must move in the forward

¹Phys. Rev., vol. 21, pp. 207, 483, 715 (1923); Nat. Acad. Sci., Proc, vol. 10, p. 271 (1924).

²Nat. Acad. Sci., Proc, vol. 9, p. 246 (1923); Phys. Rev., vol. 22, p. 524 (1923).

³Nat. Acad. Sci., Proc., vol. 10, p. 342 (1924).

⁴Nat. Acad. Sci., Proc, vol. 11, p. 25 (1925); Woo, Y. H., Phys. Rev., vol. 27, p. 242 (1926).

direction. Following Compton's prediction of their existence, these "recoil" electrons, as they are called, were discovered by C. T. R. Wilson¹ and W. Bothe.² They have been studied quantitatively by Compton and Simon³ and others. Bless,⁴ by the method of the magnetic "spectrograph," similar to Robinson's method for photoelectrons (Sec. 184), has measured the value of E_m for the recoil electrons and has found that the results are in agreement with the last equation.

(b) *The Unmodified Line.*—The presence of the so-called unmodified line (P in Fig. 167) may be accounted for as follows. In setting up our equations, it was assumed that the electron with which the photon collides is free. This assumption should be justified if the energy given to the electron is much larger than the work necessary to detach it from the atom. It is possible, however, that another type of collision may occur in which the electron remains bound to the atom. Such a collision may be regarded as taking place between the photon and the atom as a whole, the mass of which is far greater than that of the electron. If the mass of the atom be substituted for m in Eq. (304), it is readily seen that the computed change of wave length is so small as to be beyond the possibility of detection. It may well be that, in the scattering of photons, the electron sometimes behaves as if bound to the atom and sometimes as if free,—hence the unmodified and the modified scattered lines.

It must be admitted, however, that the simple picture which we have employed above does not harmonize very well with the occurrence of an *unmodified* line. Nor is it clear how the *modified* line is affected by the presence of the nuclear field that normally holds the electron within the atom. These matters, like the properties of atomic states, can only be cleared up by an application of wave mechanics.

The detailed treatment of the impact of X-rays on atoms by wave mechanics leads to the following conclusions,⁵ which, as far as tested, are confirmed by experiment. At low frequencies, the scattered radiation has the same frequency as the frequency ν of the incident radiation, constituting an unmodified scattered line. As the wave length becomes comparable with the dimensions of the electronic shells in the atoms, however, the intensity of the radiation thus scattered diminishes, especially at the larger angles of scattering. Furthermore

¹ *Roy. Soc. Proc.*, vol. 104, p. 1 (1923).

² *Zeits.f. Physik.*, vol. 20, p. 237 (1923).

³ *Phya. Rev.*, vol. 25, p. 306 (1925).

⁴ Dissertation, Cornell University (1927); *Phys. Rev.*, vol. 29, p. 918 (1927). '

⁵ Cf. BLOCH, *Phys. Rev.*, vol. 46, p. 674 (1934); HEITLER, *op. cit.*

as ν becomes larger than the absorption limits for any atomic shell, Compton scattering accompanied by the ejection of recoil electrons out of this shell sets in; and, at considerably greater frequencies, the modified line due to this type of scattering becomes easily observable.

The Compton or modified line is very broad at comparatively low frequencies. This breadth can be thought of as caused by the motion of the electrons in the atom. In our simple deduction of the Compton effect, the electron was assumed to be initially at rest; if it is assumed to have a component of velocity, positive or negative, in the direction of the incident radiation, the wave-length shift is different. Wave mechanics furnishes a probability distribution for the velocities of the atomic electrons, from which the broadening can be computed in

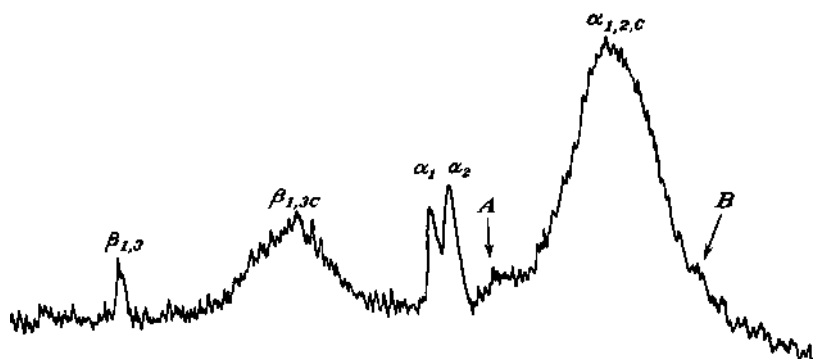


FIG. 169.—Microphotometer curve of Mo K radiation scattered by orientated Ceylon graphite with its diamagnetic axis normal to the natural reference axis of the scattering experiment. (Kirkpatrick and Dumond.)

agreement with experiment.¹ A microphotometer curve showing Mo K radiation as scattered by a block of graphite is shown in Fig. 169; it is taken from a paper by Kirkpatrick and Dumond.²

As ν is increased further, the Compton line becomes narrower. Eventually, in any direction of scattering other than that of the incident beam, the unmodified line becomes weaker than the Compton line, sooner at large angles of scattering than at small angles, and sooner for heavy atoms than for light ones. Finally, only the modified line remains in appreciable intensity; this happens, for example, in light scattered at right angles to the incident beam, for $\nu/\nu_K > 300$ in the case of carbon or $\nu/\nu_K > 50$ in the case of lead, ν_K being the absorption limit for the K shell. The modified line is then fairly sharp and approximately the same in all respects as it would be if the

¹ DUMOND, *Phys. Rev.*, vol 5, p. 1 (1933); BURKHARDT, *Ann. d. Physik*, vol. 26, p. 567 (1936).

² KIRKPATRICK and DUMOND, *Phys. Rev.*, vol. 54, p. 802 (1938).

electrons in the atom were free and at rest. According to theory, the modified line should exhibit the same features as to polarization as the unmodified line, except at extremely high frequencies.

According to the theory, furthermore, the other process by which X-rays may eject electrons from atoms, that of photoelectric absorption, should likewise decrease rapidly with increasing frequency of the incident radiation and should be relatively less important in light atoms than in heavy ones; the probability of the ejection of a photoelectron should become equal to that of the ejection of a Compton electron at about $\lambda = 0.5 \text{ \AA}$ or $v/v_K = 85$ in carbon, or $\lambda = 0.025 \text{ \AA}$ or $v/v_K = 6$ in lead. Thus at high frequencies the ejection of Compton recoil electrons should constitute effectively the only form of action of radiation upon the electrons in matter.

It may be remarked that only the *unmodified* scattered radiation contributes to the beams that are reflected (or more accurately, diffracted) in certain characteristic directions by crystals (Sec. 174) or gratings (Sec. 188) and upon which X-ray spectroscopy depends. Only to this part of the radiation can we apply the classical conception of scattered waves which are coherent with the incident waves, so that the waves scattered by different atoms possess definite relations of phase to one another and are capable of interfering constructively. The waves representing the Compton scattering must be supposed to be completely incoherent.

(c) *The Scattering Coefficient as a Function of Frequency.*—The wave-mechanical theory gives a good account of the observed decrease of the scattering coefficient at high frequencies, which was described in the last section. At frequencies much above the *K* absorption limit of the atom, as has been said, we can treat the electrons in the atom as if they were free electrons, scattering independently by the Compton process. Wave mechanics leads to the following formula for the amount of energy scattered per electron per second by a group of free electrons:¹

$$\Omega = \frac{8\pi}{3} \frac{e^4}{m^2 c^4} I \cdot \frac{3}{4} \left\{ \frac{1 + \zeta}{\zeta^2} \left[\frac{2(1 + \zeta)}{1 + 2\zeta} - \frac{1}{\zeta} \log(1 + 2\zeta) \right] + \frac{1}{2\zeta} \log(1 + 2\zeta) - \frac{1 + 3\zeta}{(1 + 2\zeta)^2} \right\}, \quad (306)$$

where I is the intensity of the incident beam of frequency ν and

$$\zeta = \frac{h\nu}{mc^2}$$

¹ KLEIN and NISHINA, *Zeits. f. Physik*, vol. 52, p. 853 (1929).

(h = Planck's constant, m = rest mass of electron, c = speed of light). As $f \rightarrow 0$, the complicated quantity in braces approaches $\frac{4}{3}$ and 0 approaches the classical value as found by Thomson [cf. Eq. (276) in Sec. 172(a)]. If f becomes large, however, as it is for hard γ -rays, approximately

$$\Omega = \frac{\pi c^4 I}{h \nu m c^2} \left(\log \frac{2 h \nu}{m c^2} + \frac{1}{2} \right), \quad (307)$$

which decreases with increasing ν , varying nearly as $1/\nu$. The scattering coefficient is then in any case

$$\sigma = \frac{n \Omega}{I},$$

where n is the number of electrons per unit volume. This formula fits the existing data very well.

(d) *The Compton Scattering Process.*—It may be of interest to remark that even the *classical* theory leads us to expect radiation that is scattered by *free* electrons to be of somewhat *longer wave length* than the incident beam. For the electrons must take up the momentum that is carried by that part of the incident radiation that is scattered, since the scattered waves, because of the symmetry of their distribution, carry away no momentum on the whole. The electrons will thus be set in motion in the direction of propagation of the incident beam (an effect which can be ascribed to radiation pressure), and the radiation scattered by them will be decreased in wave length because of the Doppler effect. It can even be shown that, if one quantum $h\nu$ is scattered by a free electron initially at rest, the average increase in wave length is of the same order as the Compton shift. But classical theory cannot explain the occurrence of a *sharp line* in the Compton effect. Furthermore, the moving electron would have to move an impossible distance in the scattering substance. Quantum theory is superior in that the entire photon is scattered as a unit by a single electron, and only this one electron is set into motion.

Finally, reference should be made to another experiment, performed much later, which further heightened the contrast between classical and quantum theory by demonstrating that the recoil electron and the scattered X-ray photon are produced *at the same time*.¹ The scattered photon was detected by observing in a cloud chamber the track of a photoelectron released by it at some other point in the scattering material, which was a gas. Recoil tracks and photoelectric

¹ BOTHE and GEIGER, *Zeits. f. Physik*, vol. 26, p. 44 (1924); vol. 32, p. 639 (1925).

tracks due to scattered X-rays were observed to occur simultaneously. This observation shows very clearly the discontinuous nature of the scattering process, which, according to classical theory, should be continuous. Most of the time nothing is happening in the scattering material. At irregular instants, however, a recoil electron is projected from one of the molecules, and almost simultaneously (presumably later owing to the time it takes light to pass from one point to the other) a photoelectron is ejected from another molecule. At such instants, we say that an X-ray photon has been "scattered" by the first molecule and absorbed by the second one.

187. The Refraction of X-rays. (a) *Refraction and Bragg's Law.*—The discovery of the refraction of X-rays and the subsequent measurement of indices of refraction are excellent illustrations of the great advances which have been made in the technique of X-ray measurements since the discovery of the action of the crystal grating. The

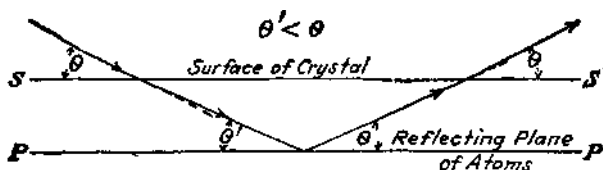


FIG. 170—Refraction of a beam of X-rays entering the surface of a crystal.

first positive evidence that X-rays are measurably refracted came from the work of Stenstrom,¹ who showed from accurate measurements of wave length that Bragg's law of the reflection of X-rays from crystals does not yield identical values when the wave length of a given line is computed from different orders of reflection. Hjalmar² found, for example, that the apparent wave length of the Fe $K\alpha_1$ line as measured in the first order by reflection from a gypsum crystal ($2d = 15.155$ angstroms) was 1.9341 Å; while measurements in the sixth order gave 1.9306 Å, nearly 0.2 percent less.

This apparent failure of the Bragg formula was shown to be due to the refraction of the beam of X-rays as it entered the crystal, the deviation of the beam being such as to indicate *that the index of refraction is less than unity*. In Fig. 170 is shown the path of a ray as it enters the surface SS of a crystal at a glancing angle θ and is incident on the Bragg plane PP at an angle θ' . Bragg's law in the form

$$n\lambda' = 2d \sin \theta',$$

¹ Dissertation, Lund (1919).

² *Zeits. f. Physik*, vol. 15, p. 65 (1923).

where λ' is the wave length *in the crystal*, gives the true law of reflection *at the crystal plane*. Measurements give 0, however, and it is λ , the wave length *in air*, that is desired. By the ordinary laws of optics

$$\mu = \frac{\lambda}{\lambda'} = \frac{\cos \theta}{\cos \theta'}.$$

Substituting values of λ' and of $\sin \theta' = (1 - \cos^2 \theta')^{1/2}$ in the preceding equation, we find $n\lambda/\mu = 2d(1 - \cos^2 \theta/\mu^2)^{1/2}$, whence

$$\begin{aligned} n\lambda &= 2d(\mu^2 - \cos^2 \theta)^{1/2} = 2d(\mu^2 - 1 + \sin^2 \theta)^{1/2} \\ &= 2d \sin \theta \left[1 + \frac{(\mu - 1)(\mu + 1)}{\sin^2 \theta} \right]^{1/2}. \end{aligned}$$

Since $1 - u$ is extremely small for X-rays, we can replace $u + 1$ by 2 in this expression, and then expand in powers of $1 - u$ and keep only the first power. Thus we obtain

$$n\lambda = 2d \sin \theta \left(1 - \frac{1 - \mu}{\sin^2 \theta} \right). \quad (308)$$

The latter formula can be used to calculate values of $\delta = 1 - u$ from observations in which different orders of diffraction are compared.

A few of the values of δ as found by Larsson¹ are as follows:

$\lambda, \text{\AA} \dots \dots \dots$	1 537	2 499	3.447	5.166	7 111	8.320
$\delta \times 10^6, \text{ mica} \dots \dots$	8 94	24 6	49 1	103	182	262
$\delta \times 10^6, \text{ calcite} \dots \dots$	8.8	22.4	41.9			

(b) *Theory of the Refraction of X-rays.*—Classical theory gives for the index of refraction u of a slightly refracting and absorbing substance, for waves of frequency ν ,

$$\mu^2 = 1 + \frac{e^2}{\pi m} \sum_i n_i \frac{\nu_i^2 - \nu^2}{(\nu_i^2 - \nu^2)^2 + \gamma_i^2 \nu^2},$$

where n_i is the number of electrons per unit volume which have a natural frequency ν_i (of undamped vibration), γ_i is a constant representing the effect of damping, and the summation is to be extended to include all of the different groups of electrons in the refracting material.² If ν is very much greater than all of the characteristic frequencies ν_i , and all γ_i 's are small, we may neglect ν_i^2 and γ_i^2 in comparison with ν^2 ; if also u differs very little from 1, we may write

¹ Quoted in SIEGBAHN, *loc. cit.*

² See WOOD, R. W., "Physical Optics," 3d ed., p. 488, 1936.

$\mu^2 = [1 + (\mu - 1)]^2 = 1 + 2(\mu - 1)$, approximately. Thus for very large ν the last formula can be written:

$$\mu = 1 - \delta, \quad \delta = \frac{ne^2}{2\pi m \nu^2}, \quad (309)$$

where $n = \sum_i n_i$, and represents the total number of electrons per unit volume. Or, substituting $\nu = c/\lambda$, we have

$$\frac{\delta}{\lambda^2} = \frac{ne^2}{2\pi mc^2}. \quad (310)$$

Thus, for frequencies much greater than any characteristic frequencies of the refracting substance, the quantity δ/λ^2 should be independent of the incident frequency.

Analogous results are obtained from wave mechanics. In place of n_i , a coefficient occurs that is mostly less than 1. The frequencies ν_i

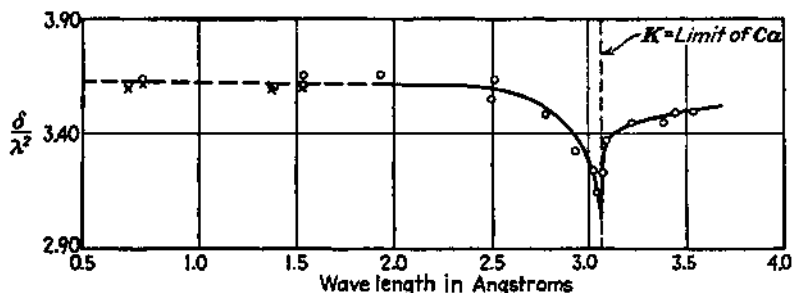


FIG. 171.—Values of $5/X^2$ for calcite, showing anomalous dispersion near the Ca K absorption limit. [X = data by Bearden and Shaw; O = data by Larsson, thesis, Univ. Uppsala, Ann. Reports (1929).]

correspond to the various transitions that the atom can undergo, either into a higher quantum state of the neutral atom or into an ionized state as the result of a process of photoelectric absorption. The terms corresponding to photoelectric absorption, however, have an appreciable effect upon the dispersion only when the incident frequency ν is close to one of the absorption limits of the atom. At other frequencies Eq. (310) should hold.

Experiment reveals a behavior of the refractive index in good agreement with these theoretical predictions. For example, in Fig. 171 is shown the observed variation of δ/λ^2 with λ for calcite in the neighborhood of the K absorption limit of Ca at 3.064 Å. The dotted portion of the curve represents the value $\delta/\lambda^2 = 3.67$, computed for calcite from Eq. (310). The equation is seen to hold very well from 0.5 to 2.5 Å.

(c) *Direct Measurement with a Prism.*—Although the index of refraction for X-rays differs only slightly from unity, its direct measurement by a prism has been accomplished by several observers. A right-angled prism may be used, the X-rays entering one face at a glancing angle of only a few minutes of arc and emerging almost perpendicularly from the opposite face; the prism may be substituted for the mirror M in the arrangement shown in Fig. 172, which is described in the next section. With modern technique, the accuracy of such methods may be very high; the prism provides, in fact, a method for the measurement of absolute X-ray wave lengths with a precision of 1 part in 10,000.¹ The wave lengths are calculated by means of the complete formula furnished by wave mechanics. The fact that wave lengths so measured agree with the same wave lengths as determined by other methods tends to confirm the correctness of the wave-mechanical theory of X-ray dispersion.

(d) *Measurement by Total Reflection.*—Since the index of refraction of a material for X-rays is less than unity, a beam of X-rays incident on a polished surface at a sufficiently small glancing angle should be totally reflected. According to the usual law of optics, this should occur for *glancing* angles less than θ_R where

$$\cos \theta_R = \mu = 1 - \delta, \quad \sin \theta_R = \theta_R = \sqrt{2\delta}$$

for small values of θ_R . Since, for a given substance, δ/λ^2 is approximately constant, it follows that, to the same approximation, $\theta_R \propto \lambda$. From the value of δ determined by methods described above, θ_R may be computed. Several representative values of θ_R are shown in Table VI.

TABLE VI.—SOME VALUES OF THE CRITICAL ANGLE OF TOTAL REFLECTION θ_R

Substance	λ , angstroms	δ	θ_R
Glass ...	0.7078	1.64×10^{-5}	6' 10''
Calcite.....	1.537	8.80	14' 25''
Calcite.....	3.734	49.2	34' 5''
Quartz.....	10.0	356	91' 40''

Reversing the procedure, one may observe B_R and compute δ .

This method was used by A. H. Compton in one of the first measurements of X-ray refraction.² The measurements of Doan³ are

¹ BEARDEN and SHAW, *Phys. Rev.*, vol. 46, p. 759 (1934); BEARDEN, *Phys. Rev.*, vol. 54, p. 698 (1938).

² *National Research Council, Bull.* 20, p. 50 (1922).

³ *Phil. Mag.*, vol. 4, p. 100 (1927).

instructive. Doan's apparatus, following Compton's method, is shown schematically in Fig. 172. A beam of X-rays from a target T , after passing through a slit S_1 , falls onto a crystal K , by means of which one of the characteristic lines in the spectrum of T is reflected at a very small glancing angle onto the mirror M . With a sufficiently small angle of incidence, this beam is totally reflected from M and

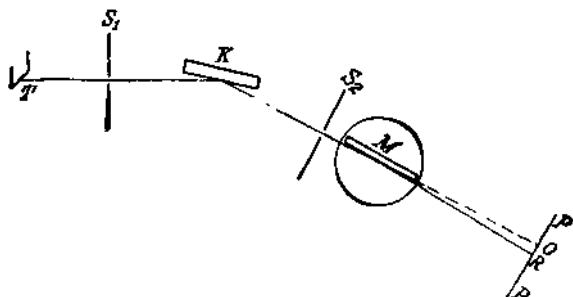


FIG. 172.—Doan's apparatus for measuring the index of refraction of X-rays by the method of total reflection.

falls on a photographic plate PP at R . The mirror M is turned very slowly during an exposure, and, when the critical angle is reached, reflection ceases. This critical angle can be determined by noting the distance on the plate between the point 0 , which indicates the position of the direct beam, and the extreme edge of the record of

TABLE VII.—INDEX OF REFRACTION OF X-RAYS BY TOTAL REFLECTION

Wave length, angstroms	Substance	Critical angle	$\delta = 1 - \mu, \times 10^8$
0.7078	Ni	10' 15"	4.42
0.7078	Ag	11' 42"	5.76
1.389	Ni	16' 9"	10.98
1.389	Cu	19' 36"	16.2
1.537	Ni	24' 40"	25.5
1.537	Cu	20' 24"	17.7
1.537	Ag	26' 42"	30
1.537	Au	31' 24"	41.6

the reflected beam. The metals under study were sputtered onto a highly polished optical glass surface.

Table VII shows some of the results obtained by Doan.

The accuracy of this method is not high, however, since the critical angle of total reflection for any given wave length is sharply defined only if the coefficient of absorption for that wave length is negligible—which is by no means the case with X-rays, particularly of

the longer wave lengths. For example, the value of δ for Cu $K\alpha_1$ radiation ($\lambda = 1.537 \text{ \AA}$) reflected by silver is approximately 30×10^{-6} , from which $\theta_R = \sqrt{2\delta} = 27$ minutes of arc (about). Were there no absorption, the coefficient of reflection for glancing angles up to 27 minutes of arc would be as shown by the dotted line in Fig. 173. Nahrings,¹ however, shows that, using classical theory and the known coefficient of absorption of silver for $\lambda = 1.537 \text{ \AA}$, the coefficient of reflection R for various glancing angles of incidence θ should be as given by the full line of Fig. 173. The dots represent Nahrings's observations.

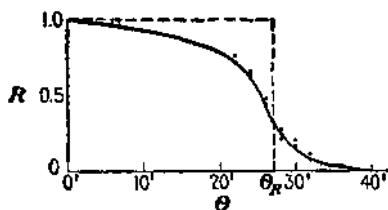


FIG. 173.—Coefficient of reflection R of silver for Cu $K\alpha_1$ radiation ($\lambda = 1.537$ angstroms) at various glancing angles of incidence θ .

We note that, although $R > 0.5$ for $\theta < 25'$, $R = 1$ only at $\theta = 0$.

188. Measurement of X-ray Wave Lengths by a Ruled Grating.

With the discovery of the total reflection of X-rays, the possibility arose that a ruled grating might be used to measure X-ray wave lengths in exactly the same way that a grating is similarly used in the optical region, provided that the glancing angle between the X-ray beam and the ruled surface is less than the critical angle for total reflection. Compton and Doan² were the first to make measurements of this kind. Using a grating of speculum metal with 50 lines per millimeter,³ they found the wave length of the $K\alpha_1$ line of molybdenum to be $\lambda = 0.707 \pm 0.003$ angstrom.

For a description of the development of the technique of using ruled gratings for the measurement of X-ray wave lengths, the reader is referred to original articles.⁴ We have already mentioned (Sec. 175) the precise work of Bearden,⁵ who measured the wave length of the Cu $K\alpha_1$ line ($1.5406 \text{ \AA}$) with gratings having 100 or 300 lines per mm. ruled on glass sputtered with gold. His results are probably correct within less than 0.01 percent.

For several years, there existed a discrepancy of 0.2 to 0.3 percent between ruled-grating and crystal values of X-ray wave lengths. As

¹Phys. Zeits., vol. 31, p. 799 (1930).

²Nat. Acad. Sci., Proc., vol. 11, p. 598 (1925).

³The arrangement used by Compton and Doan may be indicated schematically by replacing the mirror M , Fig. 172, by the grating. The glancing angle was less than 25 minutes of arc.

⁴THIBAUD, J., *de Physique et le Radium*, vol. 8, pp. 13, 447 (1927); BACKLIN, Dissertation, Uppsala (1928); BEARDEN, *Phys. Rev.*, vol. 33, p. 1088 (1929); HOWE, *Rev. Sci. Instruments*, vol. 1, p. 749 (1930).

⁵*Phys. Rev.*, vol. 48, p. 385 (1935).

stated in Sec. 175, however, this discrepancy appears now to have been removed by the discovery of an error in the crystal values, due in turn to an error in the value assumed for the electronic charge. Measurements made with ruled gratings seem to afford the most reliable means of determining the absolute magnitudes of X-ray wave lengths, since they involve no assumptions as to the homogeneity of a crystal; in fact, the only elements that enter into the determination by this method are the wave theory of light as propagated in a vacuum and such well-tested operations as the measurement of angles and the counting of lines under a micrometer microscope.

189. The Nature of Electromagnetic Radiation.—Sufficient acquaintance has now been attained with the various properties of radiation as revealed by modern experimentation to serve as a basis for final consideration of the question left unanswered in Sec. 49. What is the true nature of electromagnetic radiation? This question arises in its sharpest form in contemplating the contrasting properties of X-rays. How can an entity exhibit, on the one hand, the wave behavior that is evidenced in the diffraction of X-rays by crystals and, on the other hand, the particle properties that are revealed by the Compton effect?

According to quantum mechanics, as interpreted by Bohr and other theoretical physicists, the riddle can be solved only by extending to radiation the same limitation upon the use of ordinary space-time conceptions that was found to be necessary in dealing with electrons. Bohr insists that whenever an observation of any sort is made, its *immediate* results will always be expressible in terms of familiar ideas of space and time, since these ideas have been developed out of human experience and any observation necessarily includes as its primary stage a certain experience by a human observer. But it does not follow that it will always be possible to construct a picture of the physical reality that causes these experiences, in the same way in which we picture everyday objects. In classical theory, an electromagnetic field was assumed to have a certain character at every point in space and at every instant of time, as represented by certain values of the electric and magnetic vectors. According to the new view, such a conception of the electromagnetic field is valid at best as an approximation, and only in certain cases.

The essential significance of what we call a radiation field lies in its effects upon the motion of charged particles. In describing this motion we encounter the limitations that are expressed in the indeterminacy principle as described in Sec. 112; and these limitations make it possible for the field to exhibit contrasting characteristics under different circumstances.

At one extreme, the action of the field takes the form with which we have become familiar in the Compton effect. Here a photon appears to bounce off an electron, thereby changing the momentum of the electron suddenly and discontinuously. The change of momentum δp cannot exceed $2h\nu/c$, however, according to Eq. (301). Furthermore, this change of momentum will be definitely observable and measurable by the physicist only if it considerably exceeds the range of indeterminateness of the momentum already possessed by the particle, which cannot be less than Δp as given by Eq. (180) or $\Delta p = h/\Delta x$. The maximum possible ratio of δp to Δp is thus

$$\frac{\delta p}{\Delta p} = \frac{\frac{2h\nu}{c}}{\left(\frac{h}{\Delta x}\right)} = 2 \frac{\Delta x}{\lambda} \quad (311)$$

where $\lambda = c/\nu$ and represents the wave length of the radiation, whereas Δx measures the indeterminateness in position of the particle.

From the last equation it is evident that a clear-cut Compton scattering process can occur only when, to make $\delta p/\Delta p$ large, *the wave length of the radiation is much shorter than the diameter of the region in which the particle may be supposed to be effectively located.* For an electron in an atom, this condition can be met only for wave lengths considerably shorter than the atomic diameter, such as the wave lengths of hard X-rays or γ -rays.

In the Compton process there is no feature that can be regarded as a manifestation of an electric intensity in the wave. In order to obtain, on the other hand, an action of radiation that can be interpreted in terms of the familiar electric and magnetic vectors, two conditions must be met. The experimental conditions must be such that it is possible to follow the test particle in continuous motion along a path, so that its acceleration can be determined. It is also necessary that a segment of path which is sufficiently long to permit an adequate determination of the acceleration shall yet be short enough so that along it the field vectors do not vary appreciably in value. The first condition requires that the segment of path shall be many times as long as Δx ; and the second requires that it must be much shorter than λ . It follows that necessarily Δx is much shorter than λ .

Thus the familiar action ascribable to electric and magnetic fields in the radiation is obtained only when the particle acted on is definitely located within a region much smaller than a wave length.

The latter condition is satisfied, for example, by the electrons in a wire held parallel to the electric vector in long electromagnetic waves. In such a wire alternating electric current is observed to be produced,

varying in phase with the electric vector. Ions in the upper atmosphere acted on by such waves furnish another example.

In such cases the magnetic field in the waves will also act upon the current or on the moving ions. It thus comes about that the *average* result is a force in the direction of propagation of the waves. This force effectively constitutes an example of light pressure, and it can be regarded as analogous to the Compton recoil that occurs under other circumstances; but here there is a continuous rather than a discontinuous action. It might be thought that in such cases the photons are simply too small to be detected individually, but a better view is probably that in such cases there are no photons at all. The photon is thus an apparition that is evident under some circumstances but not under others.

The diffraction of X-rays by crystals is another example of vector action; here the phase relations of waves scattered by different atoms are involved, and in applying the indeterminacy principle the scattering electrons must be regarded as attached to the crystal as a whole rather than to individual atoms.

In intermediate cases it will be impossible to obtain clear-cut evidence either for the occurrence of photons and of Compton recoils, or for action by electric and magnetic vectors; the difference between these two modes of action will be more or less concealed under statistical variations which are closely related to the indeterminacy of motion of the particle.

Thus we are in part prevented by indeterminacy of particle motion from even asking those questions concerning the nature of radiation that are naturally suggested by classical lines of thought. It seems probable that these questions really have no complete answers, and that the radiation field cannot be pictured in the full space-time detail with which we form, for example, a picture of the waves on the sea. One kind of picture is appropriate to one phenomenon, another kind to another phenomenon; but in no case can the picture be lined in completely without going beyond observable phenomena, and no one type of picture will cover all cases.

SOME RECENT DEVELOPMENTS IN X-RAY SPECTROSCOPY

Within the scope of this book, we cannot give the "latest word" on any of the various ramifications of the subjects introduced. The selection of a few topics for very cursory discussion in this last part of the chapter on X-rays is, perforce, arbitrary. For an acquaintance with the many other interesting lines of current research in this field, the student is referred to the literature.

190. Multiple Ionization of Inner Electron Shells.—Our knowledge of the inner structure of the atom has been advanced greatly by work on such X-ray phenomena as satellites, line widths, and relative intensities of lines and the theoretical explanations of these phenomena. The explanations in question all involve atomic states characterized by the absence of 2 or more inner electrons.¹

(a) *Satellites or Second-order Lines.*—It was perhaps fortunate that the spectral apparatus available to Moseley and the early workers did not have the sensitivity and resolving power of present-day spectrometers. The lines which they observed were the more intense

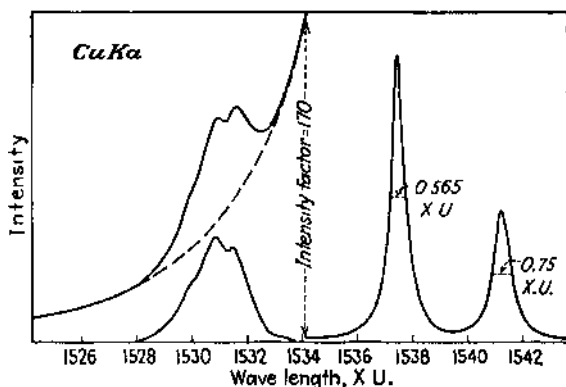


FIG. 174.—The $K\alpha$ doublet of Cu (29) with accompanying satellites. The left-hand half of the curve is plotted on a scale 170 times larger than the right-hand half, and below it is plotted the intensity due to the satellites alone, obtained by subtracting the presumable intensity due to other lines as represented by the dotted curve.

and more easily resolvable lines of X-ray spectra. These are the lines which are represented on the conventional energy-level diagram (Sec. 179) and were readily interpreted in terms of the extended Bohr theory of atomic spectra. As we have seen, these lines are due to transitions between states of *single ionization*, and are called *first-order* lines.

With improvements in technique, many other lines were subsequently discovered which did not fit into the conventional diagram. The majority of these lines were rather faint, were usually found close to and on the short-wave-length side of the more intense "diagram" lines, and hence were called "satellite" lines. A typical spectral curve of the satellite structure accompanying the $K\alpha$ lines of copper is reproduced² in Fig. 174. The satellites accompanying the $L\beta_1$,

¹ For a more detailed discussion than that given here see Richtmyer, *Rev. Modern Phys.*, vol. 9, p. 391 (1937).

² PABBATT, *Phys. Rev.*, vol. 50, p. 1, (1936).

$L\beta_{2p}$ and $L\gamma_1$ lines for Ag (47) are designated by superscripts in Fig. 154. As illustrated in these figures, the satellite structure is observed to be very complex, containing numerous component lines of various intensities. Most (if not all) first-order lines are accompanied by such satellite structure, varying from line to line as well as from element to element. The total number of component satellite lines now known far exceeds the total number of diagram lines. If such a confusing array of lines had been presented to Moseley and the other early workers, it is probable that progress in X-ray spectroscopy would have been much less rapid.

In seeking a plausible interpretation of the origin of satellite lines, we may focus our attention upon such key characteristics as the following: (1) the minimum excitation voltage, or, if the lines are observed in fluorescence, the minimum photon energy for excitation; (2) wave-length positions of the lines; (3) variation in the relative intensities of the satellite lines with the atomic number of the radiating material, or with the voltage and current in the X-ray tube.

Because of the low intensity of the satellites, reliable experimental information on such characteristics is extremely difficult to obtain. After several trials, however, the excitation voltage of certain satellites was definitely shown to be somewhat greater than the excitation voltage of the accompanying first-order or parent line.¹ In the case of the type of *K* satellites that is illustrated in Fig. 174, the energy of excitation is found to be equal to the energy required to eject a *K* electron and *in addition* an *L* electron from the atom, the energies required to eject these electrons being calculated from the corresponding absorption limits. Hence, we may assume, as a working hypothesis, that the initial state for the emission of these satellite lines is a state of *double ionization*, in which the atom has an electronic vacancy in both the *K* shell and the *L* shell. Such a state of the atom may be called a "*KL* atomic state." In a similar way, other states of double ionization, such as *KK*, *KM*, *LM*, etc., should be possible.²

An atom in a *KL* state may undergo a radiative transition into any one of a number of other states of double ionization, *e.g.*, $KL \rightarrow KM$ (an electron dropping from the *M* shell into the *L* shell), or $KL \rightarrow LL$ (an electron dropping from the *L* shell into the *K* shell). Estimates of the atomic energy indicate that the loss of energy should be slightly

¹See, for example, DRUYVESTYEN, *Zeits. f. Physik*, vol. 43, p. 707 (1927); PARRATT, *Phys. Rev.*, vol. 49, p. 132 (1934); COSTER, KUIPERS, and HUIZINGA, *Physica*, vol. 2, p. 870 (1935).

²This theory of satellites was proposed by Wentzel, *Ann. d. Physik*, vol. 66 p. 437 (1921) and modified by Druyvesteyn, *Zeits. f. Physik*, vol. 43, p. 707 (1927).

greater in the transition $KL \rightarrow LL$ than in the diagram transition $K \rightarrow L$, which gives rise to the $K\alpha$ lines; hence, the former transition should give rise to satellites close on the short-wave side of the $K\alpha$ lines. Similarly, the transition $KL \rightarrow LM$ should give rise to satellites on the short-wave side of the $K\beta$ lines. In a doubly ionized atom, the two vacancies would function spectroscopically in much the same way as would 2 valence electrons (Sec. 140). Furthermore, the vacancies may have different l values corresponding to the various subshells; for example, the satellites associated with the La diagram line are assumed to originate in transitions between the levels L_{III} , M_{IV} or L_{III} , M_V (written compactly as $L_{III}M_{IV,V}$) and the levels $M_{IV}M_{IV}$, $M_{IV}M_V$, $M_V M_V$ (or $M_{IV,V}M_{IV,V}$).

In this way we may explain the multiplicity of the observed satellite structures. As far as it has been possible to make approximate theoretical calculations, the predictions of wave mechanics lend support to the explanation just described for the origin of satellites.¹

Some satellites have been reported which are believed to originate in transitions between states of *triply ionized* atoms. Thus, we may have satellite lines of second and higher orders, in analogy with the various higher orders of optical "spark" spectra. The construction of an energy-level diagram for second- (or higher) order lines in X-ray spectra is extremely difficult, however, and in general, has not been accomplished either empirically or theoretically. The theory of satellites has scarcely advanced beyond the qualitative stage, and it is entirely possible that some of these lines originate in some type of atomic process other than that just described.²

(b) *The Auger Effect.*—In the discussion of satellites, we assumed the atom to be doubly ionized without considering how this state might be brought about. Presumably, it is possible for a cathode-ray electron to eject 2 electrons at once from an atom. If this is the origin of the doubly ionized atoms, theoretical estimates indicate that the intensity of satellites relative to the parent lines should decrease in a continuous manner with increasing atomic number. Such a variation with atomic number is found by experiment to hold for K satellites, i.e., those accompanying lines of the K series, but not for L or M satellites. *The intensity of the satellites accompanying the La line, for example, is observed to decrease rather abruptly as the atomic number increases from 47 to 50 and to increase again rather abruptly at about 75; between*

¹ For discussion and references, see Richtmyer and Ramberg, *Phys. Rev.*, vol. 61, p. 925 (1937).

² See, e.g., RICHTMYER, *Frank. Inst. J.*, vol. 208, p. 325 (1929); BLOCH, *Phys. Rev.*, vol. 48, p. 187 (1935).

atomic numbers 50 and 75, *La* satellites are practically unobservable. This anomalous behavior as to intensity prevented for a time the universal acceptance of the Wentzel-Dryvesteyn theory of satellite origin. The difficulty was resolved when Coster and Kronig pointed out the importance, in this connection, of another physical process known as the "Auger effect."¹

Under certain conditions of energy, an atom in a state of single ionization may undergo spontaneously an *Auger transition* which leaves it in a state of double ionization. The electron released in such a transition is, in the final state, expelled from the atom and left with an amount of kinetic energy E given by

$$E = E_i - E_f, \quad (312)$$

where E_i is the initial energy of the singly ionized atom and E_f is the final energy of the doubly ionized atom. This type of atomic transition does not involve the emission of a photon and, therefore, is often called a "nonradiative" transition. Nonradiative transitions are possible, in general, only when an electron can be ejected from the atom; and for this to be possible it is necessary that E as defined by Eq. (312) be equal to or greater than zero.²

Electrons ejected by such radiationless transitions were first observed by Auger, who detected the short, fat tracks made by them in the gas of a cloud chamber (Sec. 203) irradiated with X-rays.³

A great many possible Auger transitions may occur satisfying Eq. (312), starting from a given initial state, and, furthermore, many different initial states are possible. To illustrate a particular transition, suppose E_i refers to the L_I state of an atom of atomic number Z , i.e., E_i is the energy of the L_I absorption limit. If the final state for an Auger transition is $L_{III}M_{IV}$, the value of E_f is the sum of the energies required to remove, first, an electron from the L_{III} shell and, then, a second electron from the M_{IV} shell. Of these two energies, the first is that corresponding to the L_{III} absorption limit of the atom of atomic number Z ; the second is practically equal to the energy corresponding

¹ COSTER and KRONIG, *Physica*, vol. 2, p. 13 (1935).

² For the benefit of students who may later study wave mechanics, it may be remarked that atomic states from which an Auger transition is possible do not represent true stationary states, or quantum states, of the atomic system. The atom in such a state cannot be represented by a Ψ function of the type of that written in Sec. 120 but must be represented by a Ψ or wave packet (Sec. 111) of a slightly more general form. As time goes on, Ψ changes, slowly or rapidly, into a form representing an additional electron of the atom as free and the others as remaining in the atom, which is in a new state of higher ionization.

³ AUGER, *Comptes Rendus*, vol. 180, p. 65 (1925).

to the M_{IV} absorption limit of an atom of atomic number $Z + 1$, since the absence of the L_{III} electron will cause the atomic field acting on the M_{IV} shell of electrons to resemble more closely that in an atom of nuclear charge increased by unity. Hence, as a good approximation, we may write as the energy equation for the transition $L_I \rightarrow L_{III}M_{IV}$

$$E = (E_{LI})_Z \sim (E_{LII})_Z - (E_{MIV})_{Z+1}. \quad (313)$$

Clearly, nonradiative or Auger transitions provide a means by which an atom can leave a given atomic state without emitting a photon, and also a means whereby atoms singly ionized by cathode-ray

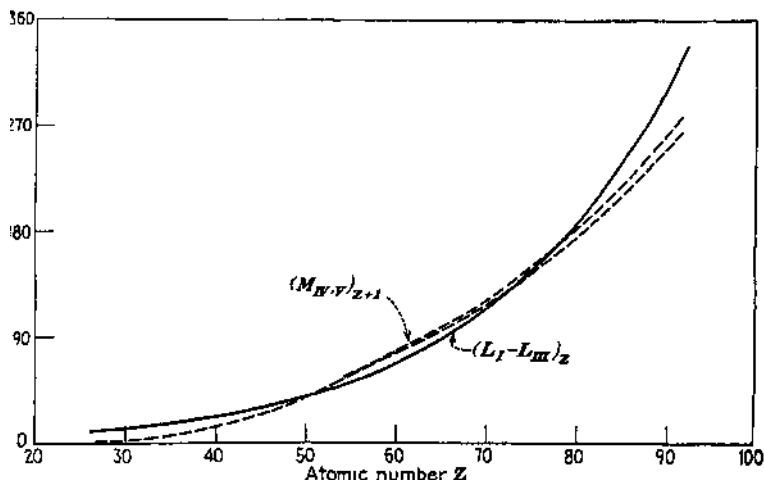


FIG. 175.—Coster-Kronig diagram for the L satellites; curves showing v/R value of the $L_I - L_{III}$ energy difference for an atom of atomic number Z , in comparison with the binding energies of M_{IV} and My electrons in atoms of atomic number $Z + 1$.

bombardment may automatically become doubly ionized. The Auger transition described in the preceding paragraph, if it occurs, will (1) weaken the L lines that start from the L_I atomic state by decreasing the number of atoms in that state, and (2) increase the intensity of the accompanying satellite lines that start from the $L_{III}M_{IV}$ states.

In order that the Auger transition $L_I \rightarrow L_{III}M_{IV}$ may occur, the right-hand member of Eq. (313) must be zero or positive, for E cannot be negative. In other words, we must have

$$(E_{LI} - E_{LII})_Z \geq (E_{MIV})_{Z+1}. \quad (314)$$

The range of elements for which this condition is satisfied can be determined from a table¹ of the energies. In Fig. 175 are plotted

¹ SIEGBAHN, loc. cit.

values of the two members of Eq. (314) and of the similar equation for the transition $L_I \rightarrow L_{III}M_{V, II}$ as functions of Z . It is evident from the figure that the Auger transitions in question are possible for $Z < 50$ and for $Z > 75$ (about). These are just the regions of Z for which the intensity of the La satellites is anomalous. Hence, if we suppose that few atoms in the states $L_{III}V, V$ are produced directly by the cathode-ray bombardment, so that such states are produced chiefly by Auger transitions, the peculiar behavior of the La satellites as to intensity is explained.

In a similar way, the Auger effect involving other nonradiative transitions has served to explain other anomalies of satellite intensity.

The second effect expected to follow from the occurrence of Auger transitions is that lines for which L_I is the initial state should be weakened and, hence, should undergo a rather abrupt change in intensity at $Z = 50$ and $Z = 75$ (about). This effect, also, has been observed experimentally, as well as similar Auger effects on other lines.¹

Finally, it may be mentioned that the Auger effect furnishes an explanation, which cannot be discussed further here, of the great broadness of all lines except those of the K series.²

191. X-ray Spectra and the Outer Part of the Atom.—The usual emphasis in the study of X-ray spectra is on phenomena involving *inner* atomic structure. It is possible largely to ignore concurrent changes that may occur in the *outer* part of the atom because these changes have relatively little effect upon the energy levels arising from vacancies in the innermost shells. Certain finer features of X-ray spectra, however, involve in their explanation a consideration of the outer part of the atom, or even, in a solid, of the surrounding material. Only the comparatively simple case of a gaseous substance will be discussed further here.

Emission spectra from gaseous atoms have not been studied in sufficient detail to reveal the effects of transitions involving the outer part of the atom because of the extremely low intensities of such spectra, but several *absorption* spectra have been observed in which such effects are noticeable.

In our discussion of the absorption of X-rays in Sec. 178(6) and elsewhere, it was always assumed that, when an X-ray photon is absorbed, an electron is removed entirely from an atom of the absorb-

¹ COOPER, J. N., Cornell Dissertation; *Phys. Rev.*, vol. 59, p. 473 (1941).

² Cf. RICHTMYER, BARNES, and RAMBERG, *Phys. Rev.*, vol. 46, p. 836 (1934); PARRATT, *Phys. Rev.*, vol. 54, p. 99 (1938); COOPER, *Phys. Rev.*, vol. 61, p. 234 (1941).

ing material. If the material is gaseous, however, it should be possible for the electron to stop in some outer vacancy in the atom. In this case the absorbed energy $h\nu$ would be less than if the electron were removed to infinity.

Wave mechanics furnishes the following basis for a theory of such transitions. With an electron removed from an inner shell, the ion will be surrounded by a field, due to the nucleus and the remaining electrons, which will approximate that of a hydrogen nucleus. For an electron moving in this field, there will exist a discrete set of quantum states, with associated wave functions. The field can be regarded as the extension outside of the atom of the equivalent central field that is introduced in the zero-order stage of perturbation theory (Sees. 118 and 120). The electrons remaining in the atom can be regarded as occupying the innermost electronic quantum states in this field, leaving all the rest of them unoccupied. It should, therefore, be possible for the electron that is removed from an inner shell, instead of proceeding to an indefinite distance, to stop in one of these unoccupied outer states.

According to this theory, we are led to expect that, instead of a single atomic level, such as the K level, as hitherto described, there would be a closely spaced *sequence of energy levels*, as suggested in Fig. 176. The uppermost of these discrete levels represents removal of the electron into a state of rest at infinity and is thus a level belonging to the ionized atom. The frequency ν_K corresponding to such removal of an electron from the K shell might be called the '*photoelectric absorption limit for a gaseous absorber*'. The lower-lying levels, corresponding to states in which the electron removed from the K shell remains attached to the atom in one of the outer, previously unoccupied electronic quantum states, correspond to states of the neutral atom (and are sometimes called resonance levels). Above the absorption limit there lies then the further continuum of levels representing energies sufficient to ionize the atom and to leave the ejected electron with an excess of kinetic energy.

A graph of the absorption coefficient should show, therefore, the well-known photoelectric absorption extending toward higher frequencies from the K photoelectric absorption edge, which is associated with ionization of the atom, and, in addition, on the long-wave side of this edge, a series of absorption lines, each arising from atomic transitions into one of those lower-lying atomic levels of the neutral

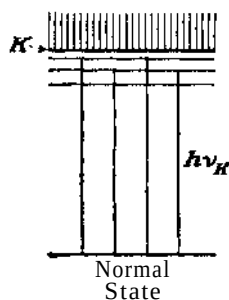


FIG. 176.-Illustration of a K photoelectric absorption limit and associated resonance levels.

limit,¹ as observed with a spectrometer of very high resolving power. There is clear evidence of the type of resonance absorption structure just described, but the absorption lines are so broad that there is great overlapping. The electronic configuration for argon ($Z = 18$), $1s^2 2s^2 2p^6 3s^2 3p^6$, will be altered in nonionizing absorption to $1s 2s^2 2p^6 - 3s^2 3p^6 np$, a p state being the only kind to which an electron can change from a $1s$ state in accordance with the selection rules. The spacing of the atomic levels under discussion will thus be almost the same as that of the optical P levels of potassium ($Z = 19$). Hence, if we ascribe the most intense line in Fig. 177 to the electronic transition $1s \rightarrow 4p$, we can locate the positions of the other resonance lines and of their series limit by using the known optical P terms for potassium. The photoelectric absorption edge should have a finite "width" when observed with such high resolving power; theory indicates that the shape should be given by an arc-tangent curve, whose width is due to the same factors that cause the width of each of the absorption lines and also of the emission lines.² The theoretical absorption edge is drawn in the figure as the left-hand dotted curve, centered at the position of the P series limit.

192. X-ray Spectroscopy of Solids.—As mentioned in the previous section, the arrangement of the electrons in the outer part of an atom must undergo considerable alteration when the atom is forced into close proximity with many other atoms, as in a solid. Consequently; we may expect the observed structure of absorption edges for atoms in a solid to be materially different from that found for a gaseous absorber. Nevertheless, some type of resonance absorption is to be expected. Just as an interpretation of a curve such as that in Fig. 177 may provide us with information about the structure of a free atom, so may an interpretation of an absorption curve for a solid absorber reveal information about the basic structure of a solid. Furthermore, with a solid target in the X-ray tube, we may study also the emission spectra involving the outer part of solid atoms and thus supplement the information obtained from the absorption curve. Based on experiments and studies of this sort, with such help as can be derived from an application of wave-mechanical theory, a *solid-state spectroscopy* is being developed.

A typical absorption curve for a solid absorber, showing the K regions for chlorine and for potassium is reproduced³ in Fig. 178.

¹ PARRATT, *Phys. Rev.*, vol. 56, p. 295 (1935). Resonance absorption structure, also found for polyatomic gases, was first identified by Kossel, *Zeits. f. Physik*, vol. 1, p. 119 (1920).

² See reference no. 2. on p. 536.

³ TRISCHKA, *Phys. Rev.*, vol. 67, p. 318 (1945).

It was obtained by passing a beam of X-rays through a thin film of KCl, which was prepared by evaporation and was estimated to be about 35,000 angstroms thick. The curves show obvious indications of resonance absorption, but no distinct lines stand out. Wave-mechanical considerations indicate that after an electron has been removed from the interior of an atom it will have an electronic wave function that represents it as belonging to the entire solid rather than to an individual atom, and that the energies associated with such transitions are distributed over broad bands of energy instead of being confined to a discrete set of values. The theory of solid-state spec-

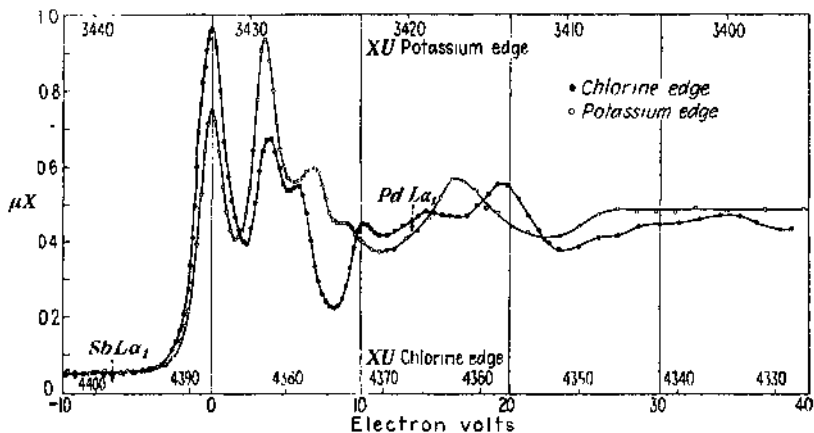


FIG. 178.—Absorption curves in two regions containing the K edges of chlorine and of potassium, respectively, brought into superposition for purposes of comparison by a shift of the wave-length scale. (Here XU denotes 10^{-11} cm.) The abscissa represents the photon energy $h\nu$ in ev. The ordinate represents μx in the factor $e^{-\mu x}$, to which the transmitted intensity is proportional, x denoting distance along the beam. {After Trischka.}

troscopy has not developed, however, to the point that we can make a clear interpretation of the details of the observed structure of the absorption spectrum. For example, we cannot locate unambiguously the wave length of the photoelectric absorption edge of a solid absorber.

The most fruitful work in emission in solid-state spectroscopy has been in the long-wave-length region,¹ from 5 to 200 Å. Spectrographs using a mica crystal were employed to about 20 Å and a grazing-incidence ruled grating for $X > 20$ Å. In such work, one studies, in particular, the X-ray emission lines that involve transitions of the atomic valence electrons, which have now come to belong to the solid as a whole. The effect of the interaction with neighboring

¹ SKINNER, *Phys. Soc. Reports*, vol. 5, p. 257 (1939), and references. See, however, reports by J. A. Bearden and collaborators, *Phys. Rev.*, vol. 58, pp. 387, 396, 400 (1940).

atoms is equivalent to replacing the sharp levels of the free atom by bands of various widths. Levels associated with vacancies deep in the interior of the atom, such as the K levels in the heavier atoms, are only very slightly broadened. At the opposite extreme, levels associated with changes in the outermost atomic electrons are converted into broad bands of allowed energies. Thus the higher lines (short-wave lengths) in the series of X-ray lines, as they are observed when emitted by free atoms, become replaced in the solid by broad emission bands, which may overlap. These bands are emitted when valence electrons associated with the solid as a whole drop into vacancies produced by cathode-ray bombardment in the interior of atoms. The shape of such emission bands will be characteristic of the structure of the solid and may furnish information in regard to that structure.

Lack of space prevents an adequate treatment of this interesting development in X-ray spectra, but, to cite one conclusion, the Fermi-Dirac energy distribution of the valence electrons in a metal (Sec. 53) seems to be confirmed by studies of the shape of the X-ray bands that are emitted in the "valence L" electronic transitions for aluminum.¹

¹ CADDY and TOMBOULIAN, *Phys. Rev.*, vol. 59, p. 381 (1941), and references there given.

CHAPTER XI

THE NUCLEUS

In Chap. VI, we referred to Rutherford's experiments on the scattering of α particles and to the hypothesis of the nuclear type of atom which these experiments suggested. This was the first evidence for the existence of nuclei within atoms. The subsequent developments of quantum theory and its use in explaining the origin of spectral lines confirmed Rutherford's hypothesis and gave some information concerning the mass and the charge of the nucleus. Practically the entire mass of the atom is contained in the nucleus, and its charge is equal to Ze , where Z is the atomic number of the atom and e is the numerical value of the charge on an electron. The data furnished by ordinary spectroscopy and by chemistry yielded little additional information. Subsequently, however, new lines of attack were developed, and at the present time our knowledge of the nucleus is well advanced. Progress has resulted from work in several fields: (1) radioactivity, natural and artificial; (2) the precision measurements of the masses of atoms by means of the mass spectrograph; (3) artificial transformations or transmutations of nuclei by bombardment with (a) particles from radioactive substances, or (b) high-speed protons or other charged particles produced by laboratory methods, or (c) neutrons, or (d) γ -rays; (4) spectroscopic investigations in the visible and the ultraviolet regions, which yield evidence regarding (a) the angular momentum or "spin" of the nucleus and its magnetic moment, (b) the relative masses of certain isotopes and, in a few cases, (c) the existence of certain isotopes, the presence of which had escaped detection by the mass spectrograph; (5) the direct measurement of nuclear magnetic moments by means of the molecular-ray methods devised by Rabi (Sec. 150). In this chapter, we shall give a brief survey of some of these fields and of the information obtained concerning the structure and the properties of atomic nuclei.

THE MASSES OF ATOMS

193. Positive Rays.—Measurements of the masses of atoms are made by observing the deflection produced on positively charged ions of the substance under study by the combined action of an electric and a magnetic field, the methods used being similar to that by which J. J.

slit S_2 into the "velocity selector" between S_2 and S_3 , in which there are crossed electric and magnetic fields, both perpendicular to the line joining S_2 and S_3 . The electric field is produced by maintaining the plates P_1 and P_2 at a fixed potential difference; the magnetic field, by an electromagnet, not shown. If F is the electric field and H the magnetic field, then only those ions will pass through S_3 which possess a velocity v given by

$$FE = HEv$$

or

$$v = \frac{F}{H}$$

where E is the charge on the ion. All other ions will be bent from the rectilinear path and will be lost from the beam. After passing S_3 , the ions enter a uniform magnetic field at right angles to the plane of the paper. In this field they travel in circles of radius R such that

$$HEv = \frac{Mv^2}{R}$$

or

$$R = \frac{v}{H} \frac{M}{E} = \frac{F}{H^2} \frac{M}{E}$$

After traversing a semicircumference, the ions fall upon a photographic plate, leaving traces (lines), the position of which depends on the ratios M/E . It is readily seen that, for ions having a given charge E , M is proportional to R , and hence the mass scale is linear. Figure 184 shows the "mass spectrum" of germanium taken by Dr. Bainbridge.

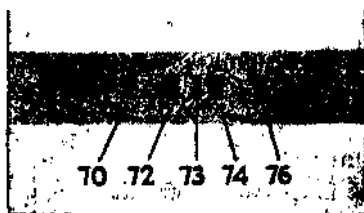


FIG. 184.—The mass spectrum of $\text{Ge}(32)$, showing the isotopes 70, 72, 73, 74, 76. (From a print kindly furnished by Dr. Bainbridge.)

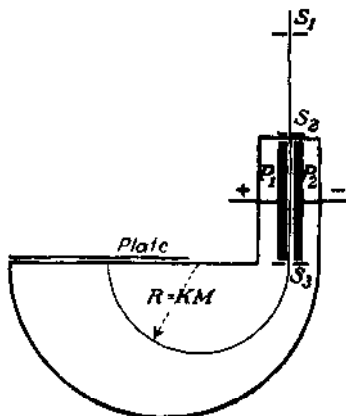


FIG. 183.—Bainbridge's mass spectrograph.

196. Isotopes, (a) Isotopes of the Elements.—All of the 92 familiar elements have now been analyzed by means of the mass spectrograph. As

has been implied in previous statements, the atomic weights of the individual isotopes are commonly expressed on a scale on which the *neujxgl* atom of the most abundant isotope of oxygen has a weight exactly

equal to 16; such weights are called *isotopic weights*, or *isotopic masses*. They are found to be within 0.1 of an integer. This integer is called the *mass number* of the isotope in question, often designated by *A*. Atoms (or nuclei) having the same mass number *A* are called *isobars* of each other; if *Z* is likewise the same, they are also called *isomers*.

To indicate a particular isotope of an element, the mass number is commonly added to the chemical symbol as a superscript; the atomic number may also be added as a subscript, preferably preceding the symbol, thus:



Confusion is thus avoided with the ordinary use of the subscript in compounds; *e.g.*, the deuterium molecule is ${}_1\text{H}_2^2$, whereas ${}_1\text{H}^1 {}_1\text{H}^2$ is a molecule containing 1 atom of ordinary hydrogen and 1 of deuterium (or heavy hydrogen, mass number 2).

The mass unit on the $\text{O}^{16} = 16$ scale, or one-sixteenth of the mass of an O^{16} atom (including its 8 electrons), may be called the *atomic mass unity* abbreviated as a.m.u. Its equivalent in grams is with sufficient approximation equal to the reciprocal of Avogadro's number (Sec. 42),¹ hence,

$$1 \text{ a.m.u.} = (6.023 \times 10^{23})^{-1} = 1.660 \times 10^{-24} \text{ gram.} \quad (315)$$

The isotopic weight or mass of an atom represents its mass expressed in atomic mass units.

A table of the known isotopes that occur naturally is given in Appendix I. Several of them were discovered spectroscopically (*e.g.*, H^2 , C^{13} , N^{16} , O^{17} , O^{18}). The relative abundance of the isotopic constituents of each element is shown in column 5. This can be determined either by comparing areas under such curves as those in Fig. 182 (allowing for any difference of scale) or from the intensities of the "line" produced on the photographic plate in the mass spectrograph.² Thus chlorine has been found to contain 75.4 percent of Cl^{35} and 24.6 percent of Cl^{37} ; silver is 51.9 percent Ag^{107} and 48.1 percent Ag^{109} . Many isotopes are normally present only in very small quantities. A study of calcium with the mass spectrograph revealed the presence, in addition to masses 40, 42, 43, 44, and 48, of an isotope with *A* = 46, forming about 0.0033 percent of ordinary calcium.³

¹ Strictly speaking, (315) gives the weight of an atom of atomic weight unity [cf. Eq. (316) below], but there is no difference to the number of figures shown.

² For method, see ASTON, *Roy. Soc. Proc.*, vol. 126, p. 511 (1930).

³ NIER, *Phys. Rev.*, vol. 53, p. 282 (1938).

The atomic masses as given refer to the neutral atom and include, therefore, the masses of the circumnuclear electrons.

A glance at the table in Appendix I reveals a number of intriguing features in the distribution of isotopes as a function of atomic number. For some 17 of the 92 elements, only a single isotope is known (e.g., Be, F, Na, Al, P, I, Cs, Au, Bi); and all but one of these elements (Be) have *odd atomic numbers*. For other elements, especially those of *even* atomic number, many isotopes are known, e.g., as many as 10 for tin ($Z = 50$). Furthermore, isotopes of odd mass number tend to be both less numerous and less abundant in nature. Reasons for such features in the distribution of isotopes should be furnished by an adequate theory of nuclear structure. We shall return to this point later (Sec. 209).

For future use, the values of several other masses in terms of atomic mass units may be noted at this point:

Electronic mass $m = 0.00055$ a.m.u.;

Mass of proton = 1.00758 a.m.u.;

Mass of neutron = 1.00893 a.m.u.

The electronic mass is found by dividing the mass of a hydrogen atom or 1.00813 by 1,837 (Sec. 97).

The mass of the proton is the mass of the hydrogen atom less the mass of 1 electron. The mass of the neutron is quoted here from Sec. 205.

(b) *Isotonic Constitution and Chemical "Atomic Weight."*—The atomic weight employed by the chemist obviously represents the average relative weight of the atoms in the mixture of isotopes with which he ordinarily works, or the *average isotonic weight* of this mixture. The *atomic weight of oxygen* is arbitrarily taken to be 16.

Oxygen, however, long thought to contain only the atom O^{16} , was later found to contain also minute traces of O^{17} and O^{18} ; the masses of the latter, referred to O^{16} as 16 exactly, are, respectively, 17.0045 and 18.005. The average mass of the atoms in the normal oxygen mixture on the $O^{16} = 16$ scale is about 16.0044, as may be determined from the data given for oxygen in Appendix I. Thus, the atomic weight M_c for any isotopic mixture is related to M_i the average isotopic weight of the atoms in the mixture, by the equation

$$M_c = \frac{16.0000}{16.0044} M_i = \left(1 - \frac{1}{3,600}\right) M_i. \quad (316)$$

As an example, ordinary lithium has been found to consist of 7.5 percent of Li^6 , with an isotopic weight $M_i = 6.0169$, and 92.5 percent

of Li^7 , for which $M_i = 7.0180$; from these data, its average isotopic weight is readily found to be $M_i = 6.943$. The chemical "atomic weight" will be a little smaller, viz., $M_c = 6.941$. This agrees excellently with the value of 6.940 measured chemically.

In this way the chemical atomic weights of a number of the elements have been calculated from the measured relative abundance and masses of their respective isotopes.¹ The results are shown in column 6 of Appendix I. For comparison, the directly observed chemical atomic weights are shown in column 7. The agreement is seen to be excellent.

(c) *Separation of Isotopes*.—Chemically, the isotopes of a given element are almost or quite indistinguishable, since chemical properties depend primarily upon the nuclear charge and are little affected by the nuclear mass. The same statement must be true, in general, of physical properties, for these are mostly determined by the force fields surrounding the molecules, which depend in turn upon the number of electrons in the molecule and the electrical charges on the nuclei but are little influenced by the nuclear mass. One exception, of course, is the density, which must be nearly proportional to the isotopic weight. Any method for the separation of isotopes, therefore, must be based either directly on the difference in nuclear mass or on the utilization of very slight differences in other molecular properties.

The mass spectrograph constitutes the most obvious means for separating the isotopes of the elements. By replacing the photographic plate *PP* in Aston's apparatus (Fig. 180) by suitably disposed slits, we should be able to collect one kind of neon atoms, say, in one compartment and the other kind in another. Or, many kinds of atoms might simply be collected as deposits on a metal plate. In this way fairly pure samples of Li^6 and Li^7 have actually been obtained, and, during the years 1945-46, an improved form of the method was employed, by dint of great effort, to separate U^{235} from U^{238} in considerable quantities.²

Several other methods of separating isotopes have been employed with some success. In the case of hydrogen, where one isotope is twice as heavy as the other, electrolytic separation happens to be very efficient, and this method is now employed on a large scale in

¹ See ASTON: *Roy. Soc, Proc*, vol. 126, p. 511 (1930); vol. 130, p. 302 (1931); vol. 132, p. 487 (1931); HAHN, FLUGGE, and MATTAUCH, *Phys. Zeits.*, vol. 41, p. 1 (1940).

² SMYTHB, H. D., "A General Account of the Development of Methods of Using Atomic Energy for Military Purposes under the Auspices of the United States Government, 1940-1945," Princeton University, 1945.

the production of heavy hydrogen, H^2 . The isotope H^1 is evolved, in proportion to the amount present, five times¹ as fast as H^2 ; thus, if a large volume of water is almost completely electrolyzed, the small residue remaining is almost pure heavy water, H_2^2O .

Another method, applicable to any gaseous element, is that of diffusion through a porous barrier composed of unglazed baked clay or similar material. The lighter molecules diffuse through the barrier faster in inverse proportion to the square root of the mass. By operating many diffusion stages suitably connected together, G. Hertz and others have been able to effect almost complete separation of neon into its isotopes. With use of a light gas containing carbon, this method can be made to yield several milligrams of the rare isotope C^{13} per day.² Another method, known as "thermal diffusion," makes use of a slight separation that occurs where heat is being conducted through a gas or liquid and is rather more useful than the method of simple diffusion. Finally, many isotopes can be separated effectively by means of "chemical exchange," a method in which advantage is taken of very slight differences in chemical properties and many successive fractionations are made. By this method it is possible to obtain, for example, a gram per day of the isotopes N^{15} , C^{13} , O^{18} , and S^{34} .

THE DISCOVERY OF NUCLEAR TRANSFORMATIONS

We have seen that the mass spectrograph yields very precise data concerning the masses of atoms and their nuclei. The fact that the masses of atoms are very nearly proportional to whole numbers raises two questions: whether the various nuclei may not be built up out of one or more common constituents; and whether or not it may be possible to transmute one nucleus into another. The first evidence suggesting an answer to these questions came from the field of radioactivity.

196. The Discovery of Radioactivity.—The discovery of the phenomenon of radioactivity, although quite accidental, resulted directly from the discovery of X-rays. Roentgen had shown that X-rays are emitted by those parts of the discharge tube which are bombarded by the cathode rays. This bombardment was also accompanied by the emission of the well-known greenish or bluish fluorescence. The question arose: Is fluorescence always accompanied by the emission of X-rays?

¹ LEWIS and MACDONALD, *J. Chem. Phys.*, vol. 1, p. 34 (1933).

² URKY, "Separation and Use of Stable Isotopes," *J. of Applied Phys.*, vol. 12, p. 270 (1941).

Several investigators had apparently found that fluorescent bodies activated by sunlight gave out a type of radiation which, like X-rays, was able to pass through black paper and to affect a photographic plate. In February, 1896, a few months after the discovery of X-rays, Henri Becquerel¹ was trying an experiment of this kind using as the fluorescing substance the double sulphate of uranium and potassium. After preparing the experiment and while waiting several days for sunshine, Becquerel discovered that even in the dark the specimen emitted a radiation which penetrated not only black paper but even thin sheets of metal, and that exposure of the fluorescing substance to sunlight had no effect on the phenomenon.

Becquerel soon found that this radiation was emitted by uranium irrespective of its state of chemical combination and that there was no connection whatever between this phenomenon and phosphorescence. Furthermore, the phenomenon was found to be quite independent of the temperature of the uranium compound. It was later discovered that these rays from uranium possess the power of discharging electroscopes by rendering the air through which they pass conducting.

This property of "radioactivity," as it was called, was soon found to be possessed by several other substances. Among them were thorium and two new elements, polonium and radium, discovered by M. and Mme. Curie, radium being more than a million times more active than uranium.

197. The Radioactive Radiations.—Like X-rays, the "rays" from radioactive materials affect a photographic plate, cause fluorescence, and ionize gases through which they pass. Unlike X-rays, however, these rays are of three types, which were given the names, before their nature was certainly known, of α -, β -, and γ -rays. If a small quantity of radium preparation is placed at the bottom of a small hole drilled into a lead block *B*, the emerging rays can be divided into the three groups by use of a strong magnetic field at right angles to the plane of the paper and directed away from the reader, as is shown in Fig. 185. One group is bent into a circular path to the right and will cause an impression on a photographic plate *PP*. These are the β -rays. From the direction of their deflection, it follows that they must be negatively charged particles. By studying quantitatively their deflection in magnetic and electric fields, it was shown that these particles are electrons; they are ejected from radioactive materials with velocities which in some cases are very high. A second type of radiation is deflected slightly toward the left. This type consists of positively charged particles called "*a* particles," which were shown to possess

¹ BECQUEREL'S papers appear in *Comptes Rendus*, vol. 122 (1896).

a ratio E/M of the order of magnitude of that for *atoms*. These *a* particles were found to have a mass 4 (taking the mass of the oxygen atom as 16) and to carry a charge $+2e$. This identified them with the nuclei of helium atoms. The third type of radiation, the γ -rays, proceeds undeviated by either electric or magnetic fields, has a very high penetrating power, and is now known to consist of electromagnetic radiations of very short wave length lying, in general, in the spectral region beyond the shortest X-rays.

Recently it has been found that many artificially produced radioactive materials emit *positrons* instead of (negative) electrons, a positron being the same as an electron in all respects except that its electrical charge is positive (see Sec. 212). The tendency of usage seems to be to apply the term " β -rays" to both electrons and positrons when emitted by radioactive substances. In Fig. 185, the path of a positron β -ray would be a circle like that shown but curving toward the left instead of toward the right.

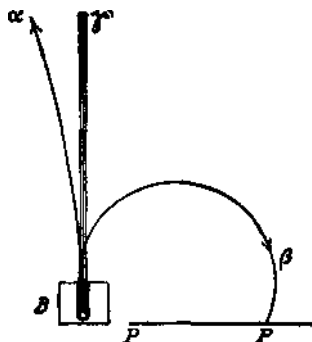


FIG. 185.—Schematic representation of the three types of rays from naturally radioactive materials.

The three types of rays are further differentiated from each other by their penetrating power. The α -rays are absorbed by a few centimeters of air at ordinary pressure. They are reduced in intensity one half by passing through 0.005 mm. of aluminum. Their initial velocities are of the order of 2×10^9 cm.-sec.¹ The β -rays are, roughly, 100 times more penetrating, since it requires something like 0.5 mm. of aluminum to reduce their intensity to half. The initial velocities of the β -rays, in some instances, exceed 99 percent of the velocity of light. The γ -rays are able to penetrate many centimeters of even so dense a metal as lead.

A short additional discussion of the properties of these radiations will be given in the next few sections. For further details, the reader is referred to special treatises¹ or to the original articles.

198. The Alpha-rays, (a) *General Properties*.—The velocity and the value of E/M for α -rays were determined by Rutherford by means

¹ RUTHERFORD, CHADTICK, and ELLIS, "Radiations from Radioactive Substances," 1930; SODDY, "The Chemistry of the Radio Elements," 1914; HEVESY and PANETH, "Radioactivity," translated by R. W. Lawson, 1936; F. RASETTI, "Elements of Nuclear Physics," 1936; J. M. CORK, "Radioactivity and Nuclear Physics," 1946.

of a modification¹ of the usual method in which magnetic and electric fields are employed. The ratio of charge to mass, E/M , was found to be almost exactly half as large for the α particles as it is for atoms of hydrogen. Additional observations were then necessary in order to determine E and M separately.

By means of the apparatus shown in Fig. 186, Rutherford and Geiger counted the number of α particles emitted per second by a given quantity of radium C.² The radioactive material was deposited on a disk D placed inside the highly evacuated vessel A at a known distance from a small circular opening S of known area, the opening being covered by a sheet of mica thin enough to allow the passage of the α particles into a brass chamber C . This chamber was evacuated to a pressure of several millimeters of mercury and had at its center

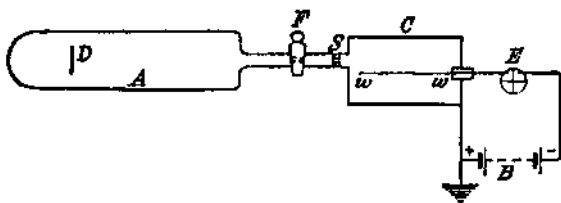


FIG. 180.—Apparatus of Rutherford and Geiger for "counting" α particles.

an insulated wire WW , which, by means of a battery B , was maintained at a potential, with respect to the walls of the cylinder, just less than the critical discharge potential. When an α particle entered C , the ionization caused by its passage through the gas in C lowered the critical potential by an amount sufficient to allow the passage through the cylinder of a momentary current, which could be detected by a "kick" in the electrometer E . In this way, Rutherford and Geiger were able to detect the passage of *single* α particles. Thus, knowing the rate at which the α particles passed through S , the aperture which S subtended at D , and the quantity of radium on D , it was possible to determine the total number of α particles emitted per second per gram of radium C.

The total charge carried by the particles was then determined with the apparatus³ shown schematically in Fig. 187. Through a window W , α particles from a known quantity of radium C deposited on the plate P were allowed to fall upon a collecting plate C ; the charge that they imparted to C was measured with an electrometer. To complete

¹ Cf. RUTHERFORD, CHADWICK, and ELLIS, *loc. cit.*

² RUTHERFORD and GEIGER, *Roy. Soc. Proc.*, vol. 81, p. 141 (1908).

³ *Ibid.*, D. 162.

the electrical screen around C, the window was covered with aluminum foil thin enough to allow all of the α particles to pass through it. The entire vessel was evacuated.

Division of the observed total charge by the known number of the α particles then gave the charge on each one. This turned out to be twice as great as the electronic charge or the charge on a hydrogen ion. From this result and the previously found value of E/M , the mass of an α particle was found to be approximately four times that of a hydrogen atom. Thus the α particle must be the nucleus of a helium atom. This conclusion was confirmed, both qualitatively and quantitatively, by the observation that *helium is produced* as a result of α -ray activity. Rutherford's observations led to the conclusion that 1 gram of radium in radioactive equilibrium would emit 1.3×10^{11} α particles per second, and that these, after becoming neutralized by picking up electrons, would, in 1 year, form 0.16 cc. of helium gas.

The *velocity* of the α particles from radium C, determined as above by Rutherford, is 2.06×10^9 cm. per sec. Since their mass is

$$\frac{4}{N_0} = \frac{4}{6.023 \times 10^{23}} = 6.64 \times 10^{-24} \text{ gram,}$$

we find for their initial energy

$$\frac{1}{2}mv^2 = 14.1 \times 10^{-6} \text{ erg,}$$

or

$$\frac{14.1 \times 10^{-6}}{1.601 \times 10^{-12}} = 8.8 \times 10^6 \text{ electron-volts.}$$

More recent measurements give an initial velocity of 1.92×10^9 cm. per sec. and an initial energy of 7.68×10^6 electron-volts. It has been found that the α particles from a given substance in some cases all have the same initial energy, whereas in other cases they fall into several sets, the particles in each set having the same energy. In the latter case, when the α particles are separated by means of a magnetic field, a "line spectrum" is obtained.¹

In Table I below are shown initial velocities and energies for the α -rays from a number of typical radioactive substances.

(b) *The Range of Alpha Particles.*—Because of its double charge and great mass, an α particle produces an enormous number of ions as

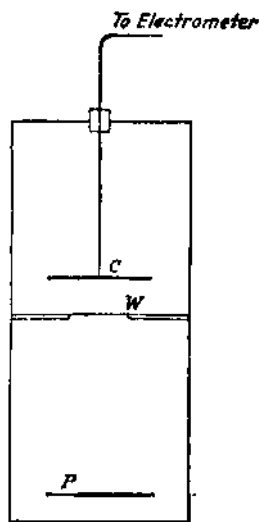


FIG. 187.—Rutherford's apparatus for measuring the charge carried by the α -rays from a known quantity of radioactive material.

¹ Cf. RASETTI, *op. cit.*, p. 114.

it passes through matter. Since energy is lost in producing these ions, and also in exciting many molecules without ionizing them, the α particle rapidly loses velocity, until finally it is moving too slowly to produce any ions at all. The total length of path along which an α particle causes ionization is called its *range*. Since the range increases

with increasing initial energy, it is commonly used as a measure of the initial energy of the α particle.

Various methods have been devised to measure the range of α particles. An approximate method makes use of the scintillations produced when α particles strike a screen of phosphorescent zinc sulfide, the range of the particles being measured by the maximum distance between source and screen, in air under standard conditions, at which scintillations can be observed. A more precise method is illustrated

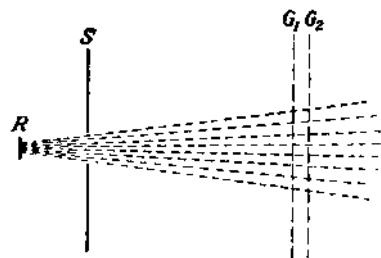


FIG. 188.—Arrangement for measuring the range of α particles from a source R by the ionization which they produce between two pieces of wire gauze, G_1G_2 .

diagrammatically in Fig. 188. The α particles from a radioactive source R pass through the opening in the screen S . The ionization produced in the space between two strips of wire gauze, G_1G_2 , 1 mm. or so apart, is measured for various distances between G_1G_2 and R

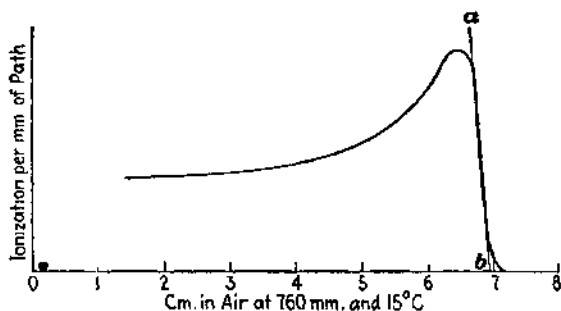


FIG. 189.—Ionization produced by an α particle from radium C at different distances from the source. The ordinate represents about 25,000 pairs of ions per centimeter at the left end of the curve, or, at the maximum, about 60,000, in air of standard density.

(by varying either distance or gas pressure). Curves are obtained similar to that shown in Fig. 189 for radium C. The general shape of the curve may be accounted for as follows. Since the impulses given to the component parts of an atomic system by the passage near or through it of an α particle depend upon both the magnitude of the forces and the time during which those forces act, and since the speed

of the particle is greatest at the beginning of its path, the smaller ionization over the first few centimeters is explained. As the particle loses energy, and therefore velocity, owing to the formation of ions, the times during which the forces act, and therefore the impulses, increase for given distances of passage between a particle and atom, and the ionization per centimeter increases to a maximum; beyond this it falls rapidly. Since the exact end of the ionization curve is difficult to determine, it became customary to take the point b , where the straight line ab intersects the axis, as a measure of the range. A range so determined is called an *extrapolated range*.

The ordinate of the curve in Fig. 189 represents about 25,000 ion pairs per cm. in standard air at the left-hand end of the curve, and about 60,000 at the maximum of the curve. An ion pair consists of the ionized molecule and the electron ejected from it. The ionization curve for an α particle of lower initial energy is practically the same as a part of the curve shown in Fig. 189 taken from the proper point outward. For example, if the particles have such an initial energy that their extrapolated range is 4.9 cm., instead of 6.9 cm. as in Fig. 189, the curve will start at the point corresponding to $6.9 - 4.9 = 2.0$ cm. in the same figure.

Such curves are affected, however, by *straggling* of the particles; even if all of them have the same energy initially, their individual ranges differ slightly because of statistical fluctuations in the number and kinds of ions or excited molecules that they produce. By modifying the arrangement shown schematically in Fig. 188, so as to *count the number* of particles that cross the space G_1G_2 , instead of measuring the total ionization produced by them, the distribution of the individual ranges can be found. In the observations of Holloway and Livingston,¹ the counting was done by connecting G_1 or G_2 through an amplifier circuit to a thyratron, which, when suitably adjusted, has the property that voltage impulses below a certain magnitude have no effect upon it, whereas those exceeding this critical voltage cause a large momentary rush of current through the thyratron;² this current was made to activate a mechanical counter by means of an electromagnet.³

Plotting the number of particles as thus observed against the distance x from R to G_1 (Fig. 188), we obtain a curve in which the ordinate represents the number of α particles that have traveled a

¹ HOLLOWAY and LIVINGSTON, *Phys. Rev.*, vol. 54, p. 18 (1938).

² See REICH, H. J., "Electron Tubes and Circuits," 1940; Dow, W. G., "Fundamentals of Engineering Electronics," 1937.

³ See also COLBY and HATFIELD, *Rev. Set. Instruments*, vol. 12, p. 62 (1941).

distance equal to x without stopping. The slope of this curve then gives the number that stop per unit of x . In this way the true *mean range* of α particles of a given initial energy can be found. Doubtless in the future it will become customary to state as *the range* of the particles the mean range rather than the extrapolated range.

In experiments where it is not convenient to measure the range in air, it can be determined in metal foils whose "stopping power" in equivalent centimeters of air has been determined empirically. The range of the α particles from radium C in aluminum, copper, and lead is, in units of 10^{-3} cm., 4.0G, 1.83, and 2.41, respectively. The "mass stopping power" of various elementary substances, or reciprocal of the range divided by the density, for α -rays, is nearly inversely proportional to the square root of the atomic weight. The (extrapolated) ranges in air of the α particles from a number of typical radioactive substances are given in Table I. Curves of range versus energy may be found elsewhere.¹

TABLE I.—RANGE, INITIAL VELOCITY, AND ENERGY OF ALPHA PARTICLES FROM TYPICAL RADIOACTIVE SUBSTANCES*

Emitter	Range, cm , in air at 760 mm. and 15°C.	Initial velocity, cm. sec. ⁻¹	Initial energy, ev
Uranium I	2.73	1.40×10^9	4.05×10^6
Radium	3.39	1 51	4 74
Radium C .	6 97	1 92	7 68
Radium F (polonium).	3 93	1 59	5 25
Thorium	2 90	1 44	4 23
Thorium C'	8 62	2 06	8 76

*Compiled from a similar table given by Rutherford, Chadwick, and Ellis.

The increase in initial energy of the α particle that is required to lengthen its range by 1 cm. represents the loss of energy by the particle in going 1 cm. If we divide this loss of energy by the number of ion pairs produced in 1 cm., we have the average loss of energy per ion pair. This loss appears to be about 35 volts per ion pair, in air. Part of the energy, however, is undoubtedly lost in exciting molecules without ionizing them.

It may be remarked that *protons* (or hydrogen nuclei) produce about $\frac{1}{4}$ as many ions and lose about $\frac{1}{4}$ as much energy per centimeter as do α particles of the same velocity, owing to the fact that the ioniza-

¹LIVINGSTON and BETHE, *Rev. Modern Phys.*, vol. 9, p. 245 (1937), plot on p. 266; HACMAN and HAXEL, *Zeits.f. Physik.*, vol. 120, p. 486 (1942).

tion produced by a charged particle is roughly proportional to the square of its charge. Since the mass of the proton is likewise 1836 as great, its initial energy is also 1836 as great as that of an α particle moving with the same velocity. This means that protons and α particles of the same velocity have about the same range. Or, again, an α particle has four times the energy of a proton of the same range. If, however, we compare protons and α particles having the *same energy*, we find that the protons have ranges 5 to 10 times as great as the α particles. A *deuteron*, or nucleus of heavy hydrogen (H^2), has twice as much energy as a proton of the same range.

199. Radioactive Transformations.—The early study of radioactivity led to the conclusion that the emission of α and β particles was associated with a chemical change in the radioactive atom. The transformation was found to follow an exponential law. During each successive element of time, a certain definite fraction of the surviving atoms of the original substance "decay" or disintegrate, so that the total number that survive after any given time t decreases in proportion to $e^{-\lambda t}$, where λ is a constant, the decay constant. Thus, during a time $T = (\log 2)/\lambda$, half of the atoms disintegrate. The time T is called the *period*, or *half-wave period*, or *half-life*, of the radioactive substance. Instead of T , a time equal to $1/\lambda$ is sometimes cited; it represents the *mean life* or the average time during which an atom of the radioactive substance exists before disintegrating. [Cf. analogous relations in the emission of radiation, Sec. 116(a).]

The half-value period of uranium, for example, is estimated at 4.5×10^9 years, which is presumably long enough to account for the considerable quantities of uranium that exist in the earth's crust. In disintegrating, uranium gives rise, indirectly, to radium, with a period of 1,600 years; this period is so short, relatively, that radium is found only in uranium-bearing rocks, where it has been produced within the last few thousand years. Radium emits α particles and thereby becomes converted into an inert gas known as radon ($Z = 86$), or radium emanation; this gas also emits α particles, but it has a period of only 3.825 days.

If the new atoms formed as the result of radioactive change are themselves noticeably radioactive, many of their chemical properties, and the boiling point of the new element composed of them, can often be determined even if their mean life is so short that the new element cannot be accumulated in visible quantity. A radioactive substance can be followed in chemical reactions by means of its radioactivity, so that the chemist can tell, for example, whether it remains in solution or is precipitated or passes off as a gas. Usually a chemically similar

substance, called a "carrier," is added, so as to have something visible to work with; it is then easy to tell whether or not the radioactive substance goes with the carrier in a chemical operation.

The nuclear theory of the atom made it possible to propose a precise theory of radioactive transformations which is in agreement with all of the known facts. When an atom, for example, loses an α particle, its mass number must decrease by 4, and its atomic number by 2, since the particle carried a double positive charge out of the nucleus. The new substance formed will thus be an isotope of an element standing two columns to the left of the original substance in the periodic table. It may be one of the stable isotopes of that element which occurs in nature, or it may be a new isotope, stable or unstable.¹ That just one α particle is given out in such processes for each atom that is transformed has been shown by counting the number of α particles given out by a known amount of a radioactive substance in a given time and comparing the number so obtained with the known rate of disintegration of the material.

The ejection of a negative β particle from a nucleus, on the other hand, *without* changing its *mass number*, *increases* its positive charge and thereby *raises* its *atomic number* by unity. The emission of a γ -ray can obviously change neither the mass number nor the atomic number. It appears that in the natural radioactive substances γ -rays are not associated with an independent type of radioactive change but constitute an incidental accompaniment of transformations involving the emission of an α or β particle.

A systematic study of radioactive substances and their disintegration products has led to the discovery that the naturally radioactive substances can be grouped into several *radioactive series*, the elements of a given series being so arranged that each is a disintegration product of the preceding element. Thus, the uranium atom of mass number 238 ($Z = 92$), called "uranium I" or U I, may emit an α particle and thereby become converted into a nucleus with atomic number $Z = 90$ and mass number $A = 234$. The new substance so produced, called "uranium X_1 " (UX_1), is found by chemical tests to be an isotope of ordinary thorium ($Z = 90$, $A = 232$). The UX_1 atom then emits a β particle (period, 24.5 days) and becomes "uranium X_2 " (UX_2 ; $Z = 91$, $A = 234$); this emits another β particle (period, 1.14 minutes) and becomes uranium II (U II; $Z = 92$, $A = 234$). Thus U II is an isotope of ordinary uranium and is chemically indistinguish-

¹ The transformed atom will promptly lose 2 of its circumnuclear electrons as well, thus bringing the number of its electrons into harmony with its reduced nuclear charge.

able from it; but its mass number is 234, not 238, and its period is much shorter than that of U^{234} .

Uranium II also emits α particles (period, about 10^6 years), forming ionium, with $Z = 90$, $A = 230$. Ionium then emits another α particle (period, 7.6×10^4 years) and becomes radium ($Z = 88$, $A = 226$). The atomic weight of radium, determined chemically, is 225.97. The end product of this series, after the emission of five more α parti-

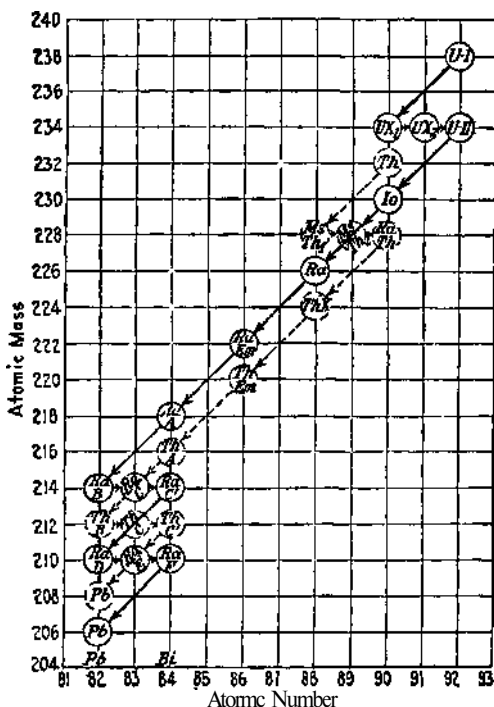


FIG. 190.—The radioactive-disintegration series of uranium and of thorium.

cles and four more β particles, is radium G or radium lead ($Z = 82$, $A = 206$), which is indistinguishable chemically from ordinary lead and has a measured atomic weight of 206.05. The entire uranium series is shown in Table II, and also, along with a similar series starting from thorium, in Fig. 190. In this figure, the emission of an α particle is indicated by an arrow pointing obliquely downward to the left, the emission of a β particle, by a short horizontal arrow pointing to the right. (Two small branches at UX_1 and at radium C are omitted.¹) Two radioactive members of the uranium series, Ra C and Ra E, are chemically indistinguishable from bismuth ($Z = 83$, $A = 209$).

¹ Cf. RUTHERFORD, CHADWICK, and ELLIS, *op. cit.*, Chap. I.

TABLE II.—THE UKANILM SERIES OF RADIOACTIVE SUBSTANCES

Nucleus.	UI	UX ₁	UX ₂	UII	Io	Ra	Rn
Z	92	90	91	92	90	88	86
A . . .	238	234	234	234	230	226	222
Ray	α		β	α	α		α
Period..	4.5×10^9 years	24.5 days	1.14 min.	10_6 years	7.6×10^4 years		1,600 years

Nucleus	Rn	RaA	RaB	RaC	RaC	RaD
Z	86	84	82	83	84	82
A	222	218	214	214	214	210
Ray Period	α 3 825 days	α 3 05 nun.	β 26 8 mm.	10 7 mm.	(10 ⁻⁶ sec.)	

Nucleus	RaD	RaE	RaF	RaG (Pb)
Z	82	83	84	82
A	210	210	210	206
Ray.. Period	β (25 years)	β 5 0 days	α 136 3 days	

(U — uranium, Io = ionium, Ra = radium, Rn = radon, RaF = polonium.)

The end products of both the uranium and the thorium series are isotopes of lead, with $A = 206$ and 208 , respectively. Ordinary lead has the atomic weight 207.2; analysis by the mass spectrograph reveals the presence of isotopes 206, 207, and 208, with a small amount of 204. Observed values of the atomic weight of lead, however, vary somewhat with the source. Uranium-bearing minerals usually contain lead—a fact in itself confirming the series of radioactive transformations from uranium to lead. The same is true of certain thorium-bearing minerals. The atomic weight of lead coming from uranium-bearing minerals is only a little over 206, whereas lead associated with Norway thorite has an atomic weight of 207.9, which is nearly the value 208 predicted for the end products of the thorium series.

In an early stage of the work on radioactivity it was decided to adopt radon as a standard radioactive substance. The amount of radon existing in a gram of radium in equilibrium with its products was called a "curie"; it is about 0.66 mm.⁸ of radon at normal pressure

and temperature. Other activities are also frequently expressed in curies, although not without awkwardness; for this purpose a curie is understood to be an activity in which the same number of disintegrations occur in a second as in a curie of radon, or about 3.7×10^{10} per second.

200. Gamma-ray Spectra.—Direct measurements of the wave lengths of some of the γ -rays have been made by use of the crystal grating. The wave length is so small, however, in comparison with the grating constants of available crystals that only the longer wave lengths can be measured in this way. Consequently, an indirect method has been developed based on the magnetic spectrum of the secondary β -rays which are produced by photoelectric absorption of the

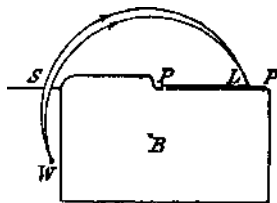


FIG. 191.—The β -ray spectrograph of Ellis.

γ -rays in matter.

The method employed by Ellis¹ has been followed, with modification, by subsequent investigators. In Ellis's experiments the source of the secondary rays was placed at *W* (Fig. 191) near a lead block *B*. *W* was a small glass tube 1 cm. long and 0.7 mm. in diameter, the inside walls of which were coated with a deposit of radium *B* which served as a source of the γ -rays under study. Around *W* was wrapped a foil of the metal *M*—Pb, Pt, W, etc.—which was to serve as the source of the secondary β -rays. The γ -rays passing out through the walls of the tube excited secondary β -rays in the metal foil. The whole apparatus was placed in a magnetic field (and, of course, in a highly evacuated enclosure). The electrons passed, in circular paths, through the wide slit *S* and were "focused" onto the photographic plate *PP* at *L*. Upon developing the plate, a number of "lines" were found corresponding to various electron energies which could be determined by knowing the magnetic field and the diameter *LW* of the circular paths.

The data in Table III, taken from Ellis's paper, will serve to illustrate the method of analyzing the results. For any given metal *M*, there was observed a number of lines on the plate of which three were more intense than the others. These same three appeared on the plate for each element but were shifted in position in such a direction as to indicate that the corresponding electrons had *less* energy the *greater* the atomic number of *M*. This is shown in the upper part of Table III, in which the energies of the electrons are expressed in electron-volts (Sec. 96). Ellis interpreted these elec-

¹ ELLIS, *Roy. Soc. Proc.*, vol. 99, p. 261 (1921).

trons as *photoelectrons* ejected from certain levels in the atoms of the emitting substance. The change in the energy of the electrons in a given line from one element to another agrees closely with the difference of the energies of the *K* levels of the 2 atoms. Hence, he concluded that the electrons forming the three lines were expelled from the *K* levels of the respective atoms by three different y-ray frequencies. Applying Einstein's photoelectric equation

$$h\nu = E_k + W_K$$

where E_k is the observed kinetic energy of the photoelectron and W_K is the energy required to remove the electron from the *K* level of the atom, one obtains at once the energy $h\nu$ of the incident γ -rays. Data from each metal are found in this way to give the *same* group of three y-rays, as the lower half of Table III shows. These, then, are three *lines* in the y-ray spectrum of radium B, the respective wave lengths of which are 0.0521, 0.0425, and 0.0355 angstrom. In a similar way energy values of many other lines in γ -ray spectra have been measured.

TABLE III.—DETERMINATION OF THE WAVE LENGTH OF GAMMA-RAYS FROM THE ENERGY OF SECONDARY BETA-RAYS (ELLIS)

Emitter	Tung- sten 74	Plati- num 78	Lead 82	Ura- nium 92		
Secondary β -ray energies in elec- tron-volts, $\times 10^{-5}$	1.66 2.20 2.76	1.58 2.12 2.69	1.49 2.03 2.60	1.22 1.74 2.31		
E_K of emitter	0.693	0.782	0.891	1.178	Mean energy	angstroms
Energy ($h\nu$) of y-rays in electron- volts, $\times 10^{-5}$	2.35 2.89 3.46	2.36 2.91 3.46	2.38 2.92 3.49	2.40 2.92 3.48	2.37 2.91 3.47	0.0521 0.0425 0 0355

Photoelectrons from shells other than the *K* shell have also been observed. In addition, there is a continuous background in the spectrum of the secondary β -rays which is due to Compton electrons resulting from *scattering* of the y-rays.

If the interpretation of such observations is correct, it will also be expected that a y-ray may be absorbed photoelectrically by one of the electrons surrounding the nucleus that emits the γ -ray itself. Radioactive substances emitting y-rays of a given wave length ought,

therefore, to emit in addition groups of β -rays forming a spectrum of β -ray lines corresponding to ejection from the K , L , M , $\cdot \cdot \cdot$ shells of the radioactive atoms. Such is, indeed, observed to be the case. The production of *secondary* β -rays in this manner is commonly called *internal conversion* of the γ -ray. It happens much more frequently in the K shell than in the L shell; the probability that the γ -ray is so converted before escaping from the atom varies in different cases from 0.001 or less to 0.1. β -ray lines produced in this manner are much sharper (the velocities of the β -rays being more nearly uniform within each group) than are the lines excited by the same γ -rays in extraneous substances. Hence most wave lengths of γ -ray lines have actually been determined from observations of the energies of the β -rays due to internal conversion.

Not all natural β -rays arise from internal conversion of γ -rays, however. As we shall see in the next section, many of them come from the nucleus itself.

All γ -ray spectra, apparently, consist of sharp lines. Some of them have wave lengths longer than the hardest X-rays obtainable from the same atom, but most of their wave lengths are shorter than the shortest K line from any element (*viz.*, 112 X.U. = 0.112 Å, from uranium). Thus the γ -rays cannot consist of radiation emitted by the circumnuclear electrons; in view of the success of the modern theory of atomic spectra, we can assume with safety that the electrons can emit no harder lines than the K lines. The γ -rays must, therefore, be emitted somehow by the nucleus.

201. Nuclear Energy Levels.—In some cases measurements of γ -ray spectra have yielded sufficient data so that it has been possible to postulate a system of energy levels in the nucleus, transitions between which give rise to γ -ray lines in exactly the same way as X-ray lines originate in transitions between levels in the extranuclear structure of the atom.

The wave lengths and energy values of some of the lines in the γ -ray spectrum of radium B, as found by Ellis and Skinner,¹ together with the "name" of the line, are shown in the first three columns of Table IV. These investigators found it possible to postulate a series of seven levels A , B , C , $\cdot \cdot \cdot$ transitions among which give rise to the 14 observed lines, as indicated in column 4. The resulting energy-level diagram is shown in Fig. 192.

In view of the difficulty of making the measurements, it is a scientific achievement of the first magnitude to have isolated and measured a set of γ -ray lines with sufficient precision for the construc-

¹ ELLIS and SKINNER, *Roy. Soc. Proc.*, vol. 105, pp. 165, 185 (1924).

TABLE IV.—SOME LINES IN THE GAMMA-RAY SPECTRUM OF RADIUM B AND THE ENERGY LEVELS IN THE RADIUM C NUCLEUS (ELLIS AND SKINNER)

Line	Wave length, angstroms	Energy (unit: 10 ⁶ electron- volts)	Origin	Proposed energy levels (unit: 10 ⁵ electron-volts)
<i>C</i> ₂ [*]	1.37	0.090	<i>B C</i>	<i>A</i> = 0
<i>C</i> ₁ [*]	0.230	0.537	<i>A B</i>	<i>B</i> = 0 537
<i>C</i> ₅ [*]	.196	0.629	<i>A C</i>	(7 = 0 625
<i>C</i> ₂ [*]	.115	1.073	<i>E F</i>	<i>D</i> = 2 572
<i>C</i> ₄ [*]	.098	1.250	<i>F G</i>	<i>E</i> = 2 942
<i>E</i> ₁	.0634	1.947	<i>C D</i>	<i>F</i> = 4 048
<i>E</i> ₂	.0607	2.035	<i>B D</i>	<i>G</i> = 5.31
<i>E</i> ₃	.0513	2.404	<i>B E</i>	
<i>E</i> ₄	.0480	2.572	<i>A D</i>	
<i>E</i> ₅	.0451	2.733	<i>D G</i>	
<i>E</i> ₆	.0419	2.942	<i>A E</i>	
<i>E</i> ₉	.0351	3.511	<i>B F</i>	
<i>E</i> ₁₆	.0263	4.684	<i>C G</i>	
<i>E_n</i>	.0257	4.800	<i>B G</i>	

* Measured by Rutheifoid and Andrade by crystal reflection.

tion of an energy-level diagram. Like most scientific advances, however, this accomplishment raises fresh questions. First of all, how do the nuclei come to be in excited states out of which they can

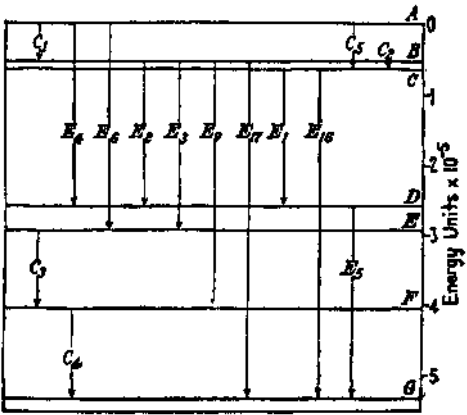


FIG. 192.—Energy-level diagram for the y-rays from radium B. (Ellis and Skinner.)

undergo transitions to lower levels? The situation is quite different from that in vacuum tubes, where atoms are continually being excited by electron impact.

A clue to the answer is furnished by the fact that (in natural radioactivity) γ -rays are emitted only by substances which also

emit α - or β -rays. It is reasonable to suppose that the emission of an α or β particle leaves the residual nucleus, not in its normal state, but in an excited state of higher energy. If this is the correct explanation, then the 7-ray spectrum commonly ascribed to radium B is really emitted by nuclei of the resulting radioactive product, which is radium C; and the energy levels deduced from these rays are levels of the radium C nucleus, not of radium B. This conclusion has been confirmed by experiment, most clearly by Ellis and Wooster,¹ who calculated the difference in the energies of the secondary β -rays produced by a given 7-ray line in the radioactive source itself and in platinum, and compared this difference with the difference in the energies of the K levels in the two substances. Agreement was found with the K"-level difference between platinum and radium C, not with that between platinum and radium B.

It appears that all *natural* 7-rays are emitted by nuclei which have just been formed as a result of α - or β -ray activity. The interval of time between the formation of the new nucleus and the emission of the 7-ray is probably extremely short, of the order of 10^{-12} or 10^{-13} sec. Hence, the 7-ray activity appears to decay at the same rate as the α - or β -ray activity to which it owes its origin, the rate of decay having reference primarily to the process of emission of the α or β particle. The association of the 7-rays with the α or β -rays in the manner stated can be checked by recording the two sets of rays in separate counters and observing that the pulses due to them appear to occur simultaneously.² One counter is surrounded by sufficient absorbing material to stop the α or β -rays, so that it counts only the 7-rays; in the other counter ionization due to the 7-rays is made negligible.

Artificially activated nuclei, on the other hand, sometimes exhibit a 7-ray activity which is not closely associated with α or β -ray activity but nevertheless decays only at a measurable rate. In such cases the upper level from which the nuclear transition starts is called a *metastable* nuclear level, in analogy with the metastable states of atoms; for some reason the transition from this level into lower lying levels has an extraordinarily small probability of occurring. An example is described in Sec. 221(a).

The emission of a 7-ray thus involves no change in the actual constitution of the nucleus, being analogous in this respect to the emission of line radiation by atoms.

202. Beta-ray Spectra.—When the β -rays from a radioactive substance are spread out into a spectrum according to their velocity,

¹ ELLIS and WOOSTER, *Cambridge Phil. Soc. Proc.*, vol. 22, p. 844 (1926).

² BOTHE and v. BAEYER, *Göttingen Nachrichten*, vol. 1, p. 195 (1935).

by means of a magnetic field, it is commonly found that the spectrum consists of narrow lines superposed upon a continuous background. In the beginning it was thought that all β -rays come from the nucleus, but later work appears to have shown, as we have indicated, that all of the *lines* are due to the ejection of *circumnuclear electrons* from the atom. In most cases these β -rays consist of *photoelectrons* produced by γ -rays, in the process called internal conversion of the γ -rays. One line from a disintegration product of radium C, however, is perhaps due to direct interaction between an excited nucleus and the K shell of electrons, the nucleus jumping to a lower nuclear energy level without the emission of radiation and thereby furnishing the energy for the ejection of a *if* electron [in analogy with the Auger transitions in the X-ray region, Sec. 190(6)].

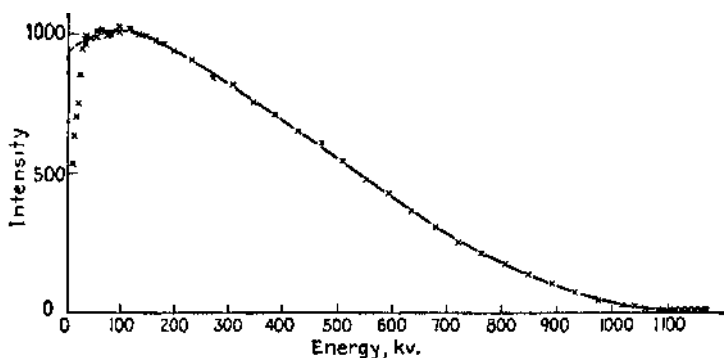


FIG. 193.— β -ray spectrum of RaE. The ordinate represents, on an arbitrary scale, the number of β -rays emitted per unit of their kinetic energy.

The β -rays forming the *continuum*, on the other hand, are believed always to come from the nucleus. In confirmation of this interpretation, measurements have shown that, in the emission of such *primary* β -rays, as they are called, just one ray is emitted for each nucleus that is transformed.

When the number of β -rays per unit of their energy is plotted against the energy as abscissa, a curve is obtained which rises quickly to a maximum and then sinks slowly to zero at an upper limit of energy. The observed maximum energies range from 6×10^6 to over 3×10^6 electron-volts. An experimental curve obtained for RaE (period, 5 days) is shown in Fig. 193.

Such curves show that the energy of the β -ray which is emitted by atoms undergoing a given type of radioactive transformation varies greatly from atom to atom. This fact presents great difficulties in regard to the conservation of energy. The kinetic energy of the β -ray should represent the energy lost by the nucleus in undergoing

the transformation. The loss in energy might have one of several different values, according as the nucleus is left by the β transformation in one or another of those nuclear levels whose existence is revealed by the T-ray spectra; but a variation due to this cause would result in a line structure in the spectrum of the primary β -rays, instead of a continuum.

It was suggested at one time by Bohr that the law of the conservation of energy might not hold for the β -decay of nuclei. There is no other indication of a failure of this law, however. At the present time, preference is given to a hypothesis proposed by Fermi. He assumes that another particle of a new kind, called a *neutrino*, is emitted along with the 0-ray. The neutrino is supposed to carry no charge and to have a very small mass, hence it can pass unhindered through miles of matter and so has eluded experimental detection. The *upper limit* of the β -ray energy is then regarded as representing the energy that is lost by the nucleus in its transformation; the difference between this energy and that of a given β -ray is carried off by the neutrino. This theory of nuclear transformations is still in the speculative stage but it is widely held today. Further evidence in its favor is described in Sec. 225.

203. Observations on Individual Charged Particles.—At this point it may be instructive to describe briefly some additional methods by which it has been found possible to make observations on *single* charged particles. The experiments described hitherto have been such as to yield only statistical results, from which the properties of individual particles could only be inferred by means of indirect reasoning.

In Sec. 198(a) above, we described the method by which Rutherford and Geiger counted the α -rays emitted from a radioactive substance. Each α particle was made to cause a separate pulse of ionization in an ionization chamber, and these pulses, detected with an electrometer, were counted. This method has been used extensively to count particles of all sorts, under the name of a *Geiger counter*. Even γ -rays can be "counted" with such devices, the ionization in the counter being caused by secondary electrons ejected from molecules of the enclosed gas or from the walls of the counter.

Various changes have been made in the ionization chamber, and still greater changes in the method of counting the impulses of ionization.¹ Nowadays the small electrical impulses from the ionization chamber are usually magnified many times by means of an electron-

¹ Cf. RASETTI, *op. cit.* p. 18; articles in *Rev. Sci. Instruments*, or articles describing observations made with the use of counters; WYNN-WILLIAMS, *Roy. Soc. Proc.*, vol. 136, p. 312 (1932).

tube amplifier. In the experiments of Holloway and Livingston referred to in Sec. 198(6) above, amplified impulses up to 250 volts were observed, each due to the entrance of a single α particle into the ionization chamber.

For α particles, an alternative but somewhat less reliable method is to count the scintillations or flashes of light produced by them when they fall upon a phosphorescent screen, which may be coated with impure zinc sulfide. By comparing counts made in this way with those made using a Geiger counter, it has been shown that each scintillation is produced by a single particle. The individual scintillations can be seen easily under a microscope.

The most instructive method of studying individual charged particles is to observe the tracks made by them in a Wilson *expansion chamber* or *cloud chamber*.¹ The method depends upon the fact that when dust-free air saturated with water vapor is expanded quickly and thereby cooled, if no ions are present, a considerable expansion is necessary in order to cause precipitation of the vapor, whereas if ions are present, a somewhat smaller expansion causes a droplet of water to condense about each ion as a center. In this way the path of an ionizing particle can be made visible as a row or "track" of cloud droplets condensed about the ions.

The cloud chamber is commonly made from a few to many centimeters deep and considerably wider, and the expansion of the air contained in it is produced by jerking the floor of the chamber downward, *e.g.*, by suddenly releasing the air below it through a valve into an evacuated vessel. In a modification due to Shimizu, the floor of the chamber consists of a tight-fitting piston which is caused to move up and down periodically. Vapors of liquids other than water, such as alcohol, are often used. The particles to be observed are admitted to the chamber just before the expansion; and, after the pressure has been restored, an electric field may be applied momentarily to clear the space of ions. The tracks are illuminated by a strong beam of light and are either observed visually or photographed through the glass top or sides of the chamber. By using two cameras pointing at different angles, stereoscopic pictures of the tracks are frequently made. {*Cf.* Fig. 226 in Sec. 235 and Fig. 231 in Sec. 237.) The whole apparatus is often arranged to operate automatically, expansions being produced at intervals of 10 to 30 seconds. Photographs may be taken on motion-picture film.

¹ WILSON, C. T. R., *Roy. Soc., Proc.*, vol. 85, p. 285 (1911); vol. 87, p. 277 (1912); vol. 104, p. 1 (1923). *Cf.* RUTHERFORD, CHADWICK, and ELLIS, *op. cit.*, p. 57; DAS GUPTA and GHOSH, *Rev. Modern Phys.*, vol. 18, p. 225 (1946).

Since an α particle produces something like 50,000 droplets per cm. in air at ordinary pressure, it makes a heavy, solid-looking track in a cloud chamber. The tracks are practically straight, except for an occasional kink where the α particle came very close to a nucleus and suffered an appreciable deflection (*cf.* Fig. 194). The distribution of such paths as to range is easily studied in a cloud chamber. Protons also make solid-looking tracks, but they are more slender. An electron, on the other hand, producing only perhaps 100 droplets per centimeter, looks thin and knobby; under a low-power microscope the individual droplets can be distinguished and counted (*cf.* Fig. 231 at Sec. 237). A γ -ray produces no track at all, except that starting from



FIG. 194.—Cloud-chamber tracks of α particles from polonium.

some point along its geometrical path there may be seen the short track of an electron ejected by it from a molecule of the gas.

The cloud chamber has been of immense service in the study of high-speed particles of all sorts, because of the possibility of observing the effects caused by the individual particles. Cloud-chamber observations have always yielded results in harmony with those obtained by other methods.

204. Artificial Transmutation by Alpha Particles.—The occurrence of natural disintegration of the nucleus in the radioactive elements early suggested the question whether it might not also be possible to disintegrate nuclei by bombarding them artificially with high-speed particles. Since, however, the particles emitted in radioactive changes have energies of the order of several million electron-volts, it was expected that bombardment energies of the same order would be

necessary. For many years the only source of particles possessing the requisite energies were the radioactive emissions themselves.

In 1919, it was shown by Rutherford¹ that nitrogen atoms, when bombarded by the α particles from radium C, would emit swiftly moving protons. A sketch of his apparatus is shown in Fig. 195. One end of a box B was closed by a metal plate E , in the center of which was a hole covered by a thin sheet of silver foil W . Just outside the window was a zinc sulfide screen, the scintillations on which were observed by a microscope M . The Source of α particles, consisting of radium C, was placed on a small disk D , whose distance x from W could be varied at will. $A A$ were side tubes through which various gases could be introduced into the enclosure.

When the box was filled with oxygen or CO_2 at atmospheric pressure, no scintillations were observed on the screen so long as x exceeded

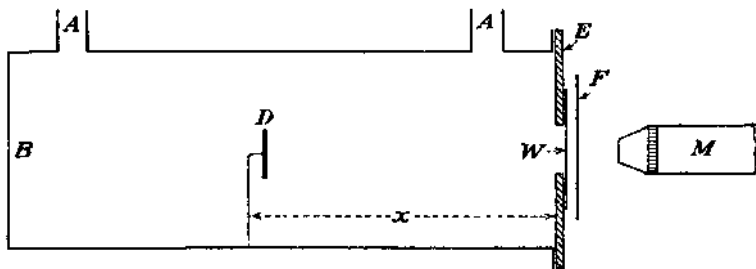


FIG. 195.—Apparatus by which Rutlieiford observed production of long-range particles when α particles collide with atoms of some of the lighter elements.

the "range" of the α particles emitted by radium C—in these gases about 7 cm. When, however, the box was filled with nitrogen, scintillations were observed when x was as great as 40 cm. Since the α particles from the radium source could not penetrate 40 cm. of air, Rutherford proposed the hypothesis that the α particles, by colliding with the nitrogen nuclei, caused the disintegration of the latter, with the emission of long-range particles which produced the scintillations. Measurements of the magnetic deflection of these particles suggested that they were protons—a surmise confirmed by subsequent work. Rutherford and Chadwick² later showed that many other light elements could be similarly transmuted. To study the phenomenon in solids, the source D was moved close to the window W , which was replaced by a thin layer (foil where possible) of the element under study, and the scintillations produced by the ejected

¹ RUTHERFORD, *Phil. Mag.*, vol. 37, p. 581 (1919).

² RUTHERFORD and CHADWICK, *Phil. Mag.*, vol. 42, p. 809 (1921) and vol. 44, p. 417 (1922); *Phys. Soc., Proc.*, vol. 36, p. 417 (1924).

THE STRUCTURE OF NUCLEI

206. The Composition of Nuclei.—A survey of the data on radioactivity in the light of Rutherford's nuclear theory of the atom at once raises the question as to the constitution of the nucleus. A mechanism capable of emitting α -rays, β -rays, and a whole spectrum of monochromatic γ -rays must possess a well-ordered, though probably very complex, structure. And, first of all, of what elementary particles is the nucleus composed?

Since electrons were known to be emitted from some nuclei, it was natural to assume in the beginning that electrons formed one constituent of many, perhaps all nuclei. Such an assumption would make it possible to carry out Prout's old hypothesis in a new form, by assuming that all nuclei consisted of electrons and protons (a proton being the nucleus of a hydrogen atom). Radioactive nuclei have never been found to emit protons, to be sure; and they do often emit α particles, which are helium nuclei. But the helium nucleus, being almost exactly four times as heavy as a proton with a positive charge twice as great, might consist of 4 protons combined with 2 electrons, the latter adding very little to the mass but serving to make the nuclear charge come out right. The fact that 4 separate protons are slightly heavier than a helium nucleus is not necessarily fatal to this hypothesis; for the attractive forces holding the 4 protons and the 2 electrons together would presumably be associated with a negative potential energy, and, according to the relativistic relation between mass and energy, this negative energy would be associated with a diminution in mass.

On the basis of wave mechanics, however, certain objections can be raised against the assumption that electrons exist inside of nuclei. According to the indeterminacy principle (Sec. 112), the indefiniteness of position of a particle in one direction, Δx , and the indefiniteness of the corresponding component of momentum, Δp , are always at least large enough so that

$$\Delta x \Delta p \cong h,$$

h being Planck's constant and $\cong$ meaning "is of the order of." Now, if we make $\Delta x = 2 \times 10^{-12}$ cm., which is of the order of nuclear diameters [c/. Eq. (322) in Sec. 208],

$$\Delta p = \frac{h}{\Delta x} = \frac{6.6 \times 10^{-27}}{2 \times 10^{-12}} = 3.3 \times 10^{-15}.$$

To estimate the indefiniteness in energy, we must use relativistic

formulas. For the total energy W of a particle with momentum p , Eq. (66) in Sec. 65 gives

$$W^2 = m^2c^4 + p^2c^2.$$

If we make p only as large as the value just found for Ap, the term $p^2c^2 = (3.3 \times 10^{-16})^2(3 \times 10^{10})^2 = 10^{-8}$ erg.² For an electron, this is easily seen to exceed greatly the term m^2c^4 . Hence for such a value of p we may write

$$W^2 = p^2c^2 = 10^{-8}, \quad W = 10^{-4} \text{ erg} = \frac{10^{-4}}{1.6 \times 10^{-12}} = 6 \times 10^7 \text{ ev}$$

[cf. Eq. (144) in Sec. 96]. Only a small part of W is represented by the rest energy of the electron, which is

$$mc^2 = (0.9 \times 10^{-27}) \frac{(3 \times 10^{10})^2}{1.6 \times 10^{-12}} = 5 \times 10^5 \text{ electron-volts.}$$

Thus a free electron confined within a space as small as a nucleus would have to have a kinetic energy of the order of 6×10^7 electron-volts. In a nucleus, no doubt, an electron would really be in some sort of quantum state; but a more rigorous wave-mechanical argument leads to the same conclusion as to its energy. Yet, experimentally, we find that when an electron issues from a nucleus as a β -ray, its kinetic energy never exceeds 4×10^6 electron-volts, or $1/15$ of that just found. In view of the contrast between these two numbers it seems improbable that nuclei can contain electrons. There are also other theoretical objections that are connected with nuclear spin, but these we shall not discuss.

In view of these difficulties with the assumption that nuclei contain electrons, the discovery of the neutron in 1932 led at once to the general adoption of a radically different theory of nuclear composition. In the theory generally held today it is assumed that

1. Nuclei are composed of protons and neutrons.
2. When a negative electron is emitted from a nucleus as a β -ray, it is created in some way at the instant of emission, a neutron in the nucleus changing at the same time into a proton. When a positron is emitted (Sees. 212 and 213), a proton changes into a neutron.

Thus the idea of the conservation of electrons as particles has been abandoned, but conservation of electrical charge is retained. Since protons and neutrons can be converted into each other in the nucleus,

they are sometimes regarded as two alternative states of a single kind of heavy particle, to which the name *nucleon* has been given.¹

According to this view, the helium nucleus, with atomic number $Z = 2$ and mass number $A = 4$, consists of 2 protons and 2 neutrons. The combined masses of 2 separate protons and 2 neutrons exceed, to be sure, the mass of the helium nucleus; but this excess is explained as in the older theory of nuclear composition by assuming that a considerable loss of energy occurs as the four particles are brought together into a helium nucleus, and this loss of energy is accompanied by a loss of mass.

A similar loss of mass presumably occurs in the formation of all heavier nuclei. The variation of the loss of mass from one isotope to the next can be reduced to a minimum by assuming that in all cases the combined number of protons and neutrons in a nucleus is represented by its mass number A as already denned; and this hypothesis is now universally adopted. The atomic number Z then represents the number of the protons alone, whereas the number of neutrons in the nucleus is $A - Z$.

207. Nuclear Mass and Rest Energy.—The comparative masses of nuclei are extremely important because of the significance that they possess in terms of nuclear energy. It will be found most convenient, however, to deal, not with the total isotopic mass itself, but with the difference between the isotopic mass (or the mass expressed in atomic mass units) and the mass number A . This difference may be called the *mass correction* (relative to 0^{16}).² In any nuclear reaction, since there is no change in the total number of protons and neutrons, and hence also none in the sum of the mass numbers, the total change in nuclear mass is also given by the change in the sum of the mass corrections.

The neutron may be considered as a nucleus with $Z = 0$, $A = 1$, and a positive mass correction equal to 0.00893. The majority of the known nuclei have negative mass corrections. This is a consequence of the choice of the 0^{16} scale. The mass correction will vanish for a nucleus in which the average mass per constituent particle happens to be exactly the same as in 0^{16} .

In Fig. 197 the mass corrections of **many** nuclei are shown plotted against the atomic mass number A . Stable isotopes of the same element have usually almost the same mass correction, hence, except for hydrogen, lithium, and boron, each element is represented by a single

¹ MOLLER, *Phys. Rev.*, vol. 59, p. 323 (1941).

² The difference between isotopic mass and mass number has been called a mass defect, but some such term as "mass correction" seems to be less confusing.

isotope. For small values of Z there is an evident tendency for the mass correction to be larger when the atomic number is odd than when it is even. The mass correction has its algebraic minimum value about at krypton, $Z = 36$.

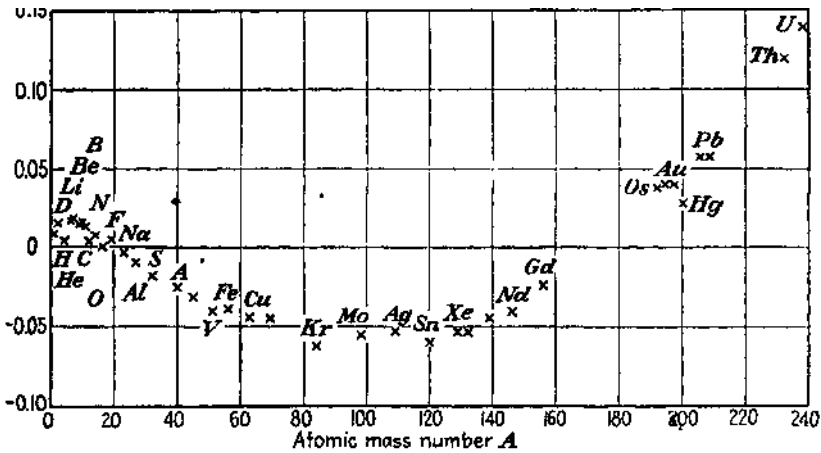


FIG. 197.—Mass correction (or mass excess) of certain nuclei, plotted against the mass number A .

Another quantity that has been used very often in the discussion of nuclear masses is the *packing fraction*, defined by Aston as the ratio of the mass correction (on the usual 0^{16} scale) to the mass number A .

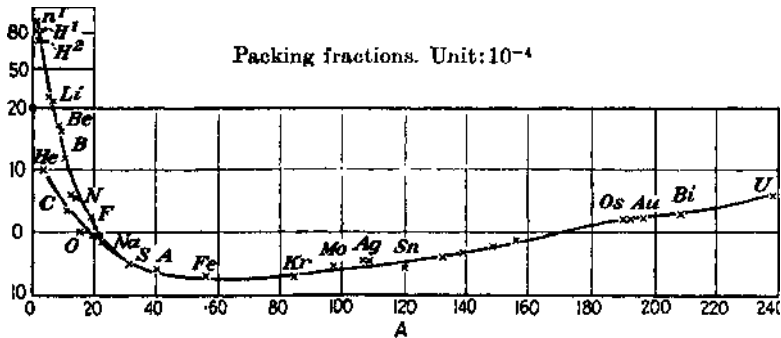


FIG. 198.—Approximate curve showing packing fraction $(M_0 - A)/A$ as a function of mass number A (M_0 = isotopic mass). The packing fractions of certain isotopes are indicated by crosses.

An approximate curve of packing fractions as a function of mass number is shown in Fig. 198; and the values of the packing fractions themselves, so far as they are known, are given in Appendix I. On the plot the packing fractions for a few isotopes are indicated by crosses.

The minimum packing fraction occurs about at iron, $Z = 26$, earlier than the minimum for the mass correction because of the progressive increase in A ,

Furthermore, according to the relation between mass and energy that is suggested by relativity, as described in Sec. 65, each gram of mass represents also c^2 grams of rest energy, where c is the speed of light in centimeters per second. The amount of energy thus associated with each atomic unit of mass, whose magnitude is given in Eq. (315) in Sec. 195, is $1.660 \times 10^{-24} \times (2.9978 \times 10^{10})^2$ or 1.492×10^{-3} erg; divided by 1.601×10^{-12} [Eq. (144) in Sec. 96], this equals 9.32×10^8 electron-volts. It is customary in nuclear physics to use ev (or eV , EV) as an abbreviation for the electron-volt and Mev (sometimes MEV , MV) for a million electron-volts (sometimes Bev for 10^9 ev). Thus we can write for the energy associated with each atomic mass unit:

$$1 \text{ a.m.u.: } 1.492 \times 10^{-3} \text{ erg} = 932 \text{ Mev.} \quad (318)$$

Any nucleus of mass M in atomic mass units thus contains 932 M million electron-volts of rest energy. For future use, we may note also the energy equivalent of the electronic mass. Calculating m as in Sec. 195(c) and using Eq. (318), we find $mc^2 = 0.000594 \times 932$ or

$$mc^2 = 0.511 \text{ Mev.} \quad (319)$$

In any nuclear reaction, a change in the sum total of the masses involved implies a proportionate change in the sum total of the rest energies. If rest mass disappears, a corresponding amount of energy must appear in other forms, usually either as kinetic energy of various particles, as the rest energy of particles newly created, or as electromagnetic radiation, usually γ -rays. When the total rest mass increases, there must be a corresponding net conversion of energy of other forms into rest energy.

In calculating the change of mass, as already stated, the masses of neutral atoms rather than nuclear masses may be used, in dealing with any change in which the total number of circumnuclear electrons remains fixed; and, in turn, mass corrections may be used instead of atomic masses, since the sum of the mass numbers of the atoms involved is always conserved (in reactions so far discovered).

Consider, for example, the process, imaginary at present, which consists of the binding of 2 protons and 2 neutrons into a helium nucleus. The sum of the initial mass corrections is

$$2 \times 0.00813 + 2 \times 0.00893 = 0.0341;$$

whereas the mass correction of a helium atom is 0.0039. Thus the reaction stated involves a loss of rest mass of magnitude

$$0.0341 - 0.0039 = 0.0302 \text{ a.m.u.}, \quad (320)$$

and a loss of rest energy equal to

$$0.0302 \times 932 = 28.1 \text{ Mev.} \quad (321)$$

This energy of 28.1 Mev is called the *binding energy* of the helium nucleus. The loss of mass may be called similarly the *mass loss in binding*.

To break a helium nucleus completely apart into its constituent protons and neutrons would thus require the expenditure of 28.1 Mev of energy. This is nearly three times the kinetic energy of the fastest α -rays known. Thus we can understand why α particles are never observed to break up into protons and neutrons as they fly about in the laboratory.

An alternative possibility, however, is that a helium nucleus might break apart into 2 deuterons, *i.e.*, into 2 nuclei of heavy hydrogen, H^2 . The mass spectrograph gives for the mass of an H^2 atom 2.01473 a.m.u. Thus the mass loss and binding energy of the deuteron are, respectively, only $0.00813 + 0.00893 - 0.01473$ or 0.00233 a.m.u., and 2.2 Mev. Accordingly, the energy released in the union of two deuterons into a helium nucleus is $28.1 - 2 \times 2.2 = 23.7$ Mev. This is more than twice as large as the largest α -ray energy, so that He^4 is also very stable as against disintegration into deuterons.

Similar calculations can be made for all isotopes. The mass losses and the binding energies are found to increase, from helium on, *almost in proportion to the mass number*. This fact stands out most clearly if the *binding energy per constituent particle* is considered, instead of the total binding energy. For helium the binding energy per particle is $28.1/4 = 7.02$ Mev. For heavier nuclei the binding energy per particle increases slowly from the smaller value of 5.3 Mev for Li^6 to a maximum of about 8.7 near iron ($Z = 26$), and then decreases very slowly to 7.94 for uranium (U^{238}). In Table V are listed, for a few nuclei, under M.l. the mass loss in atomic mass units, under B.e. the binding energy, and under B.e.p.p. the binding energy per particle, the latter two being expressed in terms of Mev as the unit of energy.

The binding energy per particle varies in close correspondence with the packing fraction, not quite in proportion to it because of the slight difference in mass between protons and neutrons. That the binding energy per particle does not change sign, as does the packing

fraction, is due, of course, to the fact that the binding energy per particle contains a constant part representing its value for O^{16} .

TABLE V.—MASS LOSS (M.L.), BINDING ENERGY (B.C.), AND BINDING ENERGY PER PARTICLE (B.e.p.p.), FOR CERTAIN NUCLEI

Nucleus	He ⁴	L ₁ ⁶	L ₁ ⁷	Be ⁹	B ¹⁰	B ¹¹
Z	2	3	3	4	5	5
M.l., a.m.u. . . .	0 03023	0.0343	0.0421	0.0622	0.0693	0 0813
B.e., Mev	28 18	31.9	39 2	57 9	64 6	75 8
B.e.p.p., Mev	7 04	5.32	5 61	6 64	0 46	6 89

Nucleus	C ¹²	N ¹⁴	O ¹⁶	A ⁴⁰	Fe ⁵⁶	Ag ¹⁰⁷	Au ¹⁹⁷
Z	6	7	8	18	26	47	79
M.l., a.m.u.	0 0984	0 1119	0 1365	0 368	0 518	0 970	1 657
B.e., Mev	91 7	104 3	127 2	343	483	904	1544
B.e.p.p., Mev.	7 64	7 45	7 95	8 57	8 62	8 45	7 71

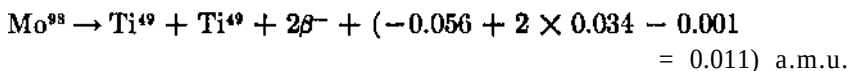
Nucleus	U ²³⁸					
Z	92					
M.l., a.m.u.	1 912					
B.e., Mev	1782					
B.e.p.p., Mev.	7 49					

The binding energy has reference to the union of fundamental particles, protons and neutrons, into a nucleus. More generally, it is interesting to consider the union of any two nuclei into a larger one, or the possibility of breaking up a large nucleus into two smaller ones. A study of the isotopic masses leads to the conclusion that, so far as the conservation of energy is concerned, it would be possible for any two of the lighter known nuclei to combine together, provided the mass number of the product nucleus so formed would not exceed about 70. Thus, nitrogen N^{14} and aluminum Al^{27} could combine into potassium K^{41} , with a loss of rest mass of

$$0.0075 - 0.01 - (-0.02G) = 0.0235 \text{ a.m.u.};$$

but a positron would have to be emitted in order to decrease Z from $7 + 13$ to 19 , requiring a rest mass of 0.0005 a.m.u., so that the energy actually released would be $0.023 \times 932 = 21.5$ Mev. Or, Si^{30} and A^{40} might unite to form $2n^{70}$ after emitting $14 + 18 - 30 = 2$ positrons, with a net loss of 0.003 a.m.u. and an emission of 1.8 Mev of energy.

On the other hand, it appears that any of the *heavier*' nuclei from molybdenum ($Z = 42$) upward could divide into two lighter nuclei, with the emission of one or more negative β -rays in order to adjust the Z values. Examples:



or **10.3 Mev; $\text{Pb}^{208} \rightarrow \text{Sn}^{124} + \text{Kr}^{84} + 4\beta^- + 0.174 \text{ a.m.u.}$ or 162 Mev.**

Only limitations set by the conservation of energy have been considered here, of course. The transformations contemplated may be impossible for other reasons. The famous fission of uranium and several other elements, described in Sec. 227, is a first step toward tapping such atomic sources of energy.

208. The General Theory of Nuclear Binding.—It is clear that the protons and neutrons composing a nucleus must be held together by attractive forces of some sort, otherwise the mutual electrostatic repulsion of the protons would disrupt the nucleus. The true nature of these attractive forces and the exact law of their action are not known with certainty. In this respect the present situation in nuclear theory is quite different from that in ordinary atomic theory. The mathematical form of the interaction between electric charges, for some reason, is the same at atomic distances as it is at physically observable distances; the only difference between the two cases is that at large distances it is possible to use the approximate, classical form of mechanics, whereas on the atomic scale the more exact wave-mechanical form must be employed. Thus, historically, observations of electrostatic interaction on the ordinary physical scale led at once to a knowledge of this interaction on the atomic scale. In constructing a nuclear theory, on the other hand, our only resource is to endeavor to infer the laws of the interaction between the nuclear particles from the observed properties of nuclei, or patiently to try various hypotheses concerning these laws and to compare deductions from them with the results of experiment. We shall not enter upon a detailed discussion of the hypotheses that have been proposed, but shall restrict the discussion to certain ideas which have proved useful and which seem destined to form a permanent part of nuclear theory, referring the student elsewhere for further information.¹

One general property that the nuclear forces must possess is indicated by certain features of the scattering of α particles by atomic

¹ Cf. RASETTI, *op. cit.*; PEIERLS, *The Theory of Nuclear Forces*, *Nature*, vol. 145, p. 687 (1940); BETHE, *Rev. Modern Phys.*, vol. 9, p. 69 (1937); and many articles in the literature.

nuclei. The scattering is found to obey the law deduced by Rutherford (Sec. 91) so long as the distance of closest approach of the α particle to the nucleus, as calculated from its velocity and the angle of scattering, remains greater than a certain distance p , which varies from one kind of nucleus to another. At smaller distances departures from Rutherford's formula occur; the scattering is then said to be "anomalous." Since Rutherford's original deduction rested on the assumption of the ordinary inverse-square law of force, and since, according to classical theory, this assumption would be justified if the nuclear charge were symmetrically distributed within a distance p of the center of the nucleus, the value of p deduced from the scattering experiments is regarded for convenience as the "radius" of the nucleus. Classical theory is not really applicable to the scattering phenomenon, to be sure, but wave mechanics, also, leads to Rutherford's formula, and it is thought that the classical interpretation of p may represent certain features of the unknown accurate nuclear theory. The values of p deduced from the experiments are represented roughly by the equation

$$\rho = 2 \times 10^{-13} Z^{1/2} \text{ cm.} \quad (322)$$

As indicated by this equation, $p \propto Z^{1/2}$, nearly. Thus the *volume of the nucleus*, $\frac{4}{3}\pi\rho^3$, is *proportional to Z* . Since Z , in turn, is nearly proportional to the number of particles in the nucleus, we reach the conclusion that the *density of all nuclei is nearly the same*. In this respect, there is a strong contrast between the nucleus and the electronic envelope of the atom, in which the mean density of the electrons increases rapidly with increasing Z . The fact that the density does not change much when one more particle is added implies that the interactions between particles which hold the nucleus together are *sharply limited in range*, so that the addition of a particle affects only those near to it. The interaction between two particles must, therefore, drop off quickly to zero when the distance between them exceeds a certain limit; and the observed values of p imply that this limiting distance must be of the order of 10^{-13} cm. or less.

A nucleus thus resembles somewhat a drop of liquid, held together by cohesion. It is doubtless more penetrable, however; perhaps a swarm of bees in the air would be a better analogy. The latter comparison illustrates also the fact that, like the electrons in the atom, the individual neutrons and protons do not have any fixed places in the nucleus. The wave function for the nucleus probably does *not* represent the component particles as *changing position from time to time*, however; it merely assigns a certain probability for each possible

position in the nucleus, this probability being the same for each neutron, or for each proton. In Fig. 199 is shown a diagram that may help in visualizing some of these conclusions.

Besides its nearly constant density, a nucleus also resembles a drop of liquid in the fact, noted in the last section, that the binding energy is almost proportional to the number of constituent particles. Just as the heat that is given off when a molecule of vapor condenses onto a drop of liquid is independent of the mass of the drop, so the energy released when one more neutron or proton is added to a nucleus is (nearly) independent of Z , amounting roughly to 8 Mev per particle.

Fig. 199.—
Illustrating a nucleus composed of protons (•) and neutrons (°).

The classical analogies must not be taken too literally, however. Classical mechanics is presumably applicable to nuclear phenomena only when the energies of the particles are high enough to make their de Broglie wave lengths small as compared with nuclear dimensions. According to Eqs. (166a,fr) in Sec. 107, to have a wave length even as small as 10^{-13} cm., a neutron or proton would have to have kinetic energy equal to

$$\frac{1}{2}mv^2 = \frac{(mv)^2}{2m} = \frac{(h/\lambda)^2}{2m} = \frac{(6.61 \times 10^{-27}/10^{-13})^2}{2 \times 1.66 \times 10^{-24}}$$

or

$$1.3 \times 10^{-3} \text{ erg} = 830 \text{ Mev.}$$

This is large as compared with the amount of energy associated with one particle in a nucleus. Presumably, therefore, the correct theory of the nucleus must be wave-mechanical in nature.

209. Nuclear Forces and **the** Ratio of Mass Number to Atomic Number, (a) *The Attraction between Neutrons and Protons and the Neutron-proton Ratio.*—Some further results that are very suggestive have been obtained by studying mathematically the force actions that must exist if the nucleus is to be held together. The existence of the deuteron, consisting of 1 neutron and 1 proton, shows conclusively that an attraction of some sort exists between a neutron and a proton. Such attractions, between neutrons and protons might conceivably be the sole action holding the nucleus together against the disruptive tendency of the electrostatic repulsion of the protons. It is possible to account for the broad characteristics of stable nuclei on the assumption that only these two force actions are effective [but see (6) below].

For a first attack on the problem, let us neglect the repulsion of

the protons, supposing only the neutron-proton attraction to exist. Then it is easily seen that all stable nuclei should contain as nearly as possible equal numbers of neutrons and protons, since the energy of a given number A of particles will be a minimum when they are half protons and half neutrons. For, suppose a nucleus contains N neutrons and P protons, and suppose that $N > P$. Then, if a neutron were to be replaced by a proton, the energy of the nucleus would be decreased by the amount of the negative potential energy due to the added proton interacting with $N - 1$ neutrons, and increased by the loss of the negative energy due to the interaction of a neutron with P protons. If $N - 1 > P$, the net effect would be a decrease in the nuclear energy. Hence, under these circumstances, there would be a tendency for a neutron to emit an electron as a β particle and to turn into a proton, thereby decreasing the nuclear energy and at the same time making N more nearly equal to P . In the same way, we may infer that, if $P - N > 1$, a proton should emit a positron and turn into a neutron.

The decrease in energy due to such a change would have to be sufficiently great, to be sure, to provide not only some kinetic energy for the emitted particle but also its rest energy mc^2 ; otherwise no change could occur. The absolute value of the energy interaction of proton and neutron must be relatively large, however. In Sec. 207 we have seen that energy equal to 2.2 Mev is required to break up a deuterium nucleus into a neutron and a proton. In all heavier nuclei the binding energy per particle is of the order of 7 or 8 Mev. In comparison with such energies, the rest energy of an electron or positron, 0.51 Mev as in Eq. (319), is rather small. Thus, we should expect a strong tendency for N and P to become nearly equal as a result of positive or negative β -ray activity.

The *electrostatic repulsion of the protons* must next be considered. The potential energy due to this repulsion adds a positive term in the energy, roughly in proportion to P^2 . Hence, its presence will cause P to tend to diminish, and the actual ratio of N to P will be that at which there is a balance between the tendency of the proton-neutron attraction to bring N and P into equality and the tendency of the proton repulsion to diminish P . Our final conclusion is thus that N should somewhat exceed P , at least in the heavier atoms.

The magnitude of the energy due to proton repulsion can be estimated from the classical picture of the nucleus described above. Equation (322) indicates that the distance between protons in a nucleus should be of the order of 10^{-13} cm. The electrostatic energy of 2 protons 10^{-13} cm. apart is only

$$\frac{(4.80 \times 10^{-10})^2}{10^{-18}} = 2.30 \times 10^{-6} \text{ erg} = 1.4 \text{ Mev.}$$

This is small as compared with the total binding energy of a helium nucleus, or 28 Mev. From this example it is clear that proton repulsion will have little effect upon the ratio N/P in all light nuclei, so that this ratio should be close to unity.

The repulsive effect will rise rapidly, however, as Z increases. If the 80 protons in a mercury nucleus, for example, were distributed uniformly over the surface of a sphere of 10^{-12} cm. radius, which is a little larger than the value of p as given by Eq. (322), their electrostatic energy, computed classically, would be

$$\frac{(Ze)^2}{2r} = \frac{(80 \times 4.80 \times 10^{-10})^2}{2 \times 10^{-12} \times 1.60 \times 10^{-12} \times 10^6} = 460 \text{ Mev.}$$

If they were scattered throughout the sphere, as they are in reality, their energy would be still larger. On the other hand, the total energy that is released in binding 120 neutrons and 80 protons into Hg^{200} , with a mass of 200.028, is only

$$120 \times 0.00893 + 80 \times 0.00812 - 0.028 = 1.69 \text{ a.m.u. or } 1,570 \text{ Mev.}$$

This is less than four times our estimate of the positive repulsive energy. Thus, in the heavier nuclei, the effect of proton repulsion becomes comparable with that of the neutron-proton attraction, and the number of protons should, in consequence, be considerably smaller than the number of neutrons in the nucleus.

The composition of the nuclei that occur in nature is in good agreement with these deductions from our tentative assumptions. A glance at the table in Appendix I shows that there is a strong tendency for N , which equals $A - Z$, to equal P or Z up through calcium, $Z = 20$. Beyond this point N always exceeds P , the ratio N/P increasing to about 1.5 in the heaviest nuclei.

The occurrence of isobars, or isotopes having the same mass number A but different atomic number Z , might seem to present a difficulty for the theory just stated, since, for a given number A of particles, there should be a unique value of P or Z that makes the energy a minimum. Among the atoms that occur in nature, however, isobars are not abundant, and, where they do occur, usually only two, rarely three, different values of Z occur for a given value of A . Perhaps in such cases the difference in energy is too small to permit transformation, by the emission of a β -ray, into the isobar of lowest energy. The known masses are not sufficiently accurate as yet to permit a general test of this explanation. Furthermore, in a few cases β -ray activity is

actually observed to be going on; for example, a rare isotope of potassium, ^{40}K , isobaric with the commonest isotope of calcium, ^{40}Ca , is responsible for the slight radioactivity which potassium has long been known to exhibit. Many isobaric isotopes have now been produced artificially, but nearly all of them are strongly radioactive.

All known nuclei with $Z > 83$ are likewise radioactive, presumably because of the overwhelming effect of proton repulsion. Probably the nonoccurrence in nature of nuclei with $Z > 92$ is due to the rapidity with which such nuclei undergo radioactive change.

(b) *The Interactions of Neutrons with Neutrons and of Protons with Protons.*—Since a proton and a neutron interact, in spite of the absence of an electric charge on the neutron, it seems probable that neutrons also interact with one other. If this is so, then it may be surmised further that a similar short-range attractive interaction may exist also between 2 protons, superposed upon their electrostatic repulsion; it may be that 2 protons, if they come close enough together, will attract each other.

The interaction between protons can be studied by observing the scattering at wide angles of high-speed protons by the element hydrogen, which is just a collection of almost stationary protons so far as the impinging proton is concerned, the deflecting effect of an electron being negligible because of its very light mass.¹ Careful experiments actually point to the existence of an attractive interaction between 2 protons which can be said very roughly to overbalance their repulsion when they are less than e^2/mc^2 or 3×10^{-13} cm. apart.² (All such statements are essentially of the nature of classical *analogies*, of course; the mathematical analysis of the data necessarily follows wave-mechanical lines throughout.)

The attractive effect between 2 protons, as inferred from the scattering experiments just mentioned, is about the same as that between a proton and a neutron. It is natural to suppose, then, that a similar attraction exists also between neutrons. *If the neutron-neutron and proton-proton attractions are equal*, it is easily seen that the conclusions drawn above concerning nuclear composition remain valid. Since these conclusions in turn agree with the facts, we may infer that the two attractive effects, if not equal, must be at least not very different.

¹ According to classical mechanics an electron would be set in motion by a proton passing close to it, but, by conservation of momentum, the effect on the velocity of the proton would be very small. Corresponding results follow from wave mechanics.

² Cf. BREIT, THAXTON, and EVERBLTTD, *Phys. Rev.*, vol. 55, p. 1018 (1939).

(c) *Proton and Neutron Spin*.—Another point to consider, in dealing with the binding of particles into a nucleus, is the possibility of proton and neutron spin. In the deuteron H^2 , the band spectra of deuterium, H^2 , indicate the existence of an angular momentum corresponding to a quantum number $l=1$. This is easily accounted for if the neutron and proton each has spin with $s = \frac{1}{2}$. In spectroscopic terminology, we may then explain the value $J = 1$ by supposing that the nucleus H^2_2 is normally in a triplet state (cf. Sees. 138 and 140).

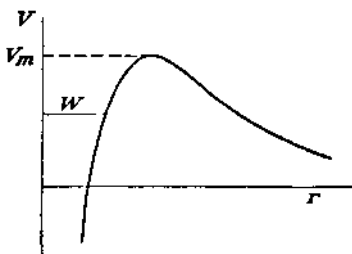
If these conclusions as to spin are correct, there may well be a tendency for neutrons, and likewise protons, to occur in pairs in the nucleus, the members of each pair having opposite spins. This is in agreement with the observed facts. The smallest nucleus that can be constructed out of such pairs is the helium nucleus, He^4 . As we have seen, the binding energy per particle of He^4 (7.0 Mev) is of the same order of magnitude as that of all heavier nuclei, and is over three times greater than that of the deuteron (2.2 Mev), which contains only a single neutron and proton. The next three elements after He have smaller binding energies per particle than has He^4 , as is evident from the fact that their packing fractions, given in Appendix I, are larger. Thus the structure of the helium nucleus or α particle presumably represents a peculiarly stable configuration, in analogy with the closed subshell of atomic theory.

As Z becomes large, the binding energy per particle increases again and finally becomes greater than that of He^4 . This suggests that heavy nuclei may be thought of as formed, in part at least, of helium nuclei, which are more lightly bound together than the particles in the helium nucleus itself. Mass numbers divisible by 4 do not seem to occupy any privileged position, however. On the other hand, there is an obvious tendency for the neutrons and protons to occur in pairs. Inspection of the table in Appendix I shows that elements with even atomic number Z have, as a rule, many more isotopes than those with odd Z . Furthermore, even Z usually goes with an even value of A the mass number, and odd Z with odd A , and this means that usually the number of neutrons is even. An odd value of Z is combined with an even value of A in only four stable nuclei (H^2 , Li^6 , B^{10} , N^{14}). It is also significant doubtless, that some 87 percent of the earth's crust is composed of elements of even atomic number.

210. The Explanation of Natural Radioactivity.—The broad features of natural radioactivity are easily understood in terms of the theory of nuclear structure just described, as is also the related fact that elements with $Z > 92$ do not occur in nature. Any nucleus containing more than 92 protons, even if it contains also the most favorable

number of neutrons, presumably sooner or later breaks up owing to the repulsions of the protons. In uranium ($Z = 92$) and thorium (90) among natural atoms, we see such a process of disintegration actually occurring, although at a slow rate; protactinium (Pa, $Z = 91$) is much scarcer and disintegrates more rapidly, as befits its odd value of Z . Stability is not reached until Z has sunk to 83 (bismuth).

That it should be an α particle that breaks off from a nucleus, under such conditions, rather than either single protons or neutrons or some other group of them, is not so easily seen. This fact must be a consequence of the especially low energy of the two-pair structure of the α particle. The eventual emission of ft particles, on the other hand, is readily explained. As successive α particles are emitted, thereby decreasing equally the number N of neutrons and the number P of protons in the nucleus, the ratio N/P , already greater than unity, becomes still larger; it may become so large that a neutron turns into a proton with the emission of an electron, thereby lowering the energy and bringing N/P nearer to its normal stable value. Thus is explained the occurrence among the naturally radioactive heavy elements of the emission of negative β -rays but never of positive ones.



Via. 200. Illustrating the "potential bairrier" for an α partiele in a nucleus.

It is sometimes helpful to employ the following semiclassical picture of the process of α -ray emission. Imagine a group of 2 neutrons and 2 protons in the form of an α particle to start out of the nucleus. At first, the group will be held back by the attraction of the remainder of the nucleus; as it moves out, therefore, the potential energy of the nucleus rises. The attraction rapidly decreases, however, until at a certain point it just equals the repulsion between the α particle and the protons in the remainder of the nucleus; at this point, the potential energy is a maximum. Beyond this point the repulsion predominates, the potential energy decreases, and the α particle is accelerated outward. The trend of the potential energy V of the nucleus, assumed zero at infinity, as a function of the distance r of the α particle from the center of the nucleus, will thus be as illustrated in Fig. 200. We can think of V as consituting a "potential barrier" tending to hold the α particle in the nucleus.

If we adhere strictly to classical theory, to be sure, we encounter serious difficulties in endeavoring to understand the *slow rate* of α -ray

emission on the basis of this picture. If an α particle is to escape across the barrier, it must have an energy W exceeding its maximum potential energy V_m when at the top of the barrier. While it is still inside the nucleus, therefore, having presumably the same total energy W but less potential energy, it must possess a considerable amount of kinetic energy and so must be in motion; why, then does it not issue immediately from the nucleus, instead of waiting for an appreciable length of time? Furthermore, why may one atom of a substance like radium disintegrate almost immediately, whereas another atom, apparently identical with the first, may wait a thousand years?

As in so many other cases, wave mechanics comes to the rescue. If $W > V_m$, wave mechanics indicates that the α particle should depart quickly, just as in classical theory. *Escape is also possible, however, if $W < V_m$* , contrary to the classical result. The process is similar to the Auger effect in atoms [Sec. 190(6)]. Provided $W > 0$, the waves leak through the barrier, at least at a slow rate, so that the probability of finding the particle outside the nucleus continually increases with the time. The higher the top of the barrier, *i.e.*, the greater the value of $V_m - W$, and the wider the barrier, the slower is the rate at which the wave packet streams outward and the smaller is the chance per second that the α particle issues from the nucleus. In this way, we can understand even such slow disintegration as that of uranium, where the chance of escape of an α particle is only 4.9×10^{-18} per second.

When the α particle does come out, its kinetic energy will be equal to its whole energy W . Since, as we have seen, $V_m > W$, the observed kinetic energy of the α particle sets a lower limit to the height of the barrier that must be assumed to surround the nucleus.

211. Nuclear Quantum States.—As has been shown in Sees. 208 and 209, many features in the behavior of nuclei can be understood by thinking of the nucleus in terms of classical imagery. A severe limitation to the use of such ideas must be recognized, however, in view of the common existence of sharply defined nuclear energy levels, as revealed by the line structure of r-ray spectra. A nucleus that is capable of suddenly jumping into a normal state of lowest energy, with the emission of a monochromatic r-ray, must be in a state very different from that, for example, of a vibrating drop of liquid. All known facts indicate that every nucleus possesses a discrete set of possible quantum states, and a corresponding array of energy levels. The lower levels are rather widely spaced, at least in light nuclei, but the spacing decreases as the energy increases. Above the true quantum levels, moreover, there is an array of pseudolevels representing

states in which the nucleus may exist momentarily before disintegrating with the emission of a material particle.

It is a recognized principle of quantum theory, however, that the measurement of energy requires time, and for this reason the *sharpness* of any energy level comes into evidence only in proportion to the length of time that the level has been occupied. [This is another example of the indeterminacy principle (Sec. 112).] If for any reason there exists a finite probability that the system will undergo a transition from a certain state into another state, then the energy associated with the first state is indefinite to an extent that is proportional to the probability of transition. Roughly, the indefiniteness or uncertainty ΔE in the energy E of any level has a magnitude h/T where h is Planck's constant and T is the mean life of the system in that level. The energy level itself is commonly regarded as possessing a certain width of magnitude ΔE . The energy that is released in a transition between two levels varies, from one transition to another, over a range represented roughly by the sum of the widths of the two levels.

The principle of level width just described appears to hold for nuclei, just as it does for the electron shell surrounding the nucleus. Only the normal state of lowest energy has zero width. Other states have widths that vary greatly according to the types of transition out of them that can occur.

From many nuclear states transitions are possible only in association with the emission of electromagnetic radiation in the form of γ -rays, resulting from the electric fields of the protons. The mean life of a nucleus in such a state may be long, perhaps even several days, but it is more commonly short, perhaps of the order of 10^{-16} sec. Even the shortest mean lives of such states, however, are long enough to permit a definition of the nuclear energy that appears to the physicist as quite sharp. Accordingly, all γ -rays, so far as is known, arise from transitions between relatively sharp levels, and the γ -rays arising from a given transition form a sensibly monochromatic line.

Other nuclear states (or pseudostates) are capable of undergoing change through the emission of a material particle, viz., a positron, a negative electron, a neutron, a helium nucleus, or, rarely, a deuteron or the nucleus H^2 . Sometimes such states are long-lived, the transition probability being very small; in such cases the nuclear energy of the initial state is relatively sharply defined. In many other cases, on the other hand, the emission of a heavy particle occurs extremely quickly, perhaps in a time of the order of 10^{-20} sec. The energy of a state that can undergo disintegration in this fashion is poorly defined and may be regarded as varying according to the manner in which the

state is produced. If a nucleus in such a state results from the absorption of a bombarding particle, then its energy increases as the energy of bombardment increases; and the kinetic energy of the particle that is subsequently emitted increases to the same degree, thereby preserving the conservation of energy. In this property of the nuclear quantum states, and in the fact that states capable of disintegration into heavy particles seem also usually to be rather closely spaced in energy, is to be found the explanation of the fact that the energy of bombardment can always be varied continuously over a wide range without great loss of effectiveness.

Many examples illustrating these ideas will be found in the remainder of this chapter.

POSITRONS, NEUTRONS, NUCLEAR REACTIONS

In the last five sections a brief discussion has been given of the attempts, on the whole strikingly successful, to develop a theory of nuclear structure. We shall now return to the description of the important sequence of fundamental discoveries that were made during the third decade of the present century. The field of known nuclear reactions will then be surveyed, with the astonishing phenomenon of nuclear fission as the closing topic.

212. Positrons.—When, during the years following 1928, the cloud chamber in a magnetic field came into general use for the study of cosmic-ray particles, it was observed that many of the particles, if assumed to be negative electrons, must have traversed the chamber in an upward direction. This was surprising, in view of evidence that the cosmic rays must originate well above the earth's surface. Some supposed that these tracks were in reality made by protons traveling downward.

In 1932, however, using a magnetic field of some 15,000 oersteds, Anderson obtained a photograph of a track that could only have been caused by a positively charged particle much lighter than a proton¹ (reproduced in Fig. 201). The track passed through a sheet of lead that had been placed in the chamber but was unequally curved on the two sides of the lead. This difference of curvature was ascribed, as in other similar cases, to a loss of energy by the particle in the lead, the direction of motion being assumed to be from the side on which the track was less curved toward that on which it was more curved, indicating a lower velocity. If this conclusion as to the direction of motion was correct, however, the *direction* of the curvature of this particular track was such as to indicate that the particle carried a

¹ ANDERSON, *Science*, vol. 76, p. 238 (1932); *Phys. Rev.*, vol. 43, p. 491 (1933).

positive charge. Yet it could not be a proton; for the curvature was so great that the particle, if a proton, would have an energy of only 300,000 volts, and a proton of this energy is known to have a range in air of only 5 mm., whereas Anderson's track showed no diminution of curvature, indicating no loss of energy by the particle, in a distance of 5 cm. Other tracks of the same nature were observed subsequently. Anderson concluded that these tracks were caused by positively charged particles having the same mass and numerical charge as the electron. He called the new particle, at first, a "positive electron," later a "*positron*." Negative electrons he proposed to call "*negatrons*" thus leaving the term electron to stand for particles of both

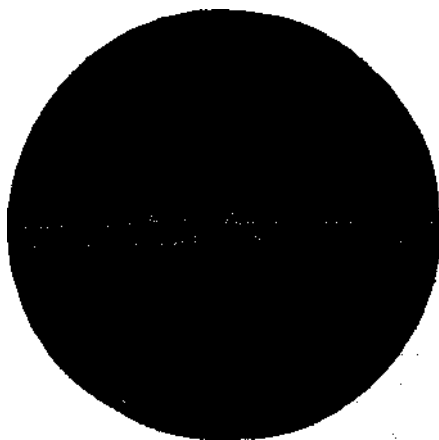


FIG. 201.—Reproduction of the first cloud-chamber photograph showing a track recognized as that of a positron. [Photograph kindly lent by Professor C. D. Anderson. *SciPhys. Rev.*, vol. 43, p. 491 (1933).]

signs; but "negatron" does not seem to be coming into use. The meaning attached to the term "electron" varies.

Subsequent work has confirmed Anderson's discovery and has revealed other experimental sources of positrons. They are observed to be emitted by many artificial radioactive substances, as described in later sections. Furthermore, they can be produced by the passage of r-rays, or even, probably, of high-speed electrons, through nuclear fields. So far as is known, they are an exact counterpart of ordinary electrons, differing only in the sign of their charge. In a cloud chamber, they produce the same thin, beady track as does an ordinary electron, but the track is curved in the opposite direction in the presence of a magnetic field.

Positrons are produced by γ -rays in a manner strikingly different from any effect of radiation so far described. It appears that when a

7-ray of sufficient energy traverses a strong *electric* field, like that near an atomic nucleus, there is a certain chance that the 7-ray may, in a sense, be *converted into an electron and a positron*. In such a process, algebraic charge is conserved, since equal amounts of positive and of negative electrification are produced. This phenomenon is called "pair production." Of the 7-ray energy $h\nu$, a part $2mc^2$ is accounted for by the rest energy of the pair of particles; the excess, $h\nu - 2mc^2$, appears as their kinetic energy. Hence, such processes can occur only if the 7-ray photon has energy $h\nu > 2mc^2 = 1.02$ Mev or a wave length

$$\lambda < 0.012 \text{ A} = 12 \text{ X.U.}$$

A partially satisfactory theory of this effect can be based on Dirac's relativistic theory of the electron;¹ and experiments have confirmed



FIG. 202.—A pair consisting of a negative electron and a positron produced by a 7-ray in the gas of a cloud chamber situated in a magnetic field. (Photograph by Curie and Joliot.)

the existence of the pair process, in conformity with the theory.² Absorption of 7-rays due to pair production has even been made the basis of a more precise method of measuring 7-ray energies; for this purpose, the kinetic energies of the pair of particles are measured in a cloud chamber placed in a magnetic field.^{3*}

A photograph obtained by Curie-Joliot and Joliot,⁴ showing a positron-electron pair produced by a 7-ray in the gas of a cloud chamber situated in a magnetic field, is reproduced in Fig. 202. Another photograph, showing three tracks, obtained by Ogle and Kruger,⁵ is shown in Fig. 203. Since two of these tracks were made by particles carrying a negative charge, this latter photograph is believed to

represent one of the rare instances in which a pair is generated in the field of an atomic electron, this electron being itself pro-

¹ Cf. HEITLER, "The Quantum Theory of Radiation," 2d ed., Chap. IV, 1945.

² ANDERSON, *Science*, vol. 77, p. 432 (1933); MEITNER and PHILIPP, *Naturwiss.*, vol. 21, p. 286 (1933); CHADWICK, BLACKETT, and OCCHIALINI, *Roy. Soc. Proc.*, vol. 144, p. 235 (1934); CURIE and JOLIOT, *Comptes Rendus*, vol. 196, p. 1581 (1933); and later articles.

³ KRUGER and OGLE, *Phys. Rev.*, vol. 67, p. 273 (1945).

⁴ CURIE and JOLIOT, *J. de Physique et le Radium*, vol. 4, p. 494 (1933).

⁵ OGLE and KRUGER, *Phys. Rev.*, vol. 67, p. 282 (1945).

jected at high speed. The threshold energy of the γ -ray for this process, according to Perrin, should be $4mc^2$; in the present instance the 2.68 Mev γ -rays from the artificial radioactive substance Na^{24} were employed, so that ample energy was present.

There are thus *three distinct ways* in which a photon can cause the *emission of an electron* from matter: by photoelectric absorption, by

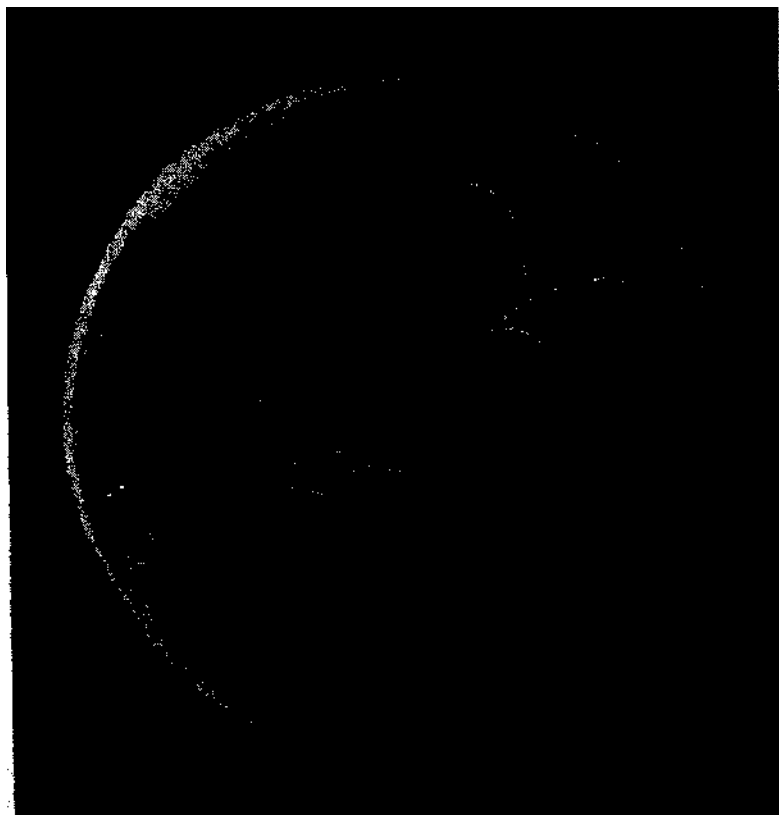


FIG. 203.—Cloud-chamber photograph showing an electron-positron pair created by a photon in the field of an electron, which is also projected and makes a third track. (Obtained by Ogle and Kruger.)

the process of Compton scattering, and by pair production. Photoelectric absorption and the pair process involve absorption of the entire photon; Compton scattering merely weakens it. In photoelectric absorption and in Compton scattering, a preexisting electron is ejected from an atom; in pair production a new electron is created, along with a positron. The relative probability of the three processes depends upon the frequency of the photon and also upon the atomic

number of the atom in which the process occurs. A beam of soft X-rays liberates many more photoelectrons than Compton electrons. As the frequency moves into the typical 7-ray region, however, the photoelectric absorption decreases rapidly, and the Compton scattering, which decreases only slowly, becomes dominant. Finally, as $h\nu$ passes $2mc^2$, pair production begins, and ultimately, at extremely high frequencies, absorption accompanied by pair production becomes the principal cause of the weakening of a 7-ray beam as it passes through matter. To give a birds-eye view of the facts, there are indicated roughly in Table VI the photon energies at which successive pairs of the three processes become equally probable in the impact of a photon upon one of 4 typical atoms.

TABLE VI

	C	Al	Cu	Pb
Photoelectric and Compton effect are equal at X.U.	0.025 500	0.05 260	0.15 85	0.5 22
Compton effect and pair production are equal at $h\nu$ (Mev)	27	15	9	5
X.U.	0.45	0.8	1.4	2.5

Since positrons can be created, we should expect that they can also be destroyed; and such indeed appears to be the fate of all positrons. The commonest process of destruction is one in which a positron unites with a free or loosely bound electron and both disappear, their entire energy passing away in the form of 2 photons of equal frequency. If their initial kinetic energy is negligible, the frequency is given by $h\nu = mc^2 = 0.51$ Mev, and the wave length is $0.024 \text{ \AA} = 24 \text{ X.U.}$ An alternative but much less probable process is the combination of a positron and an electron near a nucleus, accompanied by the emission of their combined rest energy in a single photon of energy

$$h\nu = 2mc^2 = 1.02 \text{ Mev}$$

(the initial kinetic energy being assumed negligible). Most positrons come to rest before being annihilated,¹ and are then unable to penetrate the interiors of atoms, so that the second type of annihilation mentioned becomes impossible; hence the annihilation radiation as observed

¹ Cf. BETHE, *Roy. Soc, Proc*, vol. 150, p. 129 (1935).

usually consists mostly of the frequency first mentioned or $\nu = mc^2/h$. This latter radiation has been observed by several experimenters.¹

The older observations of Gray and Tarrant² also merit a comment. They noticed that the very hard γ -rays from Th C" were scattered very strongly by lead and, upon investigation of the scattered rays, found in them a strong component of about $h\nu = 5 \times 10^6$ ev and also some indication of rays near 10^6 ev. These rays they ascribed at first to fluorescence by the nucleus. It is now believed that they arise from annihilation in the lead of positrons which are produced in the lead itself by the primary γ -rays, for which $h\nu = 2.6 \times 10^6$ ev. The mean free path of a positron so produced in lead should be only of the order of 0.5 mm, so that most of the positrons generated in the lead would also undergo annihilation in it. We have here an interesting example of the interplay that often occurs between different and apparently unconnected lines of research. Had the positron been discovered a few years earlier, Gray and Tarrant would not have wasted time elaborating their theory of nuclear fluorescence. On the other hand, had the discovery been delayed longer, the theory of nuclear fluorescence might have come to be accepted for a time.

That positrons are not commonly met with in the laboratory is due, no doubt, to the abundance of (negative) electrons in matter with which they may unite. But why are *electrons* so abundant? Perhaps because there are so many nuclei, built out of neutrons and positive protons, which serve to bind the electrons into atoms. Then why are there no *negative protons* which might similarly bind positrons into atoms? Thus does each discovery in physics raise fresh questions, answers to which it is left for future physicists to discover.

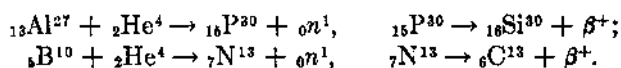
213. The Discovery of Induced Radioactivity.—Almost immediately after Anderson announced his discovery of positrons, it was reported by several experimenters that positrons are emitted from many targets when bombarded by high-speed ions. It was supposed at the time either that the positrons were produced as a direct effect of the bombardment (which is now believed never to occur), or that the bombardment gave rise to γ -rays and these produced pairs of electrons and positrons.

While studying with an ionization chamber the positrons emitted from aluminum or boron when bombarded with α particles from polonium, Curie-Joliot and Joliot made the important discovery that

¹ E.g., by KLEMPERER, *Cambridge Phil. Soc. Proc.*, vol. 30, p. 347 (1934); CRANE and LAURITSEN, *Phys. Rev.*, vol. 45, p. 430 (1934).

² GRAY and TARRANT, *Roy. Soc. Proc.*, vol. 136, p. 662 (1932); vol. 143, pp. 681, 706 (1933).

the emission of positrons did not cease at once when the polonium was removed.¹ The positron activity was found to decay exponentially, as does the activity of a radioactive substance, with a period (or time required for a decrease to half) of 3 min. 15 sec. in the case of aluminum, or 14 min. in the case of boron. It was concluded that the α particles had transformed atoms of aluminum or boron into atoms of a radioactive isotope of some other element, which then eventually disintegrated into positrons and stable atoms of some sort. The emission of a positron will decrease the atomic number Z by unity, just as, on the contrary, the emission of a negative β -ray raises it by unity. The following reactions were suggested, β^+ denoting a positron:



Here Si^{30} and C^{13} are stable isotopes known to occur in nature, but P^{30} and N^{13} are not, only P^{31} , N^{14} , and N^{15} being found. In these cases, the primary reaction results also in the ejection of a neutron.

Now this explanation can at once be put to a chemical test; for P^{30} should be chemically separable from Al, and N^{13} from B. Curie and Joliot made such tests for the aluminum reaction as follows. Aluminum which had been thoroughly activated by α -ray bombardment was dissolved in HCl and the solution evaporated to dryness, within the space of 3 min., and then the residue was tested for positron activity. It showed none. The activity had presumably gone off in the gas that was evolved. It is known that phosphorus treated in this manner would pass off as the gas PH_3 . The process was then repeated, but this time the gas itself was caught, in a tube with walls sufficiently thin to transmit the positrons. Most of the activity was then found in the gas. As a second test, the aluminum was dissolved in aqua regia, without evolution of gas, and a compound of zirconium was added which would precipitate phosphorus as zirconium phosphate, the aluminum remaining in solution; the positron activity was now found in the precipitate but not in the solution. These tests showed that the substance emitting the positrons was definitely not aluminum and was almost certainly an isotope of ordinary phosphorus. Analogous tests identified the substance produced from boron as nitrogen.

The discovery of induced radioactivity advanced nuclear work enormously, because it opened up a new possibility of establishing the chemical nature, *i.e.*, the value of the atomic number Z , of the products of nuclear reactions. If these products are radioactive, they can be followed in chemical reactions by means of their radioactivity; whereas

¹ IRENE CURIE and F. JOLIOT, *Comptes Rendu*, vol. 198, pp. 254, 408, 559 (1934).

otherwise they cannot be located and their chemical nature remains necessarily an inference from indirect evidence, such as considerations of mass and energy. It is particularly easy to demonstrate that a given radioactive substance is *not* isotopic with a given element; it is only necessary to mix the two and to show that they precipitate in different proportions, or otherwise behave differently, in some one chemical reaction.

As soon as Curie-Joliot and Joliot announced their discovery, the study of induced radioactivity was taken up in many other laboratories, and new cases of it were announced at a rapid rate. Many examples will be described in later sections of this chapter. Before discussing nuclear reactions further, however, it will be advantageous to describe next the principal methods that have been developed for the production of high-speed ions to be used in bombarding nuclei. The various lines of fundamental nuclear research during the decade from 1930 to 1940 were so interwoven that they are hard to separate.

214. Artificial Transmutation by High-velocity Protons.—It was soon realized that it would be advantageous to use artificially produced particles of high velocity instead of α particles for the production of nuclear transmutations, in order to free experimentation from the limitations incident to the use of radioactive sources. This was first done in the pioneer experiments of Cockcroft and Walton, in which nuclear transmutations were first produced by means of artificially accelerated ions.¹

The high potential for use in these experiments was produced by means of a combination of condensers and rectifiers.² The method has analogies with the old scheme of charging a number of condensers in parallel, insulating them, and then connecting them in series so as to add up their potentials; but here all connections are made electrically. Two columns of condensers are used, cross connected with rectifiers in such a way that, when the potential of either column is raised sufficiently relative to the other by means of a potential difference applied at their bases, electrical charge flows across through the rectifiers to the other column. Thus, by using an alternating potential difference, charges are pumped up to the tops of the columns. Potentials as large as 700,000 volts could be produced in this manner, using an applied voltage of a quarter as much supplied by a high-tension transformer.³

¹ COCKCROFT and WALTON, *Roy. Soc. Proc.*, vol. 137, p. 229 (1-932).

² Cf. COCKCROFT and WALTON, *Roy. Soc. Proc.*, vol. 136, p. 619 (1932); GREINACHER, *Zeits. f. Physik*, vol. 4, p. 195 (1921).

³ For later developments, see MANLEY, HAWORTH, and LUEBKE, *Rev. Sci. Instruments*, vol. 12, p. 587 (1941).

In this way Cockcroft and Walton were able to bombard a number of the lighter elements with protons having energies up to 500,000 electron-volts, the proton current being of the order of 10^{-6} ampere. Under such bombardment, a number of the elements emitted α particles, the process being thus the converse of bombardment by α particles. When lithium, for example, was bombarded by protons of 250 electron-kilovolts equivalent energy, α particles were observed whose range in air was about 8 cm., corresponding to an energy of some 8.6×10^6 electron-volts. Increasing the bombarding energy up to 500 electron-kilovolts *did not appreciably alter the range of the emitted α particles*. The α particles were observed both by means of scintillations and with an electrical counter [Sec. 198(a),(6)]; later their tracks were observed by Dee and Walton in a cloud chamber.¹

These results were explained by assuming that the proton penetrates the Li^7 nucleus, after which the product nucleus thus formed breaks up into two α particles, the net reaction being



If this interpretation is correct, a simple conclusion can be drawn concerning the directions of projection of the α particles. Their momentum is so large that the momentum of the incident proton can

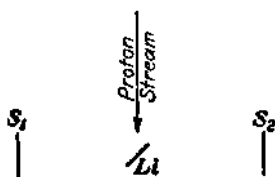


FIG. 204.

be neglected; it follows then from the conservation of momentum that the two α particles must be projected with approximately *equal velocities in opposite directions*. Cockcroft and Walton confirmed this conclusion by placing a thin lithium target at 45° to the stream of incident protons as shown in Fig.

204. They found that in a large number of cases scintillations appeared *simultaneously* on the screens S_1 and S_2 .

The *energy relations* in this reaction are particularly interesting because the energy of the α particles so greatly exceeds that of the incident protons. The excess energy must be derived from the rest energy of the Li^7 nucleus. As an energy equation we may write

$$\text{Li}^7 + \text{H}^1 + E(\text{H}^1) = 2\text{He}^4 + E(2\text{He}^4),$$

in which isotopic symbols have been written to represent rest energies and $E(\text{H}^1)$ and $E(2\text{He}^4)$ stand for the kinetic energies of the proton and of the α particles, respectively. Or, following recent custom, we may write

$$\text{Li}^7 + \text{H}^1 = 2\text{He}^4 + Q \quad (324)$$

¹ DEE and WALTON, *Roy. Soc. Proc.*, vol. 141, p. 733 (1933).

where $Q = E(2\text{He}^4) - E(\text{H}^1)$ and represents the *heat of reaction* or net amount of energy released by the reaction, in the form of kinetic energy in the present case. To test the equation, we have, from modern data, using Eq. (318):

	Mev
Rest energy of Li^7 (7.0180 a.m.u.)	 6540.8
Rest energy of H^1 (1.0081 a.m.u)	939.6
Kinetic energy of H^1	0.25
Rest energy of He^4 (4.0030 a.m.u)	3731.6
Kinetic energy of each He^4	8.6

Thus from the kinetic energies

$$Q = 2 \times 8.6 - 0.25 = 10.95 \text{ Mev};$$

whereas, from the masses and Eq. (324),

$$Q = \text{Li}^7 + \text{H}^1 - 2 \text{He}^4 = 17.2 \text{ Mev}.$$

The agreement is within the experimental error. Thus we are dealing here not only with a transmutation of the elements but also with a genuine release of energy out of the lithium atom, at the expense of its ordinary measured mass.

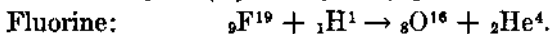
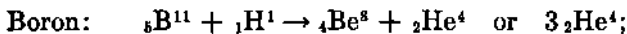
In this calculation we have used masses of the *neutral atoms*, whereas the equation refers, strictly speaking, to the *bare nuclei*. The balance of the atomic numbers in the equations implies, however, a balance in the numbers of circumnuclear electrons as well; hence, errors due to the inclusion of the electronic masses in the values used for nuclear masses cancel out.

To make such calculations significant, it is obvious that atomic masses must be known with high precision. Many of them are. The error in the mass-spectrographic values cited here is supposed to be less than 0.0002 in the case of Li^7 and less than 0.00007 for each of the others. If these errors all had the value stated and the same sign, the total error would be 0.0005 or about 0.5 Mev. It is likely, however, that the error in Q as calculated from the masses would not exceed 0.2 Mev, which is about the difference between the two values of Q as found above.

The kinetic energy of the proton is so small in this case that variation in its energy should make little difference in the range of the α particles, as was actually found to be the case.

Cockcroft and Walton also tested the elements Be, B, C, O, F, Na, Al, K, Ca, Fe, Co, Ni, Cu, Ag, Pb, U and observed an emission of α particles in most cases. Later work has shown, however, that in many cases the weak effect observed was really due to an impurity in the

target. The following nuclear reactions proposed by Cockcroft and Walton appear to have been confirmed:



Thus, in the case of boron, the product is sometimes three α particles, sometimes an α particle and a Be^8 nucleus.

For *fluorine*, the *energy* equation can be written

$$\text{F}^{19} + \text{H}^1 - \text{O}^{16} - \text{He}^4 = Q,$$

$$19.0045 + 1.0081 - 16 - 4.0039 = 0.0087 \text{ a.m.u.},$$

corresponding to

$$Q = 8.1 \text{ Mev.}$$

Here Q represents the gain in kinetic energy due to the reaction, or the kinetic energy of O^{16} and He^4 minus that of H^1 . According to later, more precise observations,¹ the α particles produced by this reaction have a range of 6.95 cm. when released by protons accelerated in a cyclotron (Sec. 215) to an energy of 1.63 Mev. The kinetic energy given to the O^{16} nucleus was not measured, but allowance for it can easily be made in the following manner.

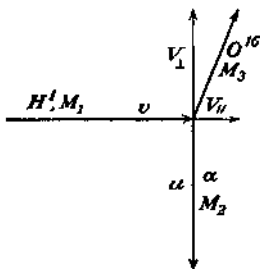


FIG. 205.

In the later experiment just mentioned, the α particles were observed in a direction at right angles to the proton beam; and under such circumstances a complete dynamical analysis of the collision is easily made. Let M_1 , M_2 , M_3 denote the masses of the proton, α particle, and residual nucleus, O^{16} , respectively. The O^{16} nucleus moves off obliquely

(Fig. 205); let us resolve its velocity into two components, $V_{\parallel}$ parallel to the proton velocity v , and $V_{\perp}$ opposite to the velocity u of the α particle. Then by conservation of momentum

$$M_1 v = M_3 V_{\parallel}, \quad \therefore V_{\parallel} = \frac{M_1 v}{M_3};$$

$$M_2 u = M_3 V_{\perp}, \quad \therefore V_{\perp} = \frac{M_2 u}{M_3}.$$

Thus

$$Q = \frac{1}{2} M_2 u^2 + \frac{1}{2} M_3 (V_{\parallel}^2 + V_{\perp}^2) - \frac{1}{2} M_1 v^2 = \frac{1}{2} M_2 u^2 \left(1 + \frac{M_2}{M_3} \right) - \frac{1}{2} M_1 v^2 \left(1 - \frac{M_1}{M_3} \right)$$

¹ HENDERSON, LIVINGSTON, and LAWRENCE, *Phys. Rev.*, vol. 46, p. 38 (1934).

or

$$Q = \left(1 + \frac{M_2}{M_1}\right) E_2 - \left(1 - \frac{M_1}{M_2}\right) E_1, \quad (325)$$

where $E_1 = M_1 v^2/2$ and $E_2 = M_2 u^2/2$, representing the respective kinetic energies of the incident proton and of the ejected α particle.

For 6.95-cm. α particles, we have $E_2 = 7.70$ Mev.¹ Substituting in (325) this value and $E_1 = 1.63$ Mev and the approximate masses $M_1 = 1$, $M_2 = 4$, $M_z = 16$, we find

$$Q = 8.10 \text{ Mev.}$$

This agrees excellently with the value of Q calculated above from the atomic masses (8.11). Every successful check of this sort not only confirms the particular nuclear reaction that has been assumed to occur but also supports once more the principle of the close relation between mass and energy.

The effects of the bombardment of fluorine with protons are actually more complex, however, than is indicated here (see Sec. 226).

215. The Production of High-speed Ions.—Many methods have been invented for the generation of ions moving at high and adjustable velocities.²

In some methods the ions are accelerated to their final velocities in a long vacuum tube, to the terminal electrodes of which there is applied a potential difference of the necessary magnitude. Mention has already been made of the condenser-rectifier cascade used by Cockcroft and Walton for the production of large potential differences (Sec. 214). Transformers with their secondary windings connected in series have also been employed for the production of voltages up to 1 million volts; to avoid having a potential difference of this magnitude between the primary and secondary of a single transformer, part of the secondary winding of each transformer is used to supply voltage for the primary of the next, only the first transformer being supplied from an independent source. A disadvantage of this method is that the final voltage produced is necessarily subject to fluctuations whose frequency is that of the alternating current.

(a) *The Electrostatic Generator*.—A very steady high potential is furnished by the electrostatic generator as developed by Van de

LIVINGSTON and BETHE, *Rev. Modern Phys.*, vol. 9, p. 245 (1937). Plot on p. 266.

² For a more complete discussion than can be given here see Livingston and Bethe, *Rev. Modern Phys.*, vol. 9, p. 245 (1937), also two articles in *J. of Applied Phys.* vol. 9 (1938), one by Wells on p. 677 and another by Kurie on p. 691.

Graaff.¹ The principle is illustrated in Fig. 200. An endless belt B is caused to run rapidly over two pulleys placed one above the other. Electric charge from a generator G is sprayed onto the lower end of the rising half of the belt by means of corona discharge from the points of a comb C . At the top, the belt enters a spherical terminal T , to which the charge is transferred from the belt by means of another comb C' ; or, still better, charge of the opposite sign is sprayed onto the belt, supplied by another generator G inside the terminal. In this way the charge on the terminal is continually built up, the potential attained being limited only by the rate at which it leaks off. A discharge tube can then be connected between the terminal and the ground, or between the oppositely charged terminals of two such generators standing side by side.

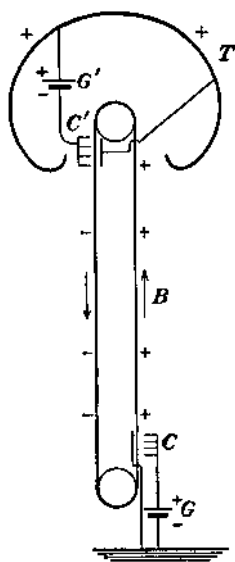


FIG. 206.—Diagram illustrating the principle of the Van de Graaff generator.

The Van de Graaff generator constructed by the Massachusetts Institute of Technology yields currents of 4 milliamperes at 2.7×10^6 volts.² A similar generator at the Westinghouse Research laboratories yields a current of comparable magnitude at voltages up to 3.7×10^6 volts, the voltage holding steady within $\frac{1}{2}$ percent for several minutes at a time.³ This latter generator is surrounded by air at 120 lb. in⁻² pressure, in order to increase the insulation; it is enclosed within a vessel 47 ft. high, made of steel $\frac{5}{8}$ to $1\frac{3}{4}$ in. thick and shaped like a pear standing on its smaller end. Two belts $31\frac{1}{2}$ ft. long and 18 in. wide are used, running at speeds of about 80 ft. per sec. The construction of such an instrument presents many technical problems and well illustrates the elaborate character of the apparatus that is needed for much of modern physical research. To smash a nucleus 10^{-14} in. in diameter, the physicist constructs a machine standing 50 ft. high!

The *vacuum tube* used for the acceleration of ions by means of high potentials is commonly made in sections, and inside it a series of insulated metal cylinders is mounted with short gaps between them

¹VAN DE GRAAF, *Phys. Rev.*, vol. 38, p. 1919 (1931); vol. 43, p. 149 (1933); L. C. VAN ATTA, NORTHROP, C. M. VAN ATTA, VAN DE GRAAFF, *Phys. Rev.*, vol. 49, p. 761 (1936).

²*Phys. Rev.*, vol. 57, p. 563 (1940); *Rev. Sci. Instruments*, vol. 12, p. 534 (1941).

³*Phys. Rev.*, vol. 58, p. 162 (1940).

(Fig. 207). The ions travel along the axis of these cylinders and the electric field in the gaps exerts a focusing action tending to prevent spreading of the beam away from the axis. In the Westinghouse installation just mentioned, the accelerating tube is about 25 ft. long and consists of 130 sections. The necessity of constructing an accelerating tube long enough to withstand the total potential difference is a drawback to any of the methods so far described for the production of high-speed ions.

(b) *The Cyclotron*.—A radically different principle is utilized in the well-known cyclotron invented by Lawrence and Livingston.¹ Instead of attempting to produce a high potential, they accelerate the ions by means of many *successive impulses*. Use is made of the fact that a charged particle moving not too rapidly in a uniform magnetic field revolves in a circle with an angular velocity depending only upon its ratio of charge to mass and upon the strength of the field, but *not upon its velocity*.

The accelerating system of a cyclotron is shown diagrammatically in Fig. 208. *BB* represents the horizontal cross section of a flat circular box which can be evacuated. Above it and below it, and only a few inches apart, are the pole pieces of a powerful electromagnet, not shown in the figure; this magnet produces a nearly uniform field of 10,000 to 20,000 oersteds. Inside the box *BB* are two hollow, semicircular electrodes often called "dees," such as might be formed by cutting another circular box in two along the diameter *EE* and slightly separating the halves. Between the dees, a rapidly alternating potential difference, of from 3 to 9 kilo volts at 8 to 15 megacycles, is produced by means of an electrical oscillator.

To produce ions, a suitable gas is introduced into the box at a pressure of 10^{-2} to 10^{-3} mm. Hg, and electrons are shot into this gas at the center of the dees (*C*), from a filament placed just above the dees and maintained at a negative potential relative to the mean potential of the dees. A positive ion thus produced at *C* will be drawn into whichever dee happens to be negative at the moment and will move along a semicircular path in the space inside the dee, where



FIG. 207. Simplified and shortened diagram of an accelerating tube for use with high potentials, shown in longitudinal section; *I*, ion source; *T*, target.

¹ LAWRENCE and LIVINGSTON, *Phys. Rev.*, vol. 45, p. 608 (1934); LAWRENCE and COOKSEY, *Phys. Rev.*, vol. 50, p. 1131 (1936). Cf. KJRIE, *J. of Applied Phys.*, vol. 9, p. 691 (1938).

there is no appreciable electric field. If the frequency of the oscillator and the magnetic field strength are properly adjusted, the ion will arrive again at the crack between the dees just at the moment when the potential difference between them has the opposite sign and will be accelerated further. The radii of the successive semicircular paths increase with the speed of the ion, but the time required to traverse a semicircular remains constant and equal to the half-period of the oscillator. Finally the ion, after 50 to 100 revolutions, is traveling near the periphery of the dees and enters the space between two deflecting plates *PP*, where an electric field deflects it slightly outward so that it passes out of the box *BB* through the window *W*. Or, instead of the window, a target may be inserted at this point to be bombarded by the ions.

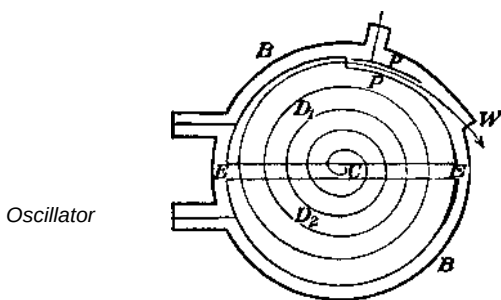


FIG. 208.—Diagram of the essentials of a cyclotron, except for the magnet, whose pole pieces would be situated just above and just below the box *B*.

Naturally a good many accessories are required, and a cyclotron is a large, complicated, and expensive instrument. The dees and pole pieces are commonly 30 in. or more across, so that many tons of iron and copper are required for the water-cooled electromagnet. With such a cyclotron a current to the target of 10^{-5} amp. of deuterons at 5 million volts is readily secured. From the 60-in., 220-ton cyclotron at the University of California a current of 10^{-5} amp., consisting of deuterons of 16 million volts energy, has been brought out into the air, where it produced a blue steamer nearly 5 ft. long.¹ In a cyclotron constructed later at the University of Illinois² the beam was made to emerge radially and could be resolved according to the energy of the emergent electrons by means of a magnetic analyzer. With this equipment a total beam of 100 microamperes could be secured, mono-energetic within 2 percent (*i.e.*, the energies of the electrons in the

LAWRENCE et al., *Phys. Rev.*, vol. 56, p. 124 (1939); photograph in an article by Mann, *Nature*, vol. 143, p. 583 (1940).

⁸ See *Rev. Set. Instruments*, vol. 15, p. 333 (1944).

beam varied at most by 2 percent); or, for experimental purposes, a beam density of 0.001 microampere per square millimeter could be obtained, monoenergetic within 0.1 percent.

A limit to the ionic energies that can be produced with a cyclotron will probably be set, ultimately, by the relativistic variation of mass with velocity, which tends to throw the ion out of step with the alternating potential on the dees. A proton with a kinetic energy of 10 Mev has a velocity, calculated nonrelativistically, of

$$\sqrt{\frac{2eV}{m}} = \left(\frac{2 \times 4.8 \times 10^{-10} \times 10^7}{300 \times 1.66 \times 10^{-24}} \right)^{1/2} = 4.4 \times 10^9 \text{ cm. sec.}^{-1}.$$

At this speed the relativistic correction on the mass is already

$$\frac{v^2}{2c^2} = \frac{(4.4/30)^2}{2} = 0.011 \text{ or 1 percent.}$$

For this reason the cyclotron cannot be used effectively for electrons.

(c) *The Induction Accelerator.*—More recently, a device has been invented that is completely free from the variable-mass difficulty and is admirably adapted for the acceleration of electrons. In this device, called an "induction accelerator" (or betatron), use is made of the electric field that is associated with a changing magnetic field. When the field is axially symmetric, like that between the closely spaced poles of an ordinary magnet, the electric lines are circles coaxial with the field. If a charged particle could be made to follow one of these circular lines, in going around many times it would acquire a high energy.

First attempts to utilize this principle were not successful, but in 1940 Kerst showed¹ how to hold the particle on its circular path by means of the magnetic field itself. Let R be the radius of the path, m the mass of the particle, q its charge in *electromagnetic unite*, p its momentum, and H the instantaneous strength of the magnetic field. Then, to keep the particle on the path, it is necessary, according to Eq. (343a) in Sec. 236, where the charge is in *electrostatic* units, that

$$p = qRH.$$

On the other hand, if the particle starts from rest at the instant at which the field begins to increase from zero, the angular momentum acquired by the particle up to the time at which the flux through the

¹ KERST, *Phys. Rev.*, vol. 58, p. 841 (1940); *Rev. Set. Instruments*, vol. 13, p. 387 (1942); *Nature*, vol. 157, p. 90 (1946). See also BLEWETT, *Phys. Rev.*, vol. 69, p. 137 (1946).

circle has attained the value φ is given by the equation

$$pR = \frac{1}{2\pi} q\varphi;$$

this equation can be obtained from Eq. (224) in Sec. 144 by replacing $AGIH$ by pR and noting that, in Eq. (224) and in the argument leading up to it, $\pi a^2 H$ must be replaced by φ in case H is not assumed to be uniform in space. Elimination of p between the two equations thus obtained gives

$$\varphi = 2\pi R^2 H. \quad (326)$$

If the field can be so designed that Eq. (326) is at all times satisfied for a certain circle of radius R , then a charged particle of any mass, once started on this circle, will go round and round indefinitely, so long as the magnetic field continues to increase. Variation of mass with velocity has no effect. The flux through the circle must be twice what it would be if the field were uniform and equal to H , or to its strength at points on the circle; this condition is easily met by suitably shaping the pole pieces of the electromagnet that furnishes the field. The motion around the circular orbit can also be shown¹ to be stable, provided the orbit lies in the plane of symmetry between symmetrical pole pieces, partly because of the direction in which the magnetic lines curve. This stability of the orbit is a very fortunate circumstance.

In practice, the magnet is constructed of laminated iron so that it can be activated by alternating current. An enclosed space between the pole pieces is highly evacuated, and charged particles are projected into this space at a point near the stable orbit, in a tangential direction. The particles enter in a spurt at a suitable instant just after the magnetic field has passed through zero, and then settle onto the stable orbit and circle round and round it, gaining energy. At a chosen point in the quarter cycle, either H at the location of the orbit or φ is suddenly changed by sending current through an auxiliary winding; the particles then spiral either outward or inward and strike a target, or perhaps issue through a thin window into other apparatus.

In a large induction accelerator built in the Radiation Laboratory of the General Electric Company at Schenectady, New York,² the stable orbit is 66 inches in diameter; the entire machine weighs 130 tons. A 60-cycle frequency is employed, and H rises to 4,000 oersteds on the stable orbit. Electrons may be injected at 30 to 70 kilovolts

¹ KERST and SERBER, *Phys. Rev.*, vol. 60, p. 53 (1941).

² WESTENDORP and CHARLTON, *J. of Applied Phys.*, vol. 16, p. 581 (1945).

velocity; if left on the orbit throughout an entire quarter cycle, they go around about 250,000 times, almost at the speed of light, and acquire an energy of a hundred million electron-volts.

To obtain higher energies, an even larger machine will probably be needed, since the time during which the acceleration acts is necessarily limited.

By allowing electrons accelerated in such a device to strike a target, X-rays may be generated whose energy varies continuously over the range that includes all known γ -rays and on upward to 10^8 ev. Thus this entire range of frequencies is opened up for investigation. X-rays and γ -rays of the same frequency are identical in nature and properties, of course, but it is convenient to employ the term "X-rays" when they are produced by the impact of electrons on a target and to reserve the term "r-rays" for electromagnetic radiation emitted by the nucleus.

(d) *The Synchrotron*.¹—It has also been proposed to adapt the cyclotron to the acceleration of electrons by introducing the electrons in successive spurts and allowing the magnetic field to increase in strength in proportion as the inertial mass of each spurt increases, thereby preserving synchronism with the alternating electric field on the dees. The accelerating effect of the superposed electric field due to the changing flux may be relatively unimportant, or it may be utilized to provide an appreciable part of the accelerating action.²

216. Experiments with Neutrons.—The discovery of neutrons by Chadwick in 1932 was followed at once by numerous other investigations.³ Neutrons are much harder to work with than charged particles, since they do not interact with the electrons in atoms and so produce no ions. For this reason their penetrating power is enormous, as we have seen, and they can be detected only by indirect means. It has been necessary to invent novel methods for handling them. In this section, a short account will be given of the properties of neutrons as revealed by experiment.

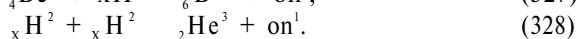
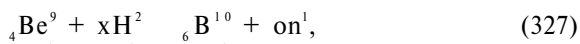
(a) *Sources of Neutrons*.—The original reaction discovered by Chadwick is still widely utilized for the production of neutrons. Radon gas obtained from radium is mixed with finely divided beryllium, from which neutrons are liberated by the α particles from the

¹ MCMILLAN, *Phys. Rev.*, vol. 68, p. 143 (1945); VEKSLER, *Jour. Phys. U.S.S.R.*, vol. 9, p. 153 (1945).

² POLLOCK, *Phys. Rev.*, vol. 69, p. 125 (1946). See also DENNISON and BERLIN, *Phys. Rev.*, vol. 70, p. 58 (1946); BOHM and FOLDY, *Phys. Rev.*, vol. 70, p. 249 (1946).

³ Cf. RASETTI, *op. cit.*, pp. 233-267; a summarizing article by Amaldi and Fermi, *Phys. Rev.*, vol. 50, p. 899 (1936); and many articles in the literature.

radon. Such a source lasts only a few days, however; and radon also emits α -rays, which may be a source of inconvenience. Hence, other sources of α particles are often employed instead of radon, such as polonium, which emits no γ -rays and has a period of 136 days. Another good method of obtaining neutrons is by bombarding a light element with deuterons from a cyclotron. The most useful target for this purpose is either beryllium or a substance containing deuterium, such as "heavy" ice or heavy paraffin, the respective reactions being



The $\text{H}^2 + \text{H}^2$ (or $d - d$) reaction requires unusually low voltages; 200,000 volts is enough. A stream of 200 microamperes of deuterons at 300,000 volts is readily produced;¹ acting on heavy ice, such a stream is equivalent to the α -rays from 10 grams of radium falling upon beryllium. When 500,000-volt deuterons act on heavy ice, 2×10^7 neutrons per second are produced for each microampere of deuterons absorbed in the target. At higher voltages, however, the beryllium target is superior; from it, a microampere of 1 million-volt deuterons yields 17×10^7 neutrons per second, or 2.7×10^{-5} neutron for each deuteron absorbed. This is about the same yield per particle as in the production of neutrons from beryllium by bombardment with α particles. In any case, as judged by the usual standards, the efficiency of all known neutron sources is very low.²

(b) *The Measurement of Neutron Intensities.*—There are several different methods by which neutrons may be detected and counted.

We may count the recoil protons or other recoil nuclei produced by them in matter; or, if the recoil nuclei are produced in a gas, the ionization caused by them may be measured in an ionization chamber.

Another method is to utilize one of the many nuclear transmutations which neutrons have been found to cause. If such a transmutation occurs in a cloud chamber and results in the ejection of a charged particle from a nucleus, a V-shaped double track will be seen, caused by the charged particle and the residual nucleus, the neutron itself producing, of course, no track (Fig. 209). Or, ionization caused by the products of the reaction may be measured. Ionization chambers lined with lithium or boron, or filled with BF_3 gas, are often used, especially for slow neutrons, the ionization being due to α particles released in one of the two reactions,



¹ MANLEY, HAWOKTH, and Lt JEBKE, *Rev. Sci. Instruments*, vol. 12, p. 587 (1941).

² Neutron sources are discussed in *Zeits. f. Physik*, vol. 42, p. 91 (1942).

The nucleus H^3 that is formed in the first of these reactions is slightly radioactive, decaying to He^3 .

Finally, Fermi and others have often employed as detectors of neutrons some substance, such as silver, which is made radioactive by them, the induced β -activity of the detector being measured with an ionization chamber.

* Some of the methods of producing and detecting neutrons are illustrated in Fig. 210.

(c) *Neutron velocities* are often determined by measuring the ranges of recoil protons or other nuclei produced by them, as in the pioneer experiments of Chadwick, described in Sec. 205. The proton

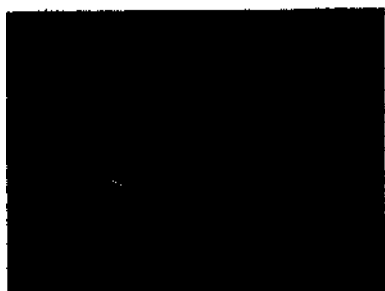


FIG. 209.—Disintegration of a nitrogen nucleus by a neutron, presumably coming from below. Tracks of the ejected particle and of the recoil nucleus are visible. {*Feather, Roy. Soc. Proc.*, vol. 136, p. 709, 1932).

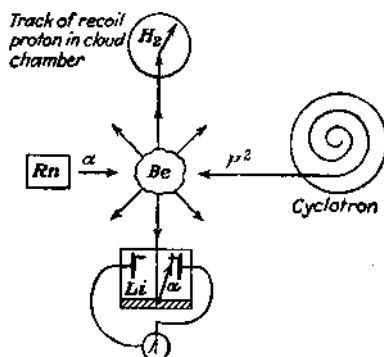


FIG. 210.—Diagrammatic representation of various methods of producing and detecting neutrons.

ranges can be measured either with an ionization chamber or in a cloud chamber.

The neutrons projected from a given target bombarded with a homogeneous parallel beam of ions have different velocities according to the direction in which they leave the target. Relative to the composite nucleus that is first formed by the capture of a bombarding ion, the neutrons should all have the same velocity; but this nucleus is given a forward component of motion by the impact. Hence, the neutrons that are projected in the same direction as the bombarding ions have the highest velocity, those projected in the opposite direction the lowest. For experimental purposes it is customary to utilize those neutrons which are projected at right angles to the ion beam, and values of neutron energies are commonly quoted for this direction of emission.

As an example of the results, the neutrons emitted by the " $d - d$ " ($H^2 + H^2$) reaction have been found to form a single, nearly homogeneous group¹; presumably if the bombarding deuterons all had exactly the same energy, so would the neutrons (in a given direction), in accordance with the equation of nuclear energy balance. Neutrons from the Rn — Be reaction, on the other hand, are found to consist of several groups, with a maximum around 4.8 Mev but ranging up to 13 Mev.

(d) *Transmutations Caused by Neutrons.*—Whereas charged particles require high energies in order to penetrate heavy nuclei against electrostatic repulsion, neutrons, carrying no charge, ought to be able to penetrate all nuclei with ease. They ought, therefore, to be especially effective agents for the transmutation of atoms. This expectation is borne out by the facts. Particularly interesting is the induced radioactivity produced in many substances by neutron bombardment, which was first observed by Fermi in 1934.² The induced activity consists of the emission of negative electrons or, less often, of positrons, accompanied sometimes by γ -rays. The period of decay varies greatly from element to element. Examples of such reactions will be given later (Sees. 217 and 220).

(e) *Slow Neutrons.*—The important observation was made accidentally by Amaldi, Fermi, and others³ that the activity induced by neutrons in silver; rhodium, and many other elements was greatly enhanced when the neutron source and the target were surrounded by material containing hydrogen, such as paraffin or water, or when such material was merely placed between the source and the target. Fermi explained this effect by assuming (1) that in the hydrogenous material the neutrons were slowed down as the result of successive impacts with hydrogen nuclei and (2) that many substances absorb slow neutrons much more strongly than they absorb fast ones.

The assumptions have been confirmed by later work. Neutrons have come to be called "slow" when their energies are less than about 1,000 electron-volts, otherwise they are "fast." A neutron will lose energy upon colliding with any nucleus that is moving with smaller momentum, but the loss is especially large in collisions with protons because of their relatively small mass. It can be shown that neutrons colliding at random with protons whose velocities are much smaller

¹ Cf. HUDSPETH and DUNLAP, *Phys. Rev.*, vol. 57, p. 97 (1940).

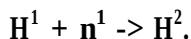
² Cf. FEKMI, AMALDI, D'AGOSTINO, RASETTI, and SEGRE, *Roy. Soc. Proc.*, vol. 146, p. 483 (1934), and references there given.

³ AMALDI, D'AGOSTINO, FERMI, PONTECORVO, RASETTI, and SEGRE, *Roy. Soc. Proc.*, vol. 149, p. 522 (1935).

will execute on the average about one collision for each reduction of their energy in the ratio $1/e$, where $e = 2.718$ standing for the Napierian base. Thus, a neutron entering hydrogenous material at room temperature with an energy of 5 Mev, provided it does not combine with some nucleus, will be reduced to the energy of a molecule at room temperature, or 0.025 ev, after some 20 collisions (since $\log, (5 \times 10^6/0.025) = 19.1$); thereafter its energy will oscillate about this latter value. Neutrons which have thus come into thermal equilibrium with matter are called *thermal neutrons*.

The mean free path of neutrons between elastic collisions in paraffin or water has been inferred to be of the order of 5 cm. at 5 Mev, but it rapidly decreases as the speed of the neutrons is diminished until for a thermal neutron it is only 3 mm. Thus, neutrons that have passed through 10 cm. of paraffin will mostly be thermal neutrons. That their velocities really are only of thermal magnitude has been shown by mechanical methods of observation.¹

There is, to be sure, a slight chance that a neutron may actually combine with a proton during the slowing-down process, according to the reaction



In this reaction the excess energy is radiated away as γ -rays, which have been detected. The probability of capture increases as the velocity of the neutron decreases, but even at thermal velocities the neutron will execute many elastic collisions with protons before being captured.

The motion of thermal neutrons in water may be studied by releasing short bursts of them at a certain point and measuring the number that arrive at various points in the water.² In this way the mean life of a free thermal neutron in water is found to be about 2.0×10^{-4} sec. Hence, moving at an average speed of about 2.2×10^5 cm. per sec, and with a mean free path of 0.3 cm., it undergoes on the average about $2.0 \times 10^{-4} \times 2.2 \times 10^5/0.3$ or 150 collisions with a water molecule before being captured. The point of capture is located on the average about 3 cm. away from the point at which the neutron originated; this distance has been called the "diffusion length" of the neutron.³

¹ DUNNING, PEGRAM, FINK, MITCHELL, and SEGRE, *Phys. Rev.*, vol. 48, p. 704 (1935); RASETTI, SEGRE, FINK, DUNNING, and PEGRAM, *Phys. Rev.*, vol. 49, p. 104 (1936).

² MANLEY, HAWORTH, and LUEBKE, *Phys. Rev.*, vol. 61, p. 152 (1942).

³ GAMERTSFELDER and GOLDBABER, *Phys. Rev.*, vol. 61, p. 556 (1942).

Presumably the ultimate fate of all neutrons is to be captured by some nucleus. Probably all nuclei are capable of picking up an extra neutron, with an evolution of energy in some form; for inspection shows that successive isotopes, so far as is known, increase in mass from one to the next by less than the mass of a free neutron, indicating a loss of rest mass and energy as an additional neutron is bound in the nucleus.

Slow neutrons are especially interesting because for them the probabilities of capture in going a given distance are relatively large.

(f) *Scattering and Absorption of Neutrons.*—A neutron passing close to a nucleus may be either captured or merely "scattered," *i.e.*, deflected and perhaps changed in energy as if by an elastic collision. The probability of either event is commonly expressed in terms of an equivalent *cross section*. The *scattering cross section* of a nucleus for neutrons can be defined as follows: Imagine a uniformly distributed beam of neutrons, all moving in the same direction and at the same speed, to pass over the nucleus. A certain number of them will be deflected or "scattered" as the result of an elastic collision with the nucleus. This number is the same as the number which, according to classical conceptions, would cross an area of a certain size drawn on a plane perpendicular to the beam. The latter area is the cross section for scattering of neutrons by the nucleus in question. In a similar way, a *cross section for capture* of a neutron by a given nucleus, denoted by σ_c , can be defined. If a large number N of nuclei are placed in a uniform beam in which n neutrons cross unit area per second, then $nN\sigma_c t$ nuclei will capture a neutron during a time t .

The *scattering cross section* for neutrons is found to increase with the atomic number of the nucleus, but to vary only moderately (in general) with neutron speed. Very roughly, it may be said that about 5 cm. of any solid or liquid material will scatter half of the neutrons out of a rapidly moving beam. Slow neutrons are scattered more readily than fast ones, especially by substances containing hydrogen.¹

Furthermore, scattering may be either elastic or inelastic. In elastic scattering, energy is transferred from the neutron to the nucleus only in the form of kinetic energy, and, except for very light nuclei, only in small amounts. In inelastic scattering, on the other hand, the nucleus is also raised into an excited state of higher energy. These two types of scattering can be distinguished either by measuring the energy of the scattered neutrons or by detecting the γ -rays that are

¹ For values of the scattering cross section see Volz, *Zeits. f. Physik*, vol. 121, p. 201 (1943).

emitted as the nucleus returns to a state of lower energy. Such 7-rays have probably been observed in some instances.¹

Neutrons scattered inelastically by magnesium have been observed directly by Little, Long, and Mandeville.² The incident neutrons had an energy of 2.50 Mev; but in the scattered beam there was a considerable component having energies of about 1.08 Mev. After various corrections had been made it was concluded that the associated excited level of the Mg nucleus lies 1.30 Mev above the normal level. Values for the lowest excited level as obtained from 7-ray observations have ranged from 1.31 to 1.49 Mev. It was estimated that the nuclear cross-section for the elastic part of the scattering was about 1.6×10^{-24} cm.², and for the inelastic part, 0.6×10^{-24} cm.²

In iron, finally, the scattering of neutrons appears to arise in part from an interaction between magnetic fields and *magnetic moments* which the neutrons possess, in spite of the fact that they carry no charge. "Unmagnetized" iron is actually divided into small domains in each of which the iron is strongly magnetized in a certain direction; in the "unmagnetized" condition, the direction of magnetization varies from one domain to another in such a way that the resultant magnetic moment is zero. Strong local magnetic fields are associated with the domains, and these fields scatter the neutrons. When the iron is magnetized, the directions of magnetization of the domains are brought more or less into parallelism and the local variations in the field are diminished; the scattering effect is then less. Magnetization of the iron almost to 'saturation' decreases the scattering of thermal neutrons by 10 to 20 percent.³

Cross sections for capture vary greatly, and irregularly, from one element to the next; and they also vary a great deal with the speed of the neutron. Capture cross sections are usually relatively small for fast neutrons and relatively large for thermal ones; for this reason, neutrons to be used for the production of transmutations are commonly first slowed down in paraffin. A few values⁴ of the capture or absorption cross section σ_c and of the scattering cross section σ_s , for thermal neutrons at room temperature, are, in units of 10^{-24} cm.², the following:

¹ LEA, *Roy. Soc. Proc.*, vol. 150, p. 637 (1935); WILDENBECK, *Phys. Rev.*, vol. 68, p. 1 (1945).

² LITTLE, LONG, and MANDEVILLE, *Phys. Rev.*, vol. 69, p. 414 (1946).

³ FBYEK, *Phys. Rev.*, vol. 70, p. 235 (1946); BLOCH, OONDT, and STATUB, *Phys. Rev.*, vol. 70, p. 972 (1946).

⁴ From VOLZ, *Zeits.f. Physik*, vol. 121, p. 201 (1943).

$\bar{Z}$	1	1	2	3	5	6	7	8	13	20
Element.	H	D	He	Li	B	C	N	O	Al	Ca
σ_c	0.25	<0.03	0+	58	500	<0.06	1.05	0+	0.42	0.28
σ_γ	47.5	7	(1.5)			4.6	8.2	4 2	1.6	9.5
Z	26	29	30	45	47	48	56	64	79	80
Element	Fe	Cu	Zn	Rh	As	Cd	Ba	Gd	An	Hg
σ_c . . .	1.6	2.2	1 2	149	58	2950	0.95	30,000	72	325
σ_γ	1 0	8±	5 2		(6)	(1)	8 2			(5)

Some of these cross sections seem, at first sight, to be astonishingly large in comparison with estimates of nuclear size. Equation (322) gives for the radius of the cadmium nucleus, for example,

$$p = 2 \times (48)^{1/3} \times 10^{-13} = 7.3 \times 10^{-13} \text{ cm.}$$

and, for its area of cross section, $\pi p^2 = 1.7 \times 10^{-24} \text{ cm.}^2$. Thus for cadmium σ_c is over 1,000 times πp^2 . And yet nuclear forces are supposed to have a very short range! The explanation undoubtedly is to be found in the fact that classical mechanics is inapplicable here; we are dealing with a wave phenomenon, not a collision between classical particles. In Sec. 48 it was remarked that, according to classical electromagnetic theory, an oscillating electron, in spite of its minute size, might absorb as much radiation as falls on an area equal to a considerable part of a square wave length. The same principle holds in wave mechanics. The de Broglie wave length of a thermal neutron, as calculated from Eq. (166a) in Sec. 107, is $1.5 \times 10^{-8} \text{ cm.}$, so that even 1/100 of a square wave length, or $2 \times 10^{-18} \text{ cm.}^2$, is far larger than any known value of σ_c .

Wave-mechanical theory indicates, further, that at sufficiently low speeds the *cross section for capture* should vary *inversely as the speed* v of the neutron, *i.e.*

$$\sigma_c \propto \frac{1}{v}.$$

It is as if the chance of capture were proportional to the length of time **during which** the neutron remains in the neighborhood of the nucleus. **For velocities of thermal** magnitude, this law has been shown by **experiment to hold** in the case of silver and boron, but it does not

hold for cadmium. In these experiments, the velocity of the neutrons relative to the target was varied by causing the target, mounted on a revolving disk, to have a known velocity toward or away from the neutron source.¹ Using another method² it was found that for boron $\delta_c = (1.55 \times 10^8/v) \times 10^{-24} \text{ cm}^2$, where v is in centimeters per second, up to a neutron energy of 50 ev. For thermal neutrons, with

$$v = 2.2 \times 10^5 \text{ cm. per sec,}$$

this gives $\sigma_c = 708 \times 10^{-24} \text{ cm}^2$.

The absorption cross section of cadmium does decrease, however, if the neutron velocity is increased sufficiently. For example, a layer of cadmium 0.5 mm. thick effectively absorbs all thermal and near-thermal neutrons out of a beam; in the process of absorption, the

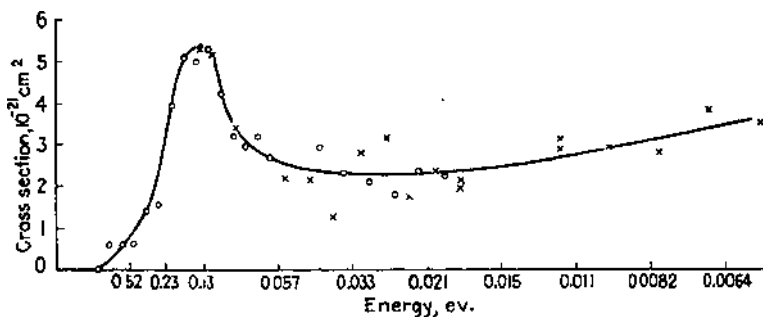


FIG. 211.—Variation of the cross section for capture of very slow neutrons in cadmium. (After Baker and Backer.)

cadmium atoms become increased in mass number by 1, and the excess energy is emitted as γ -radiation. The absorption is enormously weaker, however, for neutrons of energy above 0.4 electron-volt. For this reason cadmium screens are often used where it is desired to filter out neutrons of thermal energies and to leave only the faster ones. In Fig. 211 is reproduced a curve showing the variation of σ_c in cadmium for neutrons of extremely low energy.³

The absorption by capture of slow neutrons in cadmium is thus selective in much the same way as the absorption of light by dyes is selective. It appears to be due almost entirely to the single isotope Cd^{113} , which forms 12.3 percent of ordinary cadmium.⁴ A number of other substances have been found to possess absorption lines or bands, commonly called "resonance" levels of energy. Silver has a strong

¹*Phys. Rev.*, vol. 49, pp. 104, 777 (1936).

²BACHER, BAKER, and MCDANIEL, *Phys. Rev.*, vol. 69, p. 443 (1946).

³BAKER and BACHER, *Phys. Rev.*, vol. 59, p. 332 (1941).

⁴*Phys. Rev.*, vol. 59, p. 666 (1946).

resonance level at 5.8 Mev.¹ Nitrogen has resonance levels at 0.55, 0.70, and 1.45 Mev, associated with the $N(w, p)C$ and $N(n, \alpha)B$ reactions.² The phenomenon is of considerable theoretical interest.³ Wave mechanics predicts proportionality of σ_c and $1/v$ only for neutron energies that are small relative to the lowest resonance level.

217. Survey of Nuclear Reactions.—For the bombardment of nuclei, four principal kinds of projectiles have been used: protons, deuterons, helium nuclei (*i.e.*, α particles or "helions"), and neutrons. Each of these four kinds of particles has been observed to cause the emission from some nucleus of each of the others, with the single exception that deuterons have been observed to be emitted only in the reaction



The heavier nuclei are more easily transmuted by neutrons, which are not repelled by the nuclear charge; α particles, on the other hand, are effective chiefly on the lighter, nuclei. Furthermore, electromagnetic radiations of sufficiently short wave length, called "y-rays" when emitted by a nucleus or "X-rays" when produced by electrons, have been observed to cause the emission of neutrons from many materials; and y-rays, in turn, are often observed to be emitted as the result of bombardment of nuclei with material particles. Finally, nuclear excitation or even transmutation can be produced in some cases by means of high-speed electrons.

In the discussion of nuclear transmutations care in the use of terms is necessary. Following Bohr, a bombarding particle is always assumed to be captured by the "initial" nucleus that it strikes. In some cases the new nucleus formed by the capture represents the final product of the bombardment, although the emission of one or more y-rays may be necessary to bring it into its normal state. More commonly, however, the nucleus formed by capture subsequently undergoes further change through the emission of one or more particles. If this emission of a particle, usually an α or β particle, occurs with a sufficient average delay in time so that the process can be followed experimentally, as in radioactive decay, then the nucleus formed by capture is regarded as the final product of the reaction as such, and it is said to be radioactive. If, on the other hand, the emission of a particle occurs too rapidly for easy observation, then the nucleus formed by capture, called an *intermediate* nucleus, is commonly

¹ BACHER, BETHE, and MCDANIEL, *Phys. Rev.*, vol. 69, p. 443 (1946).

² BARSCHALL and BATTAT, *Phys. Rev.*, vol. 70, p. 245 (1946).

³ For other examples see *Phys. Rev.*, vol. 71, pp. 165 and 174 (1947).

ignored and only the residual nucleus produced by its disintegration is mentioned in describing the reaction. The residual nucleus may then, also, itself either emit γ -rays or undergo further radioactive change. Occasionally even the emission of γ -rays from a group of nuclei is sufficiently distributed in time to be followed in observation; such an emission of γ -rays, as well as that which closely accompanies the emission of α or β particles, is regarded as a form of radioactivity. Another form, furthermore, is decay by electron capture, to be described in Sec. 225. All forms of radioactivity are sometimes referred to simply as an "activity."

A bombardment process is often denoted by parentheses enclosing two symbols, the first symbol denoting the *incident* particle or ray, the second, the *emitted* particle or ray (or particles); α , p , d , n , γ are used to denote, respectively, an α particle, a proton, deuteron, neutron, or γ -ray. A symbol denoting the initial nucleus may precede the parentheses, and another denoting the residual nucleus left after the departure of an emitted particle (or particles) may follow it. The reaction (329) would thus be denoted by $\text{Be}^9 (p, d) \text{Be}^8$. In this notation the best known types of nuclear reactions may be listed as follows:

$$\begin{array}{cccccc}
 (\alpha, p) & (p, \alpha) & (d, \alpha) & (n, \alpha) & (\gamma, n) \\
 (\alpha, n) & (p, n) & (d, p) & (n, p) & \\
 & (p, \gamma) & (d, n) & (n, 2n) & \\
 & & (d, 2n) & (n, \gamma) &
 \end{array}$$

Here $2n$ denotes the emission of 2 neutrons.

Reactions of all of these types have been found, in certain instances, to result in the production of *radioactive residual nuclei*. A list published in 1944¹ contains over 420 different kinds of artificial radioactive atoms, differing from each other in mass number or in atomic number or both. Such atoms are particularly easy to identify because they can be followed by means of their radioactivity and identified by chemical procedures. Observed periods (time for the activity to decrease by one half) range from 0.022 sec. (B^{12}) to more than 1,000 years (Be^{10} , C^{14} , Cl^{86}). A radioactive isotope, natural or artificial, is now known for every element except hydrogen, including $Z = 93$ and $Z = 94$.

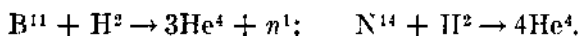
The total number of nuclear reactions known today is near 1,000. The list just mentioned contains about 840 reactions resulting in induced radioactivity, not counting fission processes; and many others are known in which the residual nucleus is one of the stable ones that also occur naturally. Out of this wealth of material we can select

¹ LIVINGOOD and SEABORQ, *Rev. Modern Phys.*, vol. 16, p. 1 (1944).

for detailed discussion only a very few reactions. These we shall group under general statements of the features which they have been chosen to illustrate.¹ Certain features requiring lengthy discussion will be made the topic of a special section; others will be listed briefly in the remainder of this introductory section.

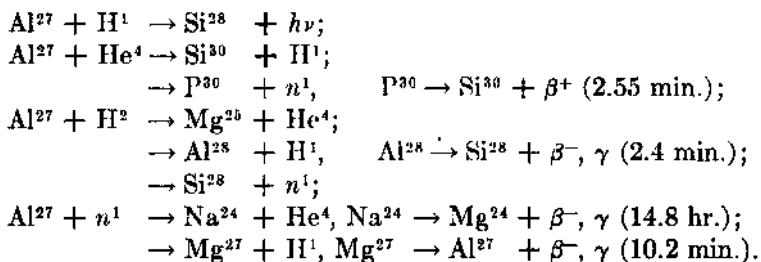
(a) *Identification of the isotope* that is undergoing reaction is easy with an element that has only 1 isotope, but can usually be inferred only indirectly when several isotopes occur. Only for a few elements, notably hydrogen, lithium, and uranium, have the isotopes been separated in sufficient quantity to test the assignment of nuclear reactions directly. In general, the principle of maximum simplicity has been followed, and with great success; out of all possible reactions, the simplest one is assumed that will fit all relevant observations, and a residual nucleus that is not radioactive and hence remains undetectable is assumed to be one of the stable isotopes that occur in nature. An example of identification from indirect evidence will be given presently under (d).

(b) *Sometimes three or more particles result from a bombardment, e.g.;*



In these examples the peculiar stability of the α particle is again in evidence. Enormous bombardment energies, such as those of cosmic-ray particles [Sec. 239 (c)] or of 100 — Mev X-rays,² may also cause the emission of three or more particles.

(c) *The result of a bombardment varies greatly with the nature of the bombarding particle, and is often not unique for a given particle.* For example,



Here β^+ indicates a positive β -ray (or positron) and β^- a negative one. The symbol $h\nu$ is written to denote γ -rays emitted by a residual nucleus that has been formed by bombardment (here Si^{28}); this nucleus is

¹ References, if not given, may be found in LIVINGSTON and BETHE, *Rev. Modern Phys.*, vol. 9, p. 245 (1937), or in LIVINGOOD and SEABORG, *loc. cit.*

² BALDWIN and KLAIBER, *Phys. Rev.*, vol. 70, p. 259 (1946).

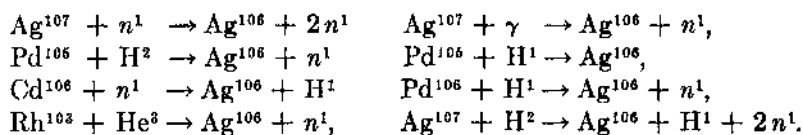
formed in an excited state and emits one or more γ -rays in returning to its normal state of lowest energy, out of which it may subsequently pass by undergoing radioactive change. The symbol γ is reserved for a γ -ray emitted by a nucleus which has itself been formed by radioactive change (here Si^{28} , Mg^{24} or Al^{27}). The distinction is purely artificial, however, and is not always maintained in writing reactions. The times given are the periods of the radioactive atoms (here P^{30} , Al^{28} , Na^{24} , Mg^{27}), or the times required for the disintegration of half of a large group of atoms that have been formed simultaneously.

It will be noted that capture of a deuteron by an Al^{27} nucleus produces an intermediate nucleus which can disintegrate in either of three ways, with the production of either Mg^{26} , Al^{28} , or Si^{28} as the residual nucleus. The intermediate nucleus in this case, Si^{29} , apparently belongs to one of the stable isotopes of silicon, but it is formed in such a high energy level that disintegration promptly occurs; and presumably there are three such levels in any one of which it may be formed. After capture of either He^4 or n^1 by Al^{27} , the disintegration may pursue either of two paths.

(d) *The Same Radioactive Isotope Can Usually be Produced, in More than One Way.* As an example, Ag^{106} , with decay reaction,



can be produced in at least eight ways:



At least six of these eight reactions also produce another form of Ag^{106} which decays to Pd^{106} by K capture (Sec. 225), with a period of 8.2 days. This is a case of *radioactive isomerism*; two forms of the same nucleus exist, having the same values of both Z and A but undergoing different radioactive changes.

The reasoning which led to the identification of the isotope as Ag^{106} may be described as an example of such reasoning.¹ The substances responsible for the 25.5-min. and the 8.2-day activities are easily identified as forms of silver by means of chemical separation. It was assumed that they were not Ag^{107} or Ag^{109} because these are the stable isotopes occurring in nature, or Ag^{108} or Ag^{110} because such isotopes ought easily to be formed by the absorption of thermal neutrons by silver, whereas bombardment with thermal neutrons was

¹ POOL, CORK, and THORNTON, *Phys. Rev.*, vol. 52, p. 380 (1933).

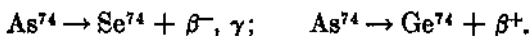
shown not to give rise to the activities in question. The relative abundance in percent of the natural isotopes of silver and of neighboring elements is shown in the following table:

	$\begin{array}{c} A \\ \diagdown \\ Z \end{array}$	103	104	105	106	107	108	109	110
Rh	45	100							
Pd	46		9.3	22.6	27.2		26.8	••••	13.5
Ag	47					51.9		48.1	
Cd	48				1.4		1.0		12.8

Now both activities can be evoked by bombarding rhodium with α particles; the emitted particle in this case can only be a neutron, since Z must increase by 2, leaving as a residual nucleus Ag^{106} , or, conceivably 2 neutrons which would leave Ag^{105} . Three neutrons have never been observed to be emitted in an ordinary nuclear reaction. Both activities are also evoked by bombarding palladium with deuterons, which may produce by a (d, n) reaction either Ag^{105} out of Pd^{104} or Ag^{106} out of Pd^{106} ; and they are also evoked in cadmium by fast neutrons, presumably in the reaction $\text{Cd}^{106} (n, p) \text{Ag}^{106}$, since a (n, d) reaction yielding Ag^{105} is highly improbable. The assignment of both reactions to Ag^{106} is thus made almost certain; and it has been further confirmed by identification of the particles that are emitted in some of the assumed reactions.

(e) *Radioactive isotopes commonly lie close on the mass scale to stable isotopes of the same element, lighter ones tending to transform radioactively with a decrease of Z , heavier with an increase.* Such relations would be expected. Typical examples are presented by the isotopes of sodium, Na^{22} (radioactive, β^+), Na^{23} (stable), Na^{24} (β^- , γ); and the isotopes of antimony, $\text{Sb}^{116,118,120}$ (all radioactive, β^+), $\text{Sb}^{121,123}$ (both stable), $\text{Sb}^{122,124,127,129,131}$ (all β^-).

(f) *A few radioactive isotopes lying between 2 stable isobars are converted sometimes into one isobar, with the emission of a positron, sometimes into the other, with the emission of a negative electron.* The period of decay as observed is necessarily the same for both processes, since the decay represents their combined effect. Example:



The period is 16 days. Gamma rays of energy 0.582 Mev also accompany **this reaction, but their** relation to the β -rays is unknown.

(g) *Artificial radioactive nuclei formed with an increase in N — P tend to exhibit negative β -ray activity, those formed with a decrease in*

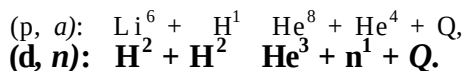
$N - P$, *positron activity*. Here N is the number of neutrons in the nucleus and P the number of protons. This would be expected as a consequence of the tendency for the ratio N/P to change toward a stable value (Sec. 209). Emission of a negative electron increases P at the expense of N , emission of a positron has the opposite effect. The principal types of reactions can be classified as follows with respect to the predominant kind of β -ray activity to which the reactions of each type give rise, numerous exceptions occurring however, except for (p, y) reactions:

Activity. .	β^-		β^+		
$A(N \rightarrow P)$	2	1	0	- 1	- 2
	(n, p)	$(\alpha, p) \ (d, p)$ $(n, \alpha) \ (n, y)$	(d, α)	$(\alpha, n) \ ((b, n) \ (p, y)$ $(n, 2n) \ (y, n)$	(p, n)

(h) *Radioactive isotopes may be formed of an element otherwise unknown.* At least eight reactions result in production of radioactive products which should have $Z = 43$ and whose common chemical properties are those to be expected for this missing element. Presumably the absence of this element in nature results from some accidental combination of features that makes every combination of neutrons with 43 protons unstable. Radioactive forms of transuranic elements with $Z = 93$ (neptunium) and $Z = 94$ (plutonium) have also been produced, the latter in large quantities (see Sec. 227).

(i) *Excitation Functions.* Increase of the bombardment energy usually increases the frequency of reaction, but resonance levels are often observed at which there occur either rapid increases of yield or actual maxima in the yield curve. Examples are described in Sees. 223 and 226.

218. Indirect Measurement of Atomic Masses.—The masses of isotopes that are excessively rare or absent in nature may often be calculated from a nuclear reaction. For example, in cyclotron work some evidence has been found for the existence in helium of a light isotope, He^3 , but in extremely minute amounts. Such nuclei should **also** be produced in the two reactions:



The masses of the other nuclei involved here are known. The **value** of Q for the first reaction was determined by Neuert, and **more** recently

by Perlow, who made very careful measurements of the range of the emitted α particles. The second reaction was studied twice by Bonner; in his second investigation, he measured, by means of recoil protons in methane, the velocities of the neutrons projected straight forward by the deuterons, thus eliminating the effects of angular spread. The date, place, and results of the four investigations are as follows:

	Place	Date	Reaction	Q (Mev)	He ³ mass
Neuert*	Cologne	1935	(p, α)		$3.0172 \pm .0007$
Bonner†. . . .	Rice Institute (Texas)	1938	(d, n)		$3.01700 \pm .00010$
Perlow‡	Chicago	1940	(P, α)	3.95	$3.01688 \pm .00011$
Bonner§. . . .	Cambridge (England)	1941	(d, n)	3.31	$3.01698 \pm .00006$

* *Phys. Zets*, vol. 36, p. 629 (1935)

† *Phys. Rev.*, vol. 53, p. 711 (1938).

‡ *Phys. Rev.*, vol. 58, p. 218 (1940)

§ *Phys. Rev.*, vol. 59, p. 237 (1941).

These results illustrate the increasing accuracy of nuclear work and the close agreement that can be obtained between nuclear measurements made in different laboratories.

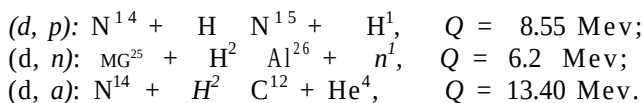
219. Exoergic and Endoergic Reactions.—*Nuclear reactions may be exoergic or endoergic, i.e., Q may be positive or negative. Q is the net amount of rest energy of particles or nuclei that is converted into other forms of energy. The value of Q may be found in either of two alternative ways. (1) If the masses of all of the particles that are involved in the reaction are known, the value of Q may at once be calculated by subtracting the combined masses of the final particles from the sum of the masses that are present initially and then converting the difference into suitable units. (2) Alternatively, Q may be calculated from the net gain in ordinary energy, by adding to the observed kinetic energy of the residual nucleus that of any particles that are emitted in the reaction, and also the energy of any γ -rays and the rest masses of any β particles whose emission may be included in the reaction, and then subtracting from the number thus obtained the known kinetic energies of the bombarding particle and of the initial nucleus. Usually, in stating Q , β -rays and γ -rays whose emission extends over an observable time are not included as parts of the reaction.*

To evoke an *endoergic* reaction, in a sensibly stationary nucleus, the bombarding particle must itself possess kinetic energy at least equal to $-Q$, since energy equal to $-Q$ disappears in the reaction and the

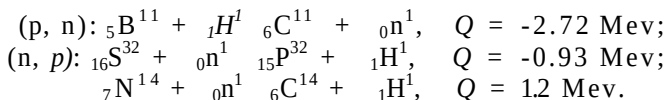
final energy, consisting of kinetic and emitted radiant energy or rest energy, cannot be negative. The actual "threshold" or minimum kinetic energy that is required in order to evoke a given reaction may, however, for other reasons, exceed $-Q$.

The (p, a) reactions discussed in Sec. 214 were all exoergic. Some reactions of this type are endoergic.

Reactions equivalent to the simple addition of a proton or neutron to a nucleus, such as (d, ri) , (d, p) , (p, y) , (n, y) are nearly always *exoergic*, the nuclear energy decreasing because of the attraction between protons and neutrons that is postulated in order to explain nuclear binding (Sec. 209). Most (d, a) reactions, also, are exoergic, the net result here being the emission of two proton-neutron pairs in the form of a closely bound a particle as against the capture of a single proton-neutron pair in the form of a loosely bound deuteron. For example:



On the other hand, (p, n) or (n, p) reactions are usually *endoergic*, A (p, n) reaction merely substitutes a proton for a neutron in the initial nucleus, whereas an (n, p) reaction substitutes a neutron for a proton; in either case Z is changed by unity without any change in the isotopic number A . In a stable nucleus, however, the division of the A constituent particles into protons and neutrons has been adjusted so as to make the energy as low as possible, or very nearly so [Sec. 209(a)]; hence any change in Z without a change in A is almost sure to increase the energy of the nucleus, at the expense of the kinetic energy of the incident proton or neutron. Examples:¹

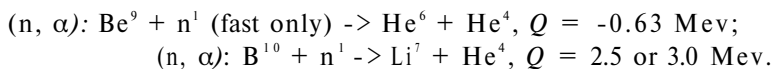


Because of the increase in nuclear energy that should result from a (p, n) or (n, p) reaction, the residual nucleus produced in such a reaction should usually be radioactive, tending to revert to the initial value of Z by the emission of a positive or negative β -ray; and this expectation appears to be confirmed by experiment.

Neutrons must obviously be *fast* in order to produce an *endoergic* reaction, in which energy of ordinary forms is lost; whereas an *exoergic*

¹ HAXBY, SHOUP, STEPHENS, and WELLS, *Phys. Rev.*, vol. 58, p. 1035 (1940); HUBER, *Helv. Phys. Acta*, vol. 14, p. 163 (1941).

reaction may be produced by either fast or slow neutrons, so far as considerations of energy are concerned. Examples:



The second of these reactions occurs especially freely with slow neutrons and is much used for their detection [Sec. 216(6)]. It appears that the Li^7 nucleus is formed only about 9 percent of the time in its normal state; more often it is formed, with a smaller Q value, in an excited state, from which it subsequently drops into its normal state with the emission of a γ -ray.¹

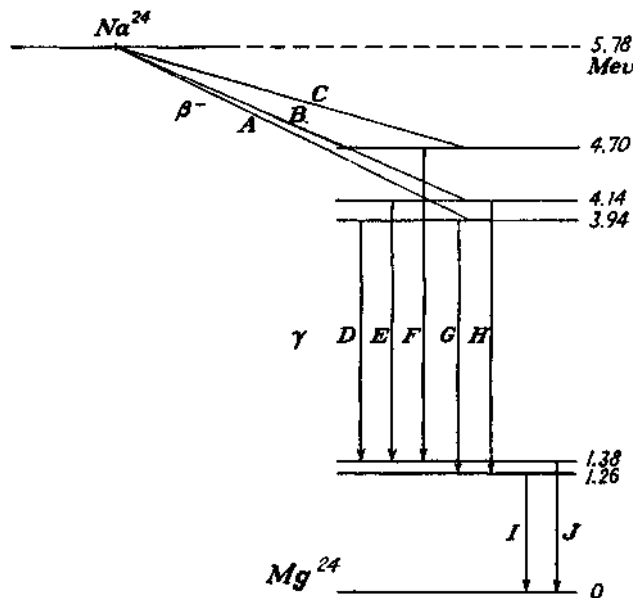
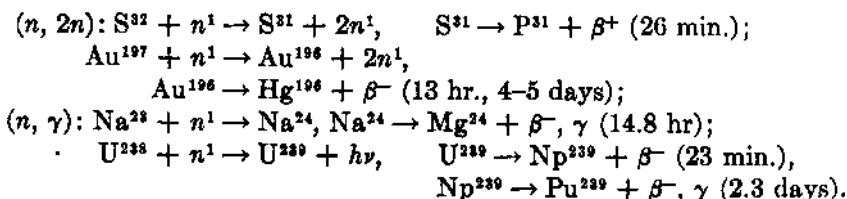


FIG. 212.—The energy levels of Mg^{24} that are involved in the disintegration of Na^{24} (according to Kruger and Ogle).

220. Isotopic Reactions.—All $(n, 2n)$ and (n, γ) reactions produce a residual nucleus isotopic with the initial one, usually detected by means of its radioactivity. Many examples are known, e.g.,



¹ WILSON, *Roy. Soc., Proc.*, vol. 177, p. 382 (1941).

Here Np stands for the transuranic element neptunium ($Z = 93$) and Pu for plutonium ($Z = 94$). The isotope Pu^{239} is a weak α -ray emitter¹ (see also Sec. 227).

The two periods given for Au^{196} indicate another example of radioactive isomerism. The bombardment creates the Au^{196} nucleus in either of two quantum states, which decay radioactively at different rates, and neither of which can pass into the other with the emission of a γ -ray (this being a "forbidden" transition).

The transformation of Na^{24} is very complex and merits description. According to Kruger and Ogle, who studied the γ -rays by measuring the energies of pair particles generated by them in a cloud chamber,² at least seven different γ -rays are emitted by the final Mg^{24} nucleus. Furthermore, the β -rays that are emitted in the disintegration of Na^{24} fall into three groups, corresponding to three alternative levels in which the Mg^{24} nucleus may be formed. The proposed level scheme and the various transitions are shown in Fig. 212. The energies of the observed γ -rays, the maximum energies of the β -rays in Mev, and the relative intensities within each group, are as follows:

β -rays			γ -rays						
1	B	C	D	E	F	G	II	I	J
1 84 515	1 63 160	1 07 40	2 56 26	2.76 15	3 24 0.9	2.68 16	2.89 9	(1.26) (25)	1.38 42

221. The Excitation of Gamma Rays.—Nuclei in excited states, out of which they can undergo transitions with the emission of γ -rays, may be produced in several different ways. Several cases have already been cited; the principal cases will now be listed here in order.

(a) *Fluorescent Gamma-rays.*—Many attempts were made to produce the analog among γ -rays of ordinary fluorescence, but success was not attained until a continuous spectrum of radiation of sufficiently high frequency was available. The reason lies no doubt in the extreme narrowness of the nuclear levels that are involved in γ -ray emission; the absorption spectrum of a nucleus consists of narrow lines, and it will be a rare accident if any γ -ray emission line coincides with the absorption line of another nucleus.

¹ OLIPHANT, *Nature*, vol. 157, p. 5 (1946).

² KBUGER and OGLE, *Pkys. Rev.*, vol. 67, p. 273 (1945). See, however, *Phys. Rev.*, vol. 70, p. 985, also p. 978 (1946).

The following example is interesting. Gold, under illumination by X-rays with an $h\nu$ value exceeding 1.22 Mev, was found by Wiedenbeck to develop a short-lived 7-ray activity, unaccompanied by β -rays.¹ The high potential required to generate the X-rays was obtained from a Van de Graaff-Herb electrostatic generator. The $h\nu$ value of the 7-rays, calculated from the absorption in aluminum foil of the internal-conversion electrons generated by them, was 0.25 Mev; their intensity was observed to decrease to half in 7.5 sec.

An excitation curve for the generation of these 7-rays, reproduced in Fig. 213, was obtained by varying the potential on the X-ray tube and therewith the upper frequency limit of the X-rays. The 7-ray activity begins to be observed at $h\nu = 1.22$ Mev, and the curve of its intensity exhibits definite breaks at 1.68, 2.15, 2.56, and 2.97 Mev.

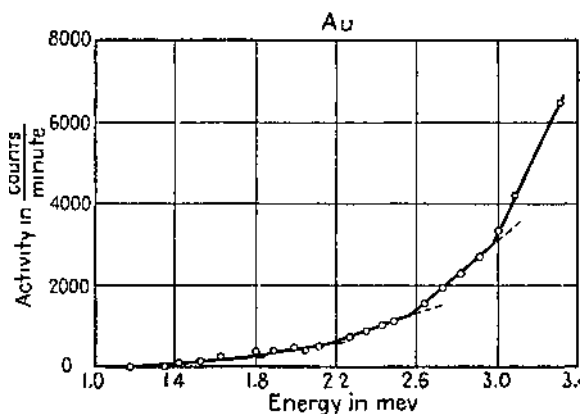


FIG. 213.—X-ray excitation curve for gold. The abscissa represents the maximum energy of the X-ray spectrum. (Obtained by Wiedenbeck.)

The observations were interpreted by Wiedenbeck to mean that the gold nucleus possesses electromagnetic absorption lines at 1.22, 1.68, 2.15, 2.56, and 2.97 Mev and corresponding energy levels lying at these distances above the normal level. From any one of these levels the nucleus is able to pass very quickly, with the emission of a high-energy fluorescent γ -ray, into a level lying only 0.25 Mev above the normal level; and a further transition then gives rise to the observed activity. The level at 0.25 Mev is metastable; the probability of a transition out of it into the normal level is so small that the half-life of nuclei in this state is 7.5 sec. For this reason, also, no absorption line at 0.25 Mev can be detected. The more energetic higher-frequency γ -rays emitted very quickly during the descent into the metastable level were not detected.

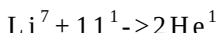
¹ WIEDENBECK, *Phys. Rev.*, vol. 67, p. 53; vol. 68, p. 1 (1945).

Rhodium was found to possess a similar metastable level lying only 0.04 Mev above the normal level.¹ The corresponding γ -rays, with energies of only 40 kilovolts, lie in the ordinary X-ray region. The half-life period is 45 min. Higher excitation levels were found at 1.26, 1.64, 2.02, 2.37, 2.71, and 3.05 Mev.

(b) *Excitation by Inelastic Bombardment.*—Wiedenbeck found that the 7-ray activity in gold just described could be evoked also by bombardment with *neutrons*. The neutrons were generated by photodisintegrating beryllium with high-frequency X-rays, so that their energy could easily be varied. It was found that activation of the gold began only when the neutron energy was raised above 1.22 Mev. It may be concluded that the neutrons also produce the metastable level at 0.25 Mev only indirectly. Since the activity exhibits the same properties whether produced by photons or by neutrons, it is to be presumed that the bombarding neutrons are not captured but merely excite the gold nucleus without otherwise altering it.

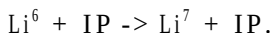
The activities in rhodium could also be evoked by bombardment with *electrons* of 2.5 Mev energy.

Another interesting example of noncapture excitation is presented by one of the first proton reactions to be discovered:



(Sec. 214). This reaction is observed to be accompanied by the emission of 7-rays. There appears to be a definite 7-ray line or band at 0.46 Mev, emitted at least whenever the energy of the bombarding protons exceeds 0.85 Mev.² These particular 7-rays have been ascribed to a noncapture excitation of the Li^7 nucleus by collision with a proton, the Li^7 nucleus emitting a 7-ray in returning to its normal state. That is, whereas some Li^7 nuclei capture the H^1 nucleus and then break up into two α particles, according to the reaction just written, others are merely excited as the result of an inelastic collision.

The assumption that the Li^7 nucleus actually possesses an excitation level 0.46 Mev above its normal state is supported by a peculiar feature of another reaction in which Li^7 nuclei appear as products, viz.,



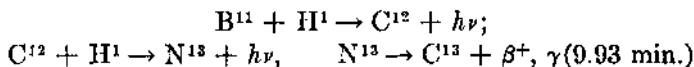
Two groups of protons are produced by this reaction (coming from different individual nuclei of course); their energy difference, as meas-

¹ WIEDENBECK, *Phys. Rev.*, vol. 67, p. 267; vol. 68, p. 237 (1945).

² HUDSON, HERB, and PLAIN, *Phys. Rev.*, vol. 57, p. 587 (1940).

ured, is 0.45 Mev.¹ It is assumed that the faster protons arise when the Li^7 nucleus is left in its normal state, the slower when it is left in an excited state lying 0.45 or 0.46 Mev higher.

(c) *Gamma Rays from a Nucleus Formed by Capture.*—In addition to many (n, γ) reactions, of which examples were given in Sec. 220, a number of (p, γ) reactions are known, in which a proton is captured to form a product nucleus in an excited state and the excess energy is then emitted as a γ -ray. The product nucleus may or may not subsequently undergo radioactive change. For example,



Other cases, in which the active nucleus is formed by capture of an electron, will be described in Sec. 225.

(d) *Emission by an Intermediate Nucleus.*—Among the γ -rays from a mass of lithium bombarded by protons there is a hard component with a photon energy of about 17.5 Mev.² It is believed that these γ -rays are emitted by the *intermediate* Be^8 nucleus, a complete statement of the reaction being as follows:



Here the symbol Be^{8*} stands for an excited beryllium nucleus and $h\nu$ for the γ -ray. The entire amount of rest mass that disappears in the reaction is:

$$\begin{aligned} \text{Li}^7 + \text{H}^1 - 2\text{He}^4 &= 7.01804 + 1.00812 - 2 \times 4.00388 \\ &= 0.01840 \text{ a.m.u.} = 17.15 \text{ Mev.} \end{aligned}$$

This energy, plus the kinetic energy of the H^1 , must supply both the energy of the γ -ray and the kinetic energy of the two α particles. The kinetic energy of the α particles, by a principle of mechanics, cannot be less than that of the Be^8 nucleus out of which the α particles are formed; and the kinetic energy of Be^8 will be indistinguishable from that of Be^{8*} . The kinetic energy of Be^{8*} , in turn, is $\frac{1}{8}$ of that of the incident H^1 , the Be^{8*} nucleus being eight times as heavy as H^1 and hence retaining $\frac{1}{8}$ of the velocity of the proton and the same fraction of its kinetic energy. Now resonance, *i.e.*, maximum excitation of the γ -rays in question, occurs when the protons have an energy of 0.44 Mev. Adding $\frac{7}{8}$ of this energy to the energy available from the rest masses, we have a maximum of 17.54 Mev available for the γ -ray. This probably exceeds the energy of the γ -rays, thus leaving something

¹ RTJMBGAUGH, ROBERTS, and HAFSTAD, *Phys. Rev.*, vol. 64, p. 657 (1940).

² Cf. HUDSON, HERB, and PLAIN, *Phys. Rev.*, *loc. cit.*

over for the kinetic energy of the α particles, which form a much slower group than those that arise from the alternative immediate disintegration of Be^8 and are described in Sec. 214.

In most instances, however, the emission of material particles tends to occur so much more quickly than the emission of radiation that r-rays can be emitted only by nuclei which are either stable toward the emission of particles or at most α - or β -active.

(e) *Gamma-rays from Residual Nuclei.*—Reactions of type (α, p) or (α, n) often result in the emission of r-rays; so do some deuteron reactions. These r-rays are believed to be emitted by the residual or product nucleus, which is left in an excited state by the departure of the proton or neutron and emits r-rays as it descends into states of lower energy. The residual nucleus may or may not undergo subsequent radioactive transformation. An example of such r-rays, due to a deuteron reaction, is discussed in Sec. 224 below.

On the other hand, (p, α) and (p, n) reactions do not seem to give rise to r-rays, the residual nucleus being always left by the reaction in its normal state.

(f) *Gamma-rays following radioactive change*, i.e., the emission of an α or β particle, are common, as we have seen. An elaborate example was described in Sec. 217(6).

222. Photodisintegration and Electron Disintegration.—Many (γ, n) reactions are known, in which a γ -ray or high-energy X-ray is absorbed by a nucleus and a neutron is emitted. Such reactions can be detected most surely by observing the neutrons, but a more convenient means of detection is often furnished by the radioactivity of the residual nucleus. As the frequency of the incident radiation is progressively increased, the reaction begins at a certain "threshold" value of the photon energy $h\nu$, and the yield then increases more and more rapidly as the frequency is further increased. If the residual nucleus is formed in its lowest state, the threshold energy represents also the work required to detach a neutron from the initial nucleus.

Since the incident radiation must be assumed to exert its direct action upon the *charged* particles in the nucleus, i.e., upon the protons, whereas it is almost always a neutron that is emitted, photodisintegration is not usually analogous to the ejection of Compton electrons from an atom; it may be compared, perhaps, with the emission of Auger electrons [Sec. 190(6)].

The photodisintegration of the *deuteron* was mentioned in Sec. 216; the threshold for this process is $h\nu = 2.17$ Mev. The effective cross section for 6.2 Mev r-rays¹ is about 1.2×10^{-28} cm.²; i.e., if the r-rays

¹ VAN ALLEN and SMITH, *Phys. Rev.*, vol. 59, p. 618 (1941).

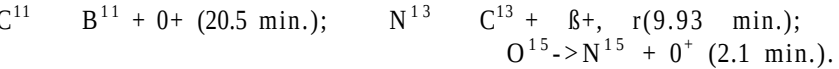
are assumed to be uniformly distributed over an area containing a large number N of deuterium nuclei, as many γ -rays will give rise to disintegrations as fall on an area equal to $1.2N \times 10^{-28}$ cm.² Cross sections for photodisintegration vary with the frequency of the incident radiation.

For the photodisintegration of *beryllium* the threshold is only 1.63 Mev. Furthermore, the same disintegration can be effected by means of electrons of the same energy as the X-rays, as was shown by Collins, Waldman, and Guth.¹ They detected the neutrons by means of radioactivity produced by them in silver, rhodium, or indium. That the effect obtained with electrons was not caused indirectly, by X-rays generated by the electrons in the beryllium, was shown by noting that almost the same emission of neutrons was obtained from beryllium targets of two different thicknesses, 0.038 cm. and 0.20 cm. The thin target was sufficiently thick to stop the incident electrons, whereas even the thick one would not appreciably absorb either the neutrons or X-rays of several Mev energy; hence the observed neutron emission would have been much greater if it were due to X-rays. The cross section for disintegration of beryllium at 2.5 Mev energy is 8×10^{-31} cm.²; the rate at which disintegrations are produced is the same as if the nucleus were a little disk of area 8×10^{-31} cm.² placed perpendicularly to the stream of electrons and all electrons striking this disk gave rise to disintegration.²

In heavier atoms the thresholds for photodisintegration lie much higher. Baldwin and Kock report the following values, in Mev:

	C ¹²	N ¹⁴	O ¹⁶	Fe ⁵⁴	Cu ⁶³	Zn ⁶⁴	Ag ¹⁰⁷	Ag ¹⁰⁹
Obs	19	11	16.3	14.2	10.9	11.6	9.5	9.3
Calc	18.8	10.6	16					

In all cases detection was effected by means of the radioactive product nucleus. The first three decay processes are



The final nuclei being thus stable and well known, it was possible in these three cases to calculate the net loss of nuclear rest energy; from this, after correcting for the emitted radiations, the values of the disintegration energy of C¹², N¹⁴, and O¹⁶ listed as "calc." were obtained.

¹ COLLINS, WALDMAN, and GUTH, *Phys. Rev.*, vol. 56, p. 876 (1939).
² WIEDENBECK, *Phys. Rev.*, vol. 69, p. 235 (1946).

These values agree satisfactorily with the observed thresholds for photodisintegration.

Good evidence for the following (r, p) reaction was obtained by Huber and his coworkers;¹



223. Groups of Emitted Protons; and Alpha Resonances.—When protons are emitted as a result of bombardment, their kinetic energy varies in step with the kinetic energy of the bombarding particles. Thus, conservation of energy is maintained. Even for a given energy of the bombarding particle, however, it is often observed that the emitted protons fall into several groups characterized by different values of the kinetic energy.

This feature, and also the phenomenon of α resonances mentioned in Sec. 217(6), will be illustrated by describing the early work on the reaction



The proton spectrum from this reaction was studied by Chadwick and Constable.² Uranine in the form of calcium fluoride was employed. The curve of proton intensity plotted against proton energy was found to show six broad maxima between the extreme proton ranges (in air) of 20 cm. and 55 cm. The energy of the bombarding α particles, which came from polonium, could be decreased in any desired ratio by interposing a suitable thickness of absorbing material. When the energy of the α particles was decreased below that corresponding to an α -particle range of 3.25 cm. in air, the proton spectrum became shortened, one pair of maxima disappearing; at about 2.7 cm., another pair disappeared; and below 2 cm. the entire proton spectrum was absent. The three pairs of maxima that thus disappeared in succession showed about the same spacing in energy between their component groups of protons.

These facts are explained if it is assumed that the components of any given pair are produced by α particles of the same energy, and that the difference in the two proton energies arises from a difference in the disintegration process. Chadwick and Constable assumed that the emission of the proton may leave the residual Ne^{22} nucleus in either of two possible quantum states. To each of these states will then correspond a different value of the reaction energy Q , which

¹ HUBER, LIENHARD, SCPIERRER, and WAFFLER, *Helv. Phys. Acta*, vol. 17, p. 139 (1944).

² CHADWICK and CONSTABLE, *Roy. Soc. Proc*, vol. 135, p. 48 (1932).

represents the (algebraic) excess of the final kinetic energy, belonging to the proton and the residual nucleus, over the initial kinetic energy, belonging to the initial nucleus and the bombarding particle. For reactions leading to the two states of Ne^{22} , Chadwick and Constable calculated from their data that $Q = 0.99$ or 1.67 Mev, respectively, so that the two levels of Ne^{22} would lie $1.67 - 0.99$ or 0.68 Mev apart.

The occurrence of three similar pairs in the proton spectrum was then accounted for by supposing that, since the α particles are slowed down to a stop in the calcium fluoride, (1) effects due to all α -particle speeds below a certain maximum were obtained, but (2) there exist three *resonance levels* for the α -particle energy at which the reaction occurs with maximum frequency. These levels were assumed to lie at α -particle ranges in air of about 2, 2.7, and 3.25 cm., and correspond presumably to three possible excited levels in which the intermediate Na^{23} nucleus may be formed when a fluorine nucleus captures an α particle. The continuous distribution of the proton energies as observed may have been due in part to straggling of the protons after leaving their point of origin or to other experimental imperfections, but it was probably due also in part to a great breadth of the Na^{23} levels, which is associated with the rapidity of the ensuing disintegration. As a result of this breadth, which implies a considerable indefiniteness in the energy of the short-lived Na^{23} nucleus, α particles can be absorbed over a rather wide range of α -particle energies; the kinetic energies of the emitted protons will then vary in step with the energy of the α particles, so as to preserve the conservation of energy.

The study of the same reaction was extended by May and Vaidyanathan, using the swifter α particles from RaC (7.0 cm. range in air).¹ They found three other groups of protons, whose actual ranges as observed were of the same order as those just described, because of the high α energies, but which corresponded to the negative Q values: -3.2 , -2.1 , -0.1 Mev. They observed no duplications resulting from the superposition of various α resonance levels, apparently because they used thin targets and thus kept their α -particle energies within a narrow range. They recalculated Chadwick and Constable's Q values, from better range-energy data for protons, as 0.87 and 1.47. Pose, using polonium particles, had found Q values² which May and Vaidyanathan recalculated as 0.6 and 1.3 Mev. They concluded, therefore, that protons with a Q value of about 1.4 Mev certainly exist, but, finding no trace of an 0.87 or 0.6 group at a point where they should have found it, they suggested that the appearance of such

¹ MAY and VAIDYANATHAN, *Roy. Soc., Proc.*, vol. 155, p. 519 (1936).

² POSE, *Zeits. f. Physik*, vol. 72, p. 528 (1931).

a group in the observations of Chadwick and Constable and of Pose may have been spurious.

If the 1.4 group is assumed to result from the production of Ne^{22} in its normal state, we have then the following energy levels for Ne^{22} , measured from the normal state:

Ne^{22} level.	0	(0.0)	1.5	3.5	4.6
Q value ..	1.4	(0.8)	-0.1	-2.1	-3.2

Here $1.5 = 1.4 - (-0.1)$, etc. Positive Q means that the protons and the recoiling Ne^{22} nucleus together have more kinetic energy than the bombarding α particle; negative, that they have less.

As calculated from the masses, which refer, of course, to the normal nuclear states,

$$Q = \text{F}^{19} + \text{He}^4 - \text{Ne}^{22} - \text{H}^1 = 19.00452 + 4.00388 - 21.99864 - 1.00812 = 0.00164 \text{ a.m.u.} = 1.53 \text{ Mev.}$$

This agrees sufficiently well with the observed value of 1.4 Mev.

In direct observations of the *yield* for both the (α, p) and the (α, n) reactions of fluorine as a function of α -particle energy, Saha likewise found indications of the existence of resonance levels for the α particles;¹ but his values for these levels do not seem to agree at all with those found by Chadwick and Constable!

This fluorine reaction has been discussed at some length, in part, for the purpose of illustrating the general fact that precise work on nuclear phenomena is difficult, and hence conflicting or imperfect data exist in the literature in regard to many nuclear reactions.

It may be remarked that fluorine bombarded by α particles also emits neutrons, according to the reaction, $\text{F}^{19} + \alpha \rightarrow \text{Na}^{22} + n$.

224. Neutron and α -ray Spectra.—Neutrons emitted as a result of bombardment commonly occur in groups, corresponding to different quantum states for the residual nucleus.

Consider the reaction, so much used as a neutron source:



Using 0.9 Mev deuterons (H^2), and determining the neutron energies from recoil proton tracks produced by them in a high-pressure cloud chamber containing methane, Bonner and Brubaker obtained the curve shown in Fig. 214 for the recoil protons.² Only proton tracks

¹ SAHA, *Zeits. f. Physik*, vol. 110, p. 473 (1938).

² BONNER and BRUBAKER, *Phys. Rev.*, vol. 50, p. 308 (1936).

pointing almost straight away from the target were counted, on the assumption that these tracks were made by protons projected in practically the same direction as the direction of approach of the incident neutrons.

The four peaks in the curve are interpreted as meaning that the B^{10} nucleus may be left in any one of four possible quantum states. Because of this fact, the neutrons arising from bombardment with deuterons of given energy and leaving the scene in a direction at right angles to the deuteron beam should have energies equal to one of four sharply defined values, and, in turn, the protons that are projected straight ahead by these neutrons should likewise be limited to one of four possible kinetic energies.

Actually, however, it is impossible to screen off high-speed neutrons effectively, so that the walls of the cloud chamber are struck by

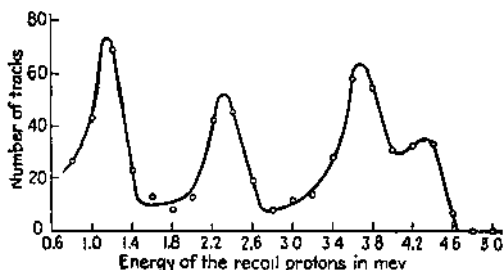


FIG. 214.—Energy distribution of recoil protons due to neutrons from the reaction $Be^9 + H^2(0.9 \text{ Mev}) \rightarrow B^{10} + n$.

neutrons over a wide area; many of them, after being deflected by a nucleus in the wall, may, by an oblique collision, produce a proton moving in the direction selected for observation but with reduced energy. Furthermore, the deuterons are slowed down somewhat in the target, which causes a corresponding variation in the neutron energies. Thus the proton spectrum, as observed, will consist of four broad peaks instead of four lines, and only the high-speed edges of these peaks correspond to neutrons projected in the ideal manner, *i.e.*, by deuterons of maximum energy and at 90° to the deuteron beam. Additional small amounts of spreading result, also, from lack of homogeneity in the incident deuteron beam and from straggling of the proton tracks themselves.

From the curve, the authors conclude that neutrons projected from beryllium at 90° by 0.9 Mev deuterons have energies of 4.52, 4.0, 2.6, or 1.4 Mev. For the neutrons of highest energy we find, from Eq. (325) in Sec. 214, in which $M_1 = 2$, $M_2 = 1$, $M_3 = 10$, nearly enough,

$$Q = ({}^{11}_{10})4.52 - ({}^8_{10})0.9 = 4.25 \text{ Mev.}$$

The masses give $Q = \text{Be}^9 + \text{H}^2 - \text{B}^{10} - n^1$ or

$$Q = 9.01497 + 2.01472 - 10.01605 - 1.00893 = 0.00471 \text{ a.m.u} \\ = 4.39 \text{ Mcv.}$$

The last figure of the masses being quite uncertain as given here, the agreement between 4.25 and 4.39 is satisfactory and indicates that in the emission of the 4.25 Mev neutrons the B^{10} nucleus must be left in its lowest state. For the other three groups of neutrons, we find similarly $Q = 3.68, 2.14, 0.82$. Subtracting these numbers from 4.25, we infer energy levels of B^{10} at 0.6, 2.1, and 3.4 Mev above the normal state, about as shown in Fig. 215.

It happens that the r-rays from this same reaction have been studied by Kruger and Green.¹ They confirm the existence of the

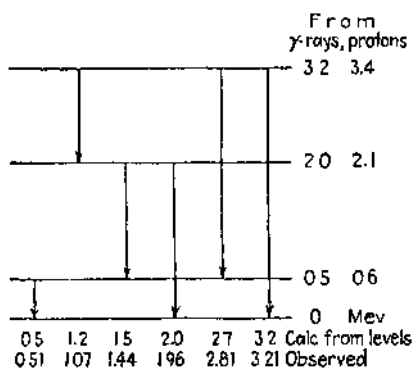
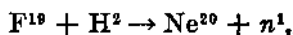


FIG. 215.—Energy levels of B^{10} . All energies, including those of this six r-rays listed at the bottom of the figure, are in Mcv.

four energy levels of B^{10} . From the distribution of the tracks of Compton electrons ejected by the γ -rays from a thin piece of mica foil in a cloud chamber, it was inferred that six r-ray lines occur, with $h\nu$ energies of 0.51, 1.07, 1.44, 1.96, 2.81, 3.21 Mev. These γ -rays could be interpreted as arising from various transitions between four levels of the B^{10} nucleus, as indicated in Fig. 215, where energy levels chosen to fit the r-ray data are represented by the diagram; the r-ray energies calculated from these levels are given below the diagram in comparison with the observed energies. The agreement is within the experimental error. The four nuclear levels as determined from the r-rays agree satisfactorily with those indicated by the neutron groups.

From another reaction,



¹ KRUGER and GREEN, *Phys. Rev.*, vol. 52, p. 773 (1937).

Bonner found seven groups of neutrons, with Q values 10.8, 9.3, 6.6, 5.4, 3.5, 1.8, and 0.7 Mev.¹

The α -rays emitted by a radioactive substance also sometimes fall into different groups, characterized by different values of their kinetic energy. Usually these groups are nearly equally spaced in energy, with a spacing of the order of 100 kilovolts, and those of lower energy are less intense. It is assumed that in such cases the groups correspond to various quantum states in which the residual nucleus left after the departure of the α -ray may be left. The γ -rays associated with transitions between these various levels are often observable. Arguments have been advanced for the belief that some of these nuclear levels may correspond to various states of rotation of the nucleus.²

An alternative possibility is that similar groups may result from differences in the quantum state of the initial nucleus that emits the α -ray, but only when this nucleus is extremely short-lived, in order to exclude the possibility of a prior γ -ray transition; examples are the nuclei of RaC (1.5×10^{-4} sec.) and ThC (3×10^{-7} sec.).

225. Radioactive Decay by Electron Capture.—If a neutron is capable of emitting an electron and turning into a proton, as is assumed in order to explain negative β -ray emission, the reverse should also be possible; a proton should be able to combine with an electron and become a neutron. This cannot happen at low energies, to be sure, for the rest mass of the electron (0.00055 a.m.u.) is insufficient to provide the required increase of mass ($0.00893 - 0.00813 = 0.00080$ a.m.u.). The impossibility of the spontaneous absorption of a slowly moving electron by a proton is doubtless the reason that nuclei containing protons can exist indefinitely even when surrounded by a shell of electrons. The reverse process, spontaneous conversion of a neutron into a proton with emission of an electron, is energetically possible; but this process has not been detected with certainty.³

In an *excited* nucleus, on the other hand, there is an additional store of energy upon which a proton might draw in order to absorb an electron and become a neutron. The hypothesis was put forward on more abstract grounds by Yukawa in 1935 and discussed by Alvarez⁴ that, as an alternative to the emission of a positron, a nucleus containing an excess of protons may suddenly capture one of its circumjacent electrons, usually out of the K shell. In this way, as well as by positron emission, the atomic number Z will be decreased by

¹ BONNER, *Roy. Soc. Proc.*, vol. 174, p. 339 (1940).

² OTJGGENHEIMER, *Roy. Soc. Proc.*, vol. 181, p. 169 (1942).

³ See, however, SUNDARACHAR, *Nature*, vol. 157, p. 286 (1946).

⁴ ALVAREZ, *Phys. Rev.*, vol. 54, p. 486 (1938).

unity without altering the mass number A . Any excess of energy may be either radiated away in γ -rays or, perhaps, carried away by a neutrino that is emitted as the electron and proton coalesce.

Many examples of this process, usually called K capture, are now known. Its occurrence can only be detected indirectly, most commonly by observing either the X-rays or the Auger electrons which are emitted as another electron drops into the hole in the K shell. These X-rays or Auger electrons should have a frequency or energy characteristic of an atom containing the final nucleus that is formed by the capture. They should not be affected by an approximately simultaneous emission of γ -rays by this nucleus as it drops into lower energy levels.

Two good examples, described by Deutsch and Elliott, are the following.¹



The Mn^{64} nuclei, which decay by K capture alone with a half-life of 500 days, were produced by bombarding iron with deuterons, separating the manganese out chemically, then letting the iron age 2 months in order to eliminate the 6.5-day activity of Mn^{52} , which is also produced by the bombardment. By using counters and observing coincidences (within the time resolution of ordinary apparatus) between the X-rays and the accompanying γ -rays, it was shown that each X-ray is accompanied by one γ -ray. It was concluded that, in addition to the X-ray emitted by the circumnuclear electrons, the Cr^{54} nucleus that is formed by K capture emits a single γ -ray, of energy 0.835 Mev, in settling into its normal state.

The case of Co^{68} is more typical in that Co^{68} is known also to emit positrons. The complete radiation from Co^{68} was studied by allowing it to eject photoelectrons from a lead foil. When a curve was drawn showing the number of photoelectrons as a function of their energy, two peaks were observed. One of these, corresponding to an incident γ -ray energy of 0.805 Mev, was ascribed to γ -rays emitted by the transformed nucleus; the other, at 0.51 Mev was ascribed to γ -rays generated in the lead foil by annihilation of the positrons. Comparison of the heights of the peaks indicated that 10 times as many γ -rays as positrons were emitted by the Co^{68} material. Coincidence studies indicated further that one γ -ray was associated with each positron, and another with each of the X-rays that were observed to be emitted. It was concluded that the Co^{58} nuclei all become transformed into nuclei of Fe^{68} in the same excited state, 10 percent of them

¹ DEUTSCH and ELLIOTT, *Phys. Rev.*, vol. 65, p. 211 (1944).

by positron emission and 90 percent by K capture; the Fe^{58} nucleus then emits a 0.805 Mev γ -ray in settling into its normal state.

The energy difference between Co^{68} and the excited Fe^{68} can easily be calculated; it is the sum of the maximum kinetic energy of the positrons, 0.470 Mev, and the rest energy of the positron, or a total of 0.98 Mev. When K capture occurs instead of positron emission, this excess energy is presumed to be carried away by a neutrino.

Still another feature is presented by Te^{121} , which decays by K capture to Sb^{121} , with the emission of mixed γ -rays of energy 0.23 and 0.61 Mev. Among the X-rays was found a weak line characteristic of *tellurium*, not *antimony*. By studying coincidences between the X-rays and the conversion electrons produced by the γ -rays,¹ Yalow and Goldhaber convinced themselves that the Sb nucleus was formed sometimes in one excited level, sometimes in another; from one level, lying 0.23 Mev above the normal state, a γ -ray is emitted so quickly (in less than 2×10^{-17} sec.) that it usually precedes the emission of an X-ray by the electrons in the same atom. This γ -ray is then internally converted half of the time, with removal of the *second* electron from the K shell. The X-ray that is emitted subsequently, coming from an antimony atom with both K electrons missing, then has n frequency approximating that of the ordinary K line from tellurium.

The hypothesis that a neutrino is emitted in the process of K capture is supported by observations of Allen on the recoil of the active nucleus.² He measured the energy with which Be^7 nuclei recoil after decaying by K capture to form Li^7 . The only way to bring the occurrence of such recoils into harmony with the conservation of momentum seems to be to assume that the excess energy is carried off by a neutrino. The energy of the recoiling atoms was found to be 40 to 45 ev, as against a theoretical value of about 58 ev.

226. Alternative Reactions.—Examples have already been given illustrating the common fact that the results of a particular bombardment may vary [*e.g.*, in Sees. 217 (c), 220, and 223]. The following additional case deserves description.

When fluorine is bombarded by protons, both long-range and short-range α particles in addition to γ -rays and pairs of positive and negative β particles are found to be emitted by the bombarded material. Elaborate investigations have been necessary in order to make out the Prelationships of these emissions.³

¹ YALOW and GOLDBABER, *Phys. Rev.*, vol. 67, p. 59 (1945).

² ALLEN, *Phys. Rev.*, vol. 61, p. 692 (1942).

³ See BECKER, FOWLER, and LAURITSEN, *Phys. Rev.*, vol. 62, p. 186 (1942), and references there given; also BENNETT, BONNER, MANDEVILLE, and WATT, *Phys. Rev.*, vol. 70, p. 882 (1946), and SCHIFF, *ibid.*, p. 891.

Important clues were finally found in excitation curves showing the observed abundance of the various products as a function of the energy of the bombarding protons. These curves, as obtained by Streib, Fowler, and Lauritsen,¹ are reproduced in Fig. 216. The resonance peaks in the curves are interpreted as corresponding to various alternative quantum states in which the Ne^{20} nucleus may be formed by the union of a proton with a F^{19} nucleus. *These resonance peaks are*

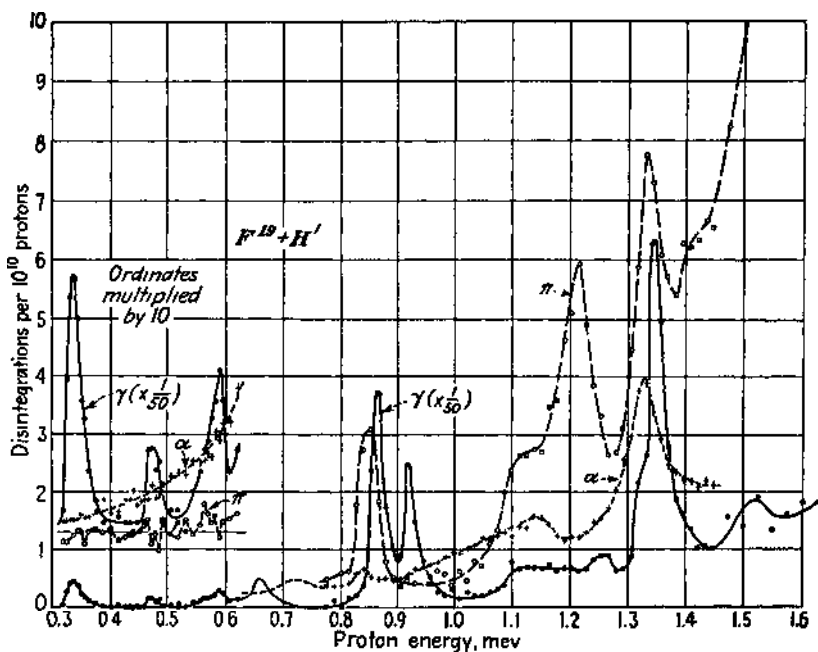


FIG. 216.—Excitation curves for the long-range α particles, γ -rays and β -ray pairs from fluorine bombarded by protons. The ordinates for the γ -ray curves are reduced in the ratio 1/50; for low proton energies, the curves are repeated with a tenfold amplification of ordinates. (From Streib, Lauritsen, and Fowler, *Phys. Rev.*, vol. 59, p. 253, 1941.)

different for the γ -rays, the long-range α particles, and the β -ray pairs. The observed resonance energies (of the incident protons) in Mev for the three emissions just mentioned are as follows:

α particles: 0.72, 0.84, 0.9 to 1.2, 1.35

γ -rays: 0.334, 0.479, 0.589, 0.660, 0.862, 0.927, 1.335, 1.363

Pairs: 0.6 to 0.8, 0.85, 1.14, 1.22, 1.35

The known facts were found to agree with the assumption that at least the following three alternative forms of the reaction occur:

¹ STREIB, FOWLER, and LAURITSEN, *Phys. Rev.*, vol. 59, p. 253 (1941).

- (A): $F^{19} + H^1 \rightarrow \alpha Ne^{20} \rightarrow O^{16} + He^4 + Q_0$
 (B): $F^{19} + H^1 \rightarrow rNe^{20} \rightarrow rO^{16} + He^4 + Q_1, \quad rO^{16} \rightarrow O^{16} + r$
 (C): $F^{19} + H^1 \rightarrow \pi Ne^{20} \rightarrow \pi O^{16} + He^4 + Q_2, \quad \pi O^{16} \rightarrow O^{16} + \beta^+ + \beta^-.$

Here the symbol " Ne^{20} " stands for a group of states of the Ne^{20} nucleus out of which it can only disintegrate into an α particle and an O^{16} nucleus in its normal state, with an evolution of energy $Q_0 = 7.95$ Mev. This value of Q_0 agrees sufficiently well with the loss of rest mass: $F + H - O - He = 0.0045 + 0.0081 - 0.0039 = 0.0087$ a.m.u., equivalent to 8.1 Mev. The emitted α particles have a kinetic energy equal to Q_0 plus the kinetic energy of the bombarding proton, minus the small kinetic energy of the residual O^{16} nucleus; hence these α -particles have a long range.

The symbol ' Ne^{20} ' stands similarly for another group of states from which Ne^{20} disintegrates into an α particle and an O^{16} nucleus in an *excited* state, denoted by rO^{16} , from which it soon passes into its normal state with the emission of a γ -ray. The energy of the γ -ray has been measured with some uncertainty as 6.2 Mev; it is observed to be independent of the energy of the bombarding protons. The α -rays given out in this reaction should thus have an energy corresponding to $Q_1 = Q_0 - 6.2 = 7.95 - 6.2 = 1.75$ Mev. Their range should accordingly be short, not much more than 1 cm. in air of normal density for a bombardment energy of less than 1 Mev. By direct observation of the tracks of these α particles in a cloud chamber using helium at lowered pressure in order to lengthen the tracks, their energy was found to correspond to $Q_1 = 1.93 \pm 0.07$ Mev, in sufficient agreement with the value calculated from the r -rays.

Finally, ${}^{\pi}Ne^{20}$ stands for still another group of states of Ne^{20} out of which it passes into a metastable excited state of O^{16} , denoted by ${}^{\pi}O^{16}$, from which a radiative transition into the normal state is impossible, or nearly so. An O^{16} nucleus in this state can, however, pass into its normal state by the simultaneous emission of a positron and a negative electron, a process that slightly decreases the nuclear mass but changes neither the atomic number nor the mass number. The sum of the kinetic energy of the pair and their combined rest energy of 1 Mev was found by Tomlinson¹ to be 6.0 ± 0.2 Mev, so that ${}^{\pi}O^{16}$ lies 6.0 Mev above O^{16} .

The Q value Q_2 for this third reaction C was also determined from the α -particle energies by Becker, Fowler, and Lauritsen, using a proton energy of 1.274 Mev, just above the strong resonance value for pair emission at 1.22 Mev. No trouble seemed to be caused by α -rays due to reaction B, from which fact it was surmised that the r -rays

¹ TOMLINSON, *Phys. Rev.*, vol. 60, p. 159 (1941).

giving rise to the broad hump in the observed excitation curve near 1.2 Mev were not due to reaction *B* but were produced in some other way, perhaps by a noncapture excitation of the F^{19} nucleus. It was found that $Q_2 = 1.93 \pm 0.07$ Mev. The sum of Q_2 and the pair energy, 6.0 Mev, is 7.93 Mev, in good agreement with the total emission of energy as represented by $Q_0 = 7.95$ Mev.

Streib, Fowler, and Lauritsen also suggested that the coincident resonance peaks at 1.35 Mev (Fig. 216) might be due to the existence of still a fourth group of Ne^{20} states, denoted by $^{\alpha\pi}Ne^{20}$ from each of

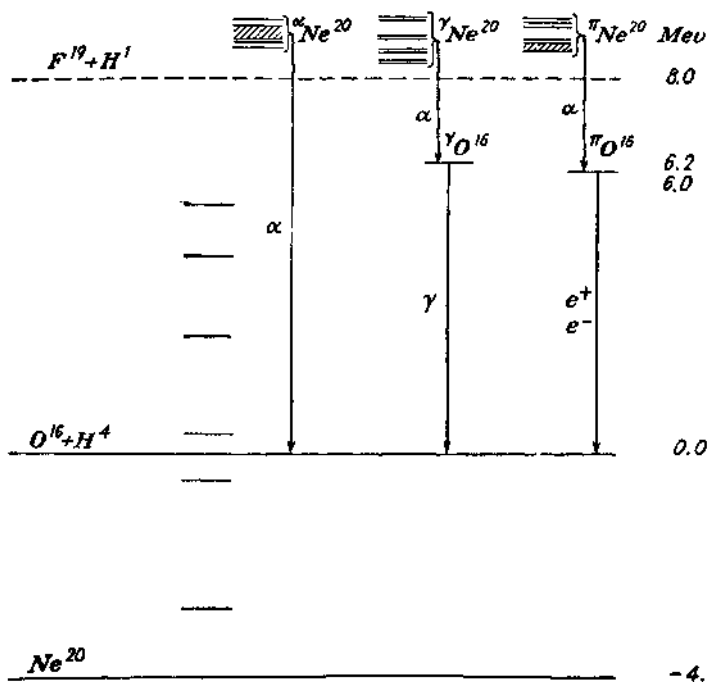


FIG. 217.—Levels of Ne^{20} involved in the reaction $F(p, \alpha)O$, and, at the left, six levels in which Ne^{20} is produced in the reaction $F(d, n)Ne$.

which the decay might be either to αO^{16} or to $^{\pi}O^{16}$; but confirmatory evidence seems to be lacking.

An approximate energy-level diagram illustrating the reactions just described is shown in Fig. 217. For completeness, the normal level of Ne is also shown, and, at the left, the six higher levels of Ne that were found by Bonner using deuterons as the bombarding agent ($F^{19} + H^2 \rightarrow Ne^{20} + n^1$).¹ In constructing the plot, values for the levels above that of $F^{19} + H^1$ were obtained by adding to the combined rest energies of fluorine and hydrogen $^{19}_{20}$ of the observed

¹ BONNER, *Roy. Soc., Proc.*, vol. 174, p. 339 (1940).

resonance energies of the bombarding protons, $^{19}_{20}$ being used in order to allow for the recoil kinetic energy of the neon nucleus.

227. Fission of the Nucleus.—There remains for discussion in the field of nuclear physics a discovery, announced early in 1939, which has already led to astounding practical consequences.

(a) *Discovery of Nuclear Fission.*¹—In 1934, Fermi tried the effect of neutrons upon thorium ($Z = 90$) and uranium ($Z = 92$) and found that β -ray activity was induced in both. The observed decay curve for the activated uranium was such as to point toward the existence of at least four different radioactive substances. Fermi found that he could separate two of these activities, with periods of 13 and 90 min., from the uranium by chemical means and thus prove that they were not isotopes of that element; and, by further tests, he satisfied himself that these substances were not isotopic with any known element having $Z > 85$. Since it did not seem possible at the time that these substances could differ greatly in Z from uranium, Fermi concluded that they were all new elements with $Z > 92$, which had been produced by one or more processes of β -decay from uranium atoms activated by the capture of a neutron.

Fermi's results were confirmed and extended, qualitatively, by others. Every experimenter seemed, however, to find a different set of periods. Furthermore, it soon became necessary to assume that the supposed "transuranic" elements ($Z > 92$), in decaying, gave rise to other radioactive elements which would be isomeric with heavy atoms already known. The number of different substances reported became so numerous and were interrelated in such a way that it was difficult to fit them into any plausible scheme.

Hahn and Strassmann then undertook a thorough investigation in the hope of establishing definitely the chemical nature of the new radioactive substances. In the course of this work, they discovered, in uranium irradiated with neutrons, a new substance with a β -ray period of 250 to 300 hr., which, according to elaborate chemical tests, was either radium ($Z = 88$) or barium (56). To settle the identity of this substance completely, they performed a series of fractional precipitations and crystallizations of the kind used for separating radium from barium. The activity definitely followed the barium, not the radium. After further experiments, Hahn and Strassmann finally felt justified in concluding that the radioactive substance was barium. To explain its origin, they proposed the radical hypothesis that *after capturing a neutron, a uranium nucleus may break up into two or more large fragments, each of the size of a*

¹ For a more detailed discussion of the early work see Turner, *Rev. Modern Phys.*, vol. 12, p. 1 (1940).

moderately heavy atom. The radioactive barium nuclei they interpreted as fragments produced in this manner by the division of uranium nuclei.¹

Once the possibility of this new type of disintegration, called by Meitner and Frisch "fission" of the nucleus,² was accepted, the facts fell rapidly into order, and fresh corroborations were rapidly obtained. Many of the fission products were recognized as substances already familiar in the study of induced radioactivity. Thorium and protactinium (Pa 91), also, were found to undergo fission when bombarded with neutrons. Radioactive isotopes of at least 25 elements, including Hg (80), Bi (83) and all from $Z = 35$ (bromine) to $Z = 57$ (lanthanum), have been identified among the fission products produced in thorium or uranium.³

The theoretical explanation of the process of nuclear fission itself presented an interesting problem. Why can it occur only in a few elements? As was remarked in Sec. 207, the nuclei of over half of the elements could break up into nuclei of lighter elements without violating the conservation of energy. Perhaps the answer is to be found along the lines of the following picture. We may think of the nucleus as being normally pulled into a spherical shape by the short-range attractive forces between neutrons and protons [Sec. 209(a)], somewhat as a suspended drop of water is made spherical by surface tension. If the nucleus became flattened through some cause, the potential energy due to these forces would be increased. The electrostatic potential energy due to repulsion of the protons would, on the contrary, decrease, but in a lesser degree, so that the equilibrium would be stable. If the deformation became large, however, the decrease in electrostatic energy would probably become predominant, and the deformation would then tend to increase further. If the nucleus actually became pinched apart into two fragments, both types of energy would decrease further as the fragments separated and rounded themselves off into spheres, and the final total energy might thus be less than the energy of the original nucleus. Now we have seen [Sec. 209(a)] that the ratio of electrostatic energy to short-range energy increases with increasing Z . Presumably for $Z > 89$ this ratio is so large that the nucleus has only a narrow margin of stability against change of shape.

(b) *Detection of Fission.*—The occurrence of fission is sometimes demonstrated, as in the original observation of Hahn and Strassmann,

¹ HAHN and STRASSMANN, *Naturwiss.*, vol. 27, pp. 11, 89 (1939).

² MEITNER and FRISCH, *Nature*, vol. 143, p. 239 (1939).

³ NISHINA, KIMURA, YASAKI, and IKAWA, *Zeits. f. Physik*, vol. 119, p. 195 (1942).

by identifying the products chemically. Another method of detection is to place the uranium, perhaps in the form of the oxide, on the wall of an ionization chamber and observe the huge pulses of ionization that are produced by the fission fragments as these are projected into the gas in the chamber at high speed.¹ Or, the heavy tracks made by these fragments in a cloud chamber may be photographed.² Finally, the fission fragments may be caught upon a receiving surface placed very close to the uranium and detected subsequently by means of their radioactivity.³ All these methods, except that of cloud-chamber observation, have been used repeatedly.

(c) *The Fission of Natural Uranium.*—The number of fissions produced in uranium was soon found to vary markedly with the speed of the incident neutrons. Fission can be produced either by fast neutrons, with energy of the order of 1 Mev or more, or by near-thermal neutrons, but not by neutrons of intermediate speed. The fission cross section of natural uranium for fast neutrons is about 0.1×10^{-24} cm.²; for thermal neutrons it averages about 2×10^{-24} cm.², varying with the neutron speed v , at least roughly as $1/v$.⁴

An interesting theoretical problem arose when these facts were considered in relation to the following phenomenon. Among the products produced in the bombardment of uranium with neutrons, Meitner, Hahn, and Strassmann had isolated a substance, chemically identical with uranium, which shows negative β -ray activity, with a period of 23 minutes. The yield curve for the production of this activity showed a resonance maximum at a neutron energy of 25 electron-volts. Now a uranium nucleus activated by a neutron of 25 ev energy ought surely to contain more energy and, hence, to be more likely to undergo fission, than a nucleus that has captured a neutron of thermal energy. Why, then, do not neutrons of 25 ev likewise cause fission in uranium, as neutrons with still more energy actually do? As an explanation, Bohr suggested⁶ that perhaps the thermal neutrons can cause fission only in the isotope U^{236} , which forms 0.71 percent of natural uranium, whereas the radioactivity just mentioned results from absorption of a neutron by the principal isotope U^{238} , as had been shown to be almost certain by Meitner, Hahn, and Strassmann. He supposes that fission of the product nucleus U^{239} , or of U^{288} itself, requires a neutron energy greatly in excess of 25 ev.

¹ Cf. several notes on pp. 416-418 of *Phys. Rev.*, vol. 55 (1939).

² JOLIOT, *Compte8 Rendus*, vol. 208, p. 647 (1939); CORSON and THORNTON, *Phys. Rev.*, vol. 55, p. 509 (1939).

³ JOLIOT, *Comptes Rendm*, vol. 208, p. 341 (1939).

⁴ Cf. *Phys. Rev.*, vol. 55, p. 512 (1939).

⁵ BOHR, *Phys. Rev.*, vol. 55, p. 418 (1939).

Bohr's explanation was soon shown to be the correct one by means of experiments upon samples of the separated isotopes, which were prepared in a mass spectrograph; the sample of U^{288} weighed 1.7×10^{-5} gram.¹ The corresponding *capture cross sections* have been measured. The capture cross section of U^{235} for fission has been found to vary in inverse proportion to the neutron speed and to be of the order of 2×10^{-24} cm.² for thermal neutrons.² The isotope U^{238} has a fission cross section of about 0.5×10^{-24} cm.² for neutrons of energy from 1 Mev to 10 Mev;³ fission ceases below about 0.35 Mev (in thorium, below 1.1 Mev).⁴

The *fission fragments* have been studied by various observers. Further observations were reported by Jentschke, with a thorough discussion of previous work.⁵ He allowed neutrons from a radon-beryllium source to pass through a screen of 6 cm. of lead and to fall upon a uranium foil 0.15 mm. thick; the two fission fragments from a disintegrating nucleus, emerging into the surrounding gas on opposite sides of the foil, produced pulses which were measured simultaneously with electrometers. From the total ionization as thus measured, the initial kinetic energies of the fragments were calculated, on the assumption that each ion pair in the gas, which was argon, required 27.5 ev for its production; each fragment was assumed to lose 1 Mev initially in traversing part of the thickness of the foil. The initial momentum of the uranium nucleus being negligible, the ratio of the observed initial kinetic energies of the fragments is also the inverse ratio of their masses. Since, furthermore, the sum of the masses must be equal to that of the uranium nucleus plus the incident neutron and minus any neutrons that are emitted in the act of fission, the actual masses of the fission fragments can be calculated.

The neutrons as they came from the source were "fast" and were found to produce four fissions, ascribed to the isotope U^{288} , per hour; when slowed down to thermal energy in paraffin, they produced 60 fissions per hour, presumably in U^{235} . The results were nearly the same in the two cases; hence only those obtained with fast neutrons will be described here.

Using the fast neutrons, it was found that the mass number of the lighter product nucleus varied from 79 to 114, 97 being slightly the commonest; that of the heavier, from 125 to 160, 142 being commonest.

¹ NIER, BOOTH, DUNNING, and GROSSE, *Phys. Rev.*, vol. 58, p. 475 (1940).

² See *Phys. Rev.*, vol. 55, p. 512 (1939).

³ AGENO *et al*, *Phys. Rev.*, vol. 60, p. 67 (1941).

⁴ HAXBY, SHOUPP, STEPHENS, and WELLS, *Phys. Rev.*, vol. 58, p. 199 (1940), for U; vol. 57, p. 1088 (1940), for Th.

⁵ JENTSCHKE, *Zeits.f. Physik*, vol. 120, p. 165 (1943). See also GRUMMITT and WILKINSON, *Nature*, vol. 158, p. 163 (1946).

The total kinetic energy of the products was observed to range from 133 to 192 Mev, with a commonest value of about 163 Mev. Jentschke remarks that, on the other hand, Japanese investigators, using neutrons of 17 Mev energy, were able to observe fissions in which the two fragments had nearly equal masses.

The kinetic energy of the fragment nuclei does not represent the whole of the energy released, of course, since part of it may be carried off by high-speed neutrons that are emitted (see below), or may be reserved for emission in association with subsequent radioactivity. Henderson measured the emission of energy calorimetrically and found an average value of 177 Mev per fission.¹ This would presumably include the energy of any quickly emitted β -rays, but might not include energy carried off by neutrons or γ -rays.

The enormous release of energy that must result from the breaking up of any heavy nucleus was discussed by Meitner and Frisch.² To illustrate the theoretical possibilities, a synthetic U^{239} nucleus, if no neutrons were lost, might be imagined to break up into two equal parts which would become tin nuclei, Sn^{119} and Sn^{120} , after the emission of 4 γ -rays each. The total energy emitted in such a fission, as estimated from the isotopic masses would be that corresponding to a loss of about 0.262 a.m.u. or 244 Mev, in addition to the kinetic energy of the incident neutron. Unequal division yields less energy; for example, division into Kr^{84} and Gd^{166} yields 0.225 a.m.u. or 209 Mev, in addition to the incident energy. Even this latter energy considerably exceeds the measured energies.

(d) *Other Means of Inducing Fission.*—Apparently any means of introducing enough energy into a fissionable nucleus suffices to cause fission. High-speed protons may be used.³ Deuterons are effective if their energy exceeds 8 Mev; the fission cross section at 9 Mev is 5×10^{-27} cm.² for uranium and nearly 7×10^{-27} cm.² for thorium.⁴ Either high-energy X-rays from an induction accelerator⁵ or γ -rays are effective, the fission cross section for 6-Mev γ -rays from a $F(p, \gamma)$ reaction being about 3.5×10^{-27} cm.² for uranium and 1.7×10^{-27} cm.² for thorium.⁶ Finally, α particles accelerated to 32-Mev energy in a cyclotron will induce fission in uranium.⁷

¹ HENDERSON, *Phys. Rev.*, vol. 58, p. 774 (1940).

² MEITNER and FRISCH, *Nature*, vol. 143, p. 239 (1939).

³ DESSATJER and HAFNER, *Phys. Rev.*, vol. 59, p. 840 (1941).

⁴ GANT, *Nature*, vol. 144, p. 707 (1939); JACOBSEN and LASSEN, *Phys. Rev.*, vol. 58, p. 867 (1940).

⁵ BALDWIN and KOCH, *Phys. Rev.*, vol. 67, p. 1 (1945).

⁶ HAXBY, SHOUPP, STEPHENS, and WELLS, *Phys. Rev.*, vol. 59, p. 57 (1941).

⁷ FERMI and SEGRE, *Phys. Rev.*, vol. 59, p. 680 (1941).

There is also evidence that fission occurs spontaneously in uranium and thorium, although at a very slow rate.¹ Maurer and Pose studied the slight spontaneous emission of neutrons from uranium, which they ascribed to fissions (see below). For this purpose a boron neutron counter 5 cm. in diameter and 35 cm. long was surrounded by a cylindrical mass of 9 kg. of uranium, and the whole was then immersed in thick paraffin to slow the neutrons down. Careful tests were made in order to rule out possible spurious effects, and, definitely to eliminate cosmic-ray effects, the observations were repeated by Pose in a deep mine. On the assumption that 1 neutron is emitted per fission, Pose found that, if uranium disintegrated only by fission, its half-value life would be 3.1×10^{15} years. For thorium the corresponding figure was 1.7×10^{17} years. If more than 1 neutron is emitted per fission, these numbers would be increased.

(e) *Emission of Neutrons*.—It was early found that fast neutrons are emitted in the fission process,² about 2 per fission on the average (the incident neutron being assumed to be captured).³

Furthermore, some of the emitted neutrons issue only after an appreciable lapse of time. These delayed neutrons, emitted with $\frac{1}{2}$ to 1 Mev of kinetic energy, appear to constitute a form of neutron radioactivity occurring in some of the fission products. They were studied by Snell, Nedzel, and Ibser,⁴ who surrounded a cylindrical shell containing 106 lb. of U_3O_8 with 2 in. of paraffin and also filled it with paraffin, except for an axial hole left for a counter filled with BF₃. The shell was bombarded with neutrons from a beryllium target bombarded by 8-Mev deuterons; after a certain time the bombardment of the shell was stopped and the decay curve of the neutron activity was noted, up to a maximum of 13 min. At least four different neutron activities seemed to be present; their half-life times and their relative intensities / when excited to saturation by sufficiently long bombardment, were as follows:

<i>L</i>	2.5 sec.	7 sec.	24 sec.	57 sec.
<i>I.</i>	1.2 sec.	1.2 sec.	1.2 sec.	0.135 sec.

¹ FLEROV and PETBZHAK, *Jour. of Physics U.S.S.R.*, vol. 3, p. 275 (1940); MATJRER and POSE, *Zeits. f. Physik*, vol. 121, p. 285 (1943); POSE, *Zeits. f. Physik*, vol. 121, p. 293 (1943).

² See SMYTHE, H. D., "A General Account of the Development of Methods of Using Atomic Energy for Military Purposes under the Auspices of the United States Government, 1940-1945," Princeton University, 1946; ANDERSON, FERMI, and HANSTEIN, *Phys. Rev.*, vol. 55, p. 798 (1939).

³ SZILARD and ZINN, *Phys. Rev.*, vol. 55, p. 799 (1939).

⁴ See SMYTHE, *loc. cit.*

It was estimated that 1 percent of all the neutrons emitted as a consequence of fission are delayed by "at least 0.01 sec, and 0.07 percent of them by at least 1 min.

(/) *Chain Reactions*.—The discovery that fission of a nucleus caused by a neutron results in the emission of more than 1 neutron opened up at once a startling possibility. If each neutron emitted could be made to cause another fission, resulting in the emission of more neutrons, a chain reaction would occur; the number of neutrons present would increase rapidly and the fission process would go on at an increasing pace until the whole of the fissionable material had been disintegrated. A tremendous amount of energy might thus be released in a small fraction of a second. In this way a bit of fissionable material might perhaps be made to act as an immensely powerful explosive; or, if the process could be controlled, we might have at last an atomic source of power.

Several factors, however, were seen at once to militate against the occurrence of a chain reaction. The neutrons may be captured by various nuclei without producing a fission; they will be slowed down by successive elastic impacts with nuclei; and they will tend to leave the scene by diffusing out through the surface of the active material. For several years it was not known whether a chain reaction was actually possible or not.

The problem continued to be studied by physicists, however, and, as every one knows, from 1942 onward it was attacked on a grand scale under the pressure of war urgency and with lavish support by the United States Government.¹ Qualitatively, the relevant scientific facts were already known in 1942, but the values of certain constants, especially capture cross sections, had not yet been measured with sufficient accuracy. Furthermore, it turned out that all operations had to be carried out on a large scale. For, the loss of neutrons through the surface of a mass of fissionable material becomes relatively less serious as the volume of the mass is increased, since surface increases with size less rapidly than does volume. For this reason, a block of fissionable material must contain at least a certain minimum mass if a chain reaction is to occur in it; and, because of the high penetrating power of neutrons, this minimum mass is not small. The ultimately successful utilization of chain reactions for the construction of bombs in 1945 may probably be rated the most stupendous feat of chemical or physical engineering that has ever been accomplished.

It was early realized that in pure natural uranium a chain reaction is not possible. Most of the neutrons will be slowed down by elastic

¹ SMYTHE, *loc cit*.

collisions with uranium nuclei before they can produce fission in U^{238} ; then, as they pass through the resonance region around 25 ev, many of them will be captured by U^{238} without causing fission, and thus only a fraction of them will be available to produce fission in U^{235} . To overcome these difficulties, it was proposed by Fermi and Szilard to arrange the uranium in small masses immersed in a substance composed of light nuclei that would be especially effective in slowing the neutrons down. With this arrangement, most of the neutrons produced in a block of uranium would wander out of it into the surrounding or "moderator" substance and would there be slowed down; if not captured or lost through the surface of the whole mass, they would then eventually find their way back into a block of uranium somewhere, with a high probability of causing fission in a U^{235} nucleus. For the moderator, carbon was suggested by Fermi and Szilard because its nuclei are light and at the same time have almost no tendency to capture neutrons. Alternative substances that have been considered are beryllium, hydrogen, and especially deuterium. Thorough elimination of impurities from both the uranium and the moderator is essential in order to minimize loss of neutrons by capture.

A successful chain-reacting "pile" of this sort was first got into operation at Chicago in December, 1942. It was made of rods of carefully purified uranium fitted into holes in a matrix of graphite, which had been purified more thoroughly than had ever been done before. In the graphite the neutrons had a mean free path of about 2.5 cm. and made something like 200 elastic collisions before finding their way back into a uranium rod. The pile was also surrounded by a thick layer of graphite in order to scatter as many as possible of the emerging neutrons back into it, but, even so, in order to keep the surface loss of neutrons within allowable limits, the pile had to be given a diameter considerably greater than the height of a man.

The fact that some of the fission neutrons are delayed in time makes it easy to control the operation of a uranium pile. To accomplish this, the pile is made only just large enough to sustain a chain reaction; the activity in it then builds up gradually because of the delayed neutron emission. The pile is divided by several removable screens of cadmium, which absorb the neutrons strongly. A single screen suffices to keep the pile inactive; as this last screen is drawn out past a critical point, activity begins, as evidenced by an evolution of heat and of neutrons, whose abundance is measured with a counter. The activity is unstable; if left to itself, an active pile will eventually run away and melt down; but the activity is easily held approximately at any desired level by making occasional slight adjustments of the cadmium screen.

A uranium pile of this sort derives its energy from fission of the rare isotope U^{236} and must cease action when all of this isotope has been destroyed, leaving only IT^{238} . The rate of release of energy in a single pile, in the form of heat, may run well up toward a million kilowatts. If technical difficulties can be overcome and the heat can be used to drive engines, uranium may be used as a nuclear source of power. The energy obtained from fission of the 235 isotope in 1 pound of natural uranium, if estimated as **200** Mev per fission, would be equal to that obtained from nearly 5,000 tons of coal. Whether the cost can be brought below that of coal is not yet known (in 1947).

The development of nuclear power presents a distinctly different problem from that of producing a nuclear explosive. For a practical explosive, two possibilities have been explored. One method is to separate out and utilize the isotope U^{235} , which is fissionable by slow neutrons. The other method is to utilize the fact that, when uranium is bombarded by neutrons, many of them are captured to form U^{239} , which disintegrates quickly, with emission of a negative β -ray, to form Np^{239} (Sec. 220); the neptunium then emits a second β -ray rather quickly and becomes an isotope of plutonium, Pu^{239} , which is only feebly α active and is thus for all practical purposes stable. Thus, as a pile operates, plutonium accumulates in it, indirectly at the expense of both U^{236} and U^{238} ; the plutonium may then be separated out chemically. Plutonium, also, is fissionable by slow neutrons.

A mass of U^{235} or of plutonium of sufficient size, but no heavier than a man can lift, is capable of undergoing a chain reaction, at least if suitably combined with a moderator. Such a mass can be used, as is now well known, as the active element in a bomb of catastrophic power. The mass must either be subdivided by a cadmium screen, which is jerked out when the bomb is to be fired, or disposed in two or more parts, which are quickly brought into contact at the desired moment. The reaction is halted in mid-course by expansion of the reacting mass, due to its own release of energy, but even so the energy released is equivalent to that released by the detonation of from 10,000 to 20,000 tons of trinitrotoluene (TNT). A further feature is that a blast of visible, ultraviolet, X-ray and γ -ray radiation is emitted, and also a blast of neutrons. Finally, in addition to radioactivity induced by the γ -rays and the neutrons in surrounding material, a large quantity of radioactive fission products is projected outward and then either precipitated on surrounding objects or carried off by surrounding water or air.

The fearsome effects of such nuclear bombs (universally miscalled "**atomic**" bombs) are well known. For further details **concerning their development the reader is referred to Smythe's report.**¹

¹ SMYTHB, *loc. cit.*

CHAPTER XII

COSMIC RAYS

The two most active fields of research in physics at the present time are nuclear research and the phenomena attending cosmic rays. Many of the features of cosmic rays lie more properly in the realm of meteorology than of physics, but within the last fifteen years some of these phenomena have turned out to possess novel physical features of supreme interest. Nature has, in fact, provided in the atmosphere a gigantic laboratory in which atoms are bombarded, not with projectiles of energies limited to a mere 5 or 10 million electron-volts, but with rays of various kinds having energies up to many billions of volts. In this chapter a short account will be given of the more prominent phenomena due to this bombardment.¹

228. Early Work on Cosmic Rays, (a) *Discovery of the Rays.*—It has been known for nearly 100 years that the air possesses a slight electrical conductivity. In 1899-1900 a careful study of the phenomenon was made by Elster and Geitel² and by C. T. R. Wilson.³ The method used was to mount an electroscope on an insulating support such as amber, inside a closed vessel, and to observe the rate at which electrical charge was lost by the electroscope. Wilson took the precaution of connecting the other end of the insulating support to a source of potential equal to the initial potential of the electroscope, so that leakage along the support would tend to maintain the charge; thus the observed loss of charge could only be due to its neutralization by ions collected out of the air. The conductivity of the air enclosed in the electroscope was found to be permanent, in spite of the continual removal of ions from it by the electrical field. From this fact it may be inferred that the ions are continually being regenerated in the air by some agency.

¹ For further information the student may be referred to F. Rasetti, "Elements of Nuclear Physics," 1936, and to various articles in *Rev. Modern Phys.*, vols. 10, 11, 13; also A. H. Compton, *Rev. Sci. Instruments*, vol. 7, p. 71 (1936); R. A. Millikan, "Cosmic Rays," 1939; PYRKOSCH, *Phys. Zeits.*, vol. 45, p. 148 (1944); and P. AUGER, "What Are Cosmic Rays?," 1945; "Cosmic Radiation," by W. Heisenberg, ed., Dover Publications, New York, 1946.

² ELSTER and GEITEL, *Phys. Zeits.*, vol. 1, p. 11 (1899); GEITEL, *Phys. Zeits.*, vol. 2, p. 116 (1900).

3 C. T. R. WILSON, *Cambridge Phil. Soc. Proc.*, vol. 11, p. 32 (1900).

For a time, the cause of this natural ionization of air was supposed to lie in traces of radioactive material residing in the apparatus or on its surroundings, or perhaps in the earth or the atmosphere. Other experimenters found that the ionization varied with the material of which the electroscope or ionization chamber was constructed, and that it could be partially screened off by surrounding the apparatus with a heavy layer of water or lead. In the expectation that the ionization should be much less at a considerable height above the earth, Gockel ascended in a balloon and made observations at various heights. He found that the ionization did decrease with increasing height, but only slightly.¹ In 1911-1914 Hess and Kolhorster extended "balloon observations"² to much higher altitudes and found that above a few hundred meters the conductivity *begins to increase again*. Kolhorster showed that the increase was continuous up to 9,000 meters, at which height 93 ions per cm.³ per sec. were produced in the ionization chamber attached to his electroscope, as against 13 at the earth's surface.

To explain the greatly increased ionization at these high altitudes, Hess proposed the novel hypothesis that it was caused by a penetrating radiation falling upon the earth from the outside. He remarked that this radiation could not be coming in large part from the sun, since the ionization was found to be sensibly the same by night as by day. Its penetrating power must be much greater than that of the hardest r-rays known. For the very hard r-rays from Th C", the value of the mass absorption coefficient μ/p , or absorption coefficient divided by density (Sec. 183), is 0.036 in Al or 0.041 in Pb, hence perhaps 0.035 in air. Thus a beam of these r-rays would be reduced to half strength in a thickness x of air at standard density given by $e^{-\mu x} = 1/2$, where $M/P = 0.035$ and $p = 0.001293$, or in a thickness $x = 153$ meters. In his balloon observations, Kolhorster found that the ionization decreased to half value in a descent of roughly 2,200 meters, at an average height where the density of the atmosphere is $5/12$ of its density at the surface. At normal density this would mean a decrease to half in about 1,000 meters, or $6\frac{1}{2}$ times the distance required for the r-rays.

The observations of Hess and of Kolhorster may be regarded as constituting the discovery of cosmic rays. They became known in Germany as "Hohenstrahlung" or "Ultrastrahlung"; the name, cosmic rays, is due to Millikan and Cameron (1925).

¹ GOCKEL, *Phys. Zeits.*, vol. 11, p. 280 (1910); vol. 12, p. 595 (1911).

² HESS, *Phys. Zeits.*, vol. 12, p. 998 (1911); vol. 13, p. 1084 (1912); vol. 14, p. 610 (1913); KOLHORSTER, *D. phys. Ges., Verhandlungen*, vol. 15, p. 1111 (1913); vol. 16, p. 719 (1914).

(b) *First Observations of Cosmic Rays under Water.*—An important step forward was taken by Millikan and Cameron¹ when, in 1928, improving upon earlier observations by Kolhorster,² they lowered sealed electroscopes to various depths below the surfaces of snow-fed mountain lakes, which should be especially free from radioactive contamination. The type of instrument used is described in the next section. Even above the water the ionization was observed to be less than over the land, presumably because r-rays from the earth were screened off by the water. As the electroscope was lowered into the lake, the ionization fell off rapidly in the first meter or so; thereafter it decreased more slowly with increasing depth, and less rapidly as the depth grew greater, in the manner illustrated by Fig. 218 below.

The ionization did not vary as a simple exponential function of the depth, however, as it would if it were caused by homogeneous radiation entering the water vertically from above. In part, the observed variation could be explained by assuming that the incident rays are distributed in direction; for rays entering the water obliquely would pass through more water and so would be weakened more in descending to a given depth than would rays entering vertically. Even allowing for this complication, however, the observations indicated that the rays causing the ionization at great depths were much harder (*i.e.*, more penetrating) than those that are responsible for most of the ionization at small depths.

Observations were made in two lakes in California, Arrowhead Lake, at an elevation of 5,100 ft. (latitude 34°), and Gem Lake at 9,080 ft. (latitude 38°). In Gem Lake, the ionization was still decreasing at 50 meters below the surface. When the ionization found in the two lakes was compared at the same depth below the surface in both lakes, it was seen to be less in the lake at the lower elevation. When, however, the comparison was made in a different manner, *viz.*, between points in the lower lake at a given depth below the surface and points in the upper lake so chosen as to be situated farther below the surface, the increase in the depth of water in the upper lake being equivalent in weight to the column of air between the elevations of the two lakes, then the observed values of the ionization were found to be in close agreement. Hence, by plotting cosmic-ray intensities, as measured by the ionization, against the total depth below the top of the atmosphere, expressed in equivalent meters of water, the observations in the two lakes could be represented by a single curve. This fact is explained

¹ MILLIKAN and CAMERON, *Phys. Rev.*, vol. 28, p. 851 (1926); vol. 31, p. 921 (1928).

² KOLHORSTER, *Preuss. Akad. Verhandlungen*, p. 366 (1923).

if we assume that (1) equal weights of water and of air absorb the rays equally and (2) the initial intensity of the rays as they enter the earth's atmosphere is the same in the two localities (which were 250 miles apart).

The part of the combined curve lying just above and just below the equivalent of sea level, treated as approximately exponential, gives a value of 0.28 per meter for the *apparent* absorption coefficient of the cosmic rays in water, or an apparent mass-absorption coefficient $n/p = 2.5 \times 10^{-3} \text{ cm.}^2 \text{ gram}^{-1}$. In air ($p = 0.001293$) the apparent absorption coefficient $/x$ would then be 0.32 per kilometer. Such numbers do not represent true absorption coefficients, however, for two reasons: (1) the rays are undoubtedly not homogeneous, and (2) they vary widely in their directions of motion.

These results reinforced so strongly the evidence from the balloon observations that belief in the existence of a penetrating radiation of extraterrestrial origin then became general. From this time on the number of workers in the field rapidly increased. Until about 1932, the main object of investigation was the magnitude of the ionization due to cosmic rays as a function of elevation, position on the earth, time of day, and so on. During this period the radiation itself was generally believed to consist of photons resembling r-rays but of much shorter wave length; and speculation was common as to the possible place of origin of such photons. Apparently they could not be coming from the sun or even from the stars; for the cosmic-ray ionization has always been found to be almost, or quite, uninfluenced by changes in the position of the sun or of the star masses composing the galaxy. Yet it has been estimated that the total amount of energy brought to the earth by cosmic rays is comparable with that brought in by starlight.¹

229. The Measurement of Cosmic-ray Ionization.—To measure the ionization, a string electrometer of very small capacity was usually used. Millikan and Cameron, in the work discussed in the last section, used an instrument of the Wulf type, the essentials of which are shown diagrammatically in Fig. 218.² The two fibers *FF*, of platinized quartz, hung from an insulating quartz rod, are charged by means of the charging wire *C* and tend to spread apart because of their mutual repulsion. The spreading is resisted by the downward pull of the quartz bow *B*. The positions of the wires are read by means of a microscope (not shown); the wires are kept in a plane perpendicular

¹ RBGENER, *Zeits. f. Physik*, vol. 80, p. 666 (1933).

² MILLIKAN and OTIS, *Phys. Rev.*, vol. 27, p. 645 (1926).

to the axis of the microscope by the attraction of induced charges on the metal wires *A A*.

Sometimes only one fiber is employed, opposite an earthed plate. In any case, the position of the fiber is either observed visually from time to time, or it may be projected upon a slowly moving photographic film driven by clockwork, so that the instrument is self-recording. In the latter case, the fiber may either be illuminated intermittently by an attachment connected to the clockwork, or a small spot on it may be made to produce a continuous trace on the film. The clock can also be made to charge the electroscope from time to time. Carefully made self-recording apparatus of this kind has been mounted on ships by Millikan and sent on long voyages over the ocean in order to compare the ionization at different geographical locations. The apparatus required no other attention than the periodic winding of the clock.

In any instrument for measuring ionization there is a zero effect due to residual radioactive contamination of the instrument itself. This zero effect has usually been determined either by sinking the instrument deep under water or by taking it down a coal mine, the residual ionization there observed being assumed to be due to radiation from the walls of the instrument. In later instruments, the residual effect was rendered unimportant by using large ionization chambers attached to the electroscope, so as to increase the volume of the enclosed gas in comparison with the area of the walls, and also by raising the pressure of the gas (*e.g.*, to 30 atmospheres). Often other gases than air are used, in which the ionization is greater. Reduction to air at standard density is then made by means of comparison tests using γ -rays from radium; it is assumed that the observed ionization varies in the same way with the density of the gas whether it is produced by cosmic rays or by γ -rays. In making measurements near the surface of the earth, the apparatus is commonly shielded with 5 to 10 cm. of iron or lead, in order to cut off radioactive rays from the surroundings.

In comparing ionization at different places, the same instrument is used at all locations. Thus the comparative results do not depend upon the assumption that the ionization per cubic centimeter is the same in instruments of different design.

The *absolute magnitude* of the ionization, or the number of ions

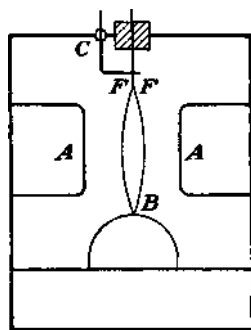


FIG. 218.—The electroscope of Millikan and Otis for measuring cosmic-ray ionization.

formed per cm.⁸ per sec, may be determined by an elementary calculation from the known capacity of the electroscope system and the observed rate of decrease in its potential. Millikan¹ obtained in this way a sea-level value of 2.5 ions per cm.³ per sec, in standard air, whereas Clay and Jongen² found 1.3, and Compton³ found 1.2. The cause of these discrepancies is not known. On the other hand, data on the ratios of the ionization at different places, obtained by different observers, agree very well.

230. The Altitude-depth Curve.—Until about 1930, the primary subject of investigation in cosmic-ray work was the variation of the

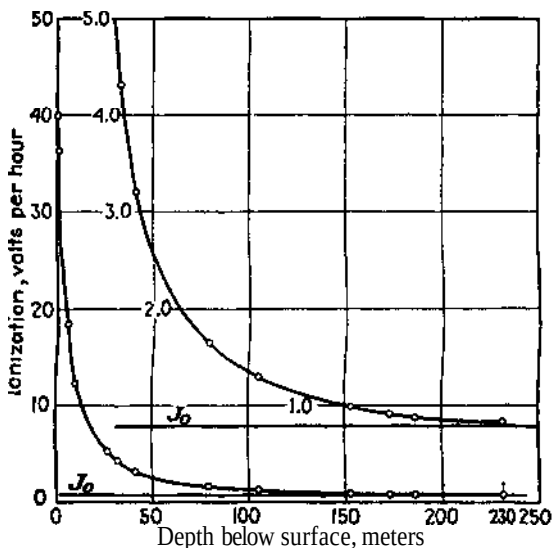


FIG. 219.—Variation of cosmic-ray ionization below the surface of Lake Constance (Regener). The ordinate gives the ionization in terms of volts lost per hour by the electrometer, the scale for the upper curve being 10 times as great as for the lower.

ionization as a function of altitude in the atmosphere or of depth under absorbing material. A few typical results may be cited.

(a) *Observations under Water.*—The most precise observations of cosmic-ray intensity under water are perhaps those of Regener.⁴ In Fig. 219 is shown his curve for the ionization at various depths below the surface of Lake Constance (altitude, 395 meters above sea level,

¹ MILLIKAN, *Phys. Rev.*, vol. 39, p. 397 (1931).

² CLAT and JONGEN, *Physica*, vol. 4, p. 245 (1937).

³ COMPTON and TURNER, *Phys. Rev.*, vol. 52, p. 799 (1937); cf. also COMPTON, WOLIAN, and BENNETT, *Rev. Set. Instruments*, vol. 5, p. 415 (1934).

⁴ REGENER, *Zeits. f. Physik*, vol. 74, p. 433 (1932); cf. also *Phys. Zeits.*, vol. 34, p. 306 (1933).

latitude, 47.5° N.). The upper curve repeats part of the lower one on a tenfold greater scale. We note that at 200 meters below the surface the ionization, reduced to 2 percent of its surface value, is still decreasing with increasing depth. The value labeled J_0 in the figure was presumed to represent the residual effect in the instrument. If J_0 is subtracted from the ordinates of the curve for the greater depths, the result is a fairly good exponential curve corresponding to a value of the mass-absorption coefficient (in water) of about $\mu/p = 0.19 \times 10^{-3}$. This is almost 200 times smaller than the value of $\mu/p = 0.035$ cited above, in air, for the hard r-rays from Th C". Thus the cosmic rays that cause ionization in deep water must be something like 100 times more penetrating than the hardest r-rays from radioactive substances. At 8 meters below the surface, the apparent value of n/p according to Regener's curve is about 0.76×10^{-3} . At a similar depth (10 meters below the surface at sea level) the data of Millikan and Cameron give about 0.9×10^{-3} .

(b) *Cosmic-ray Intensity at Various Altitudes above the Earth.*—As a typical example illustrating the variation of the ionization with altitude may be cited the results obtained by Millikan, Neher, and Haynes¹ near Fort Sam Houston in Texas (latitude 39°). Instruments were carried up into the stratosphere by sounding balloons, each instrument containing a recording electroscope and a recording barometer and thermometer, all operated by clockwork. The electroscope was recharged every 4 min. from a charged condenser, which was so well constructed that it lost by leakage only $\frac{1}{2}$ percent of its charge per hour. The instrument weighed 2.5 lb. and was carried aloft by five 1-meter balloons. As the intended top of the flight was approached, the rate of ascent was diminished by the bursting of two or three balloons; and after $3\frac{1}{2}$ hr. the instrument, automatically detached from the balloons, floated back to earth supported by a parachute. The parachute was of silk colored red, so as to attract attention, and an envelope offering a reward for the return of the instrument was attached. Four out of five instruments were eventually recovered, two within 12 hr., one from a distance of 80 miles.

The averaged results from two of these flights are shown by curve C in Fig. 224 (Sec. 234). The ordinate represents the number of ions per cm.⁸ per sec. that would be produced by the cosmic rays in air at standard density, which serves as a measure of their intensity. The actual ionization at high altitudes is much less because of the lowered density of the air. The abscissa in the figure represents the depth below the top of the atmosphere in equivalent meters of water (calcu-

¹ MILLIKAN, NEHER, and HAYNES, *Phys. Rev.*, vol. 50, p. 992 (1936).

lated as pressure in cm. Hg times 10.33/76, sea level being, therefore, at 10.33 on the axis of abscissas). The altitude itself can also be calculated from the recorded pressure, being about 92,000 ft. for the last point shown on the curve ($p = 1.29$ cm. Hg, 98 percent of the Avey to the top of the atmosphere).

The *pressure* should be a more significant quantity than the *altitude* for cosmic-ray work, however, since it gives directly the mass of air above the point in question. For this reason—to digress for a moment—the cosmic-ray ionization at a given point ought to vary with the barometric reading; and such is indeed found to be the case. From a long series of readings at different barometric pressures, Compton and Turner deduced¹ a decrease at Chicago of 1.6 percent for each increase of 1 cm. Hg in the barometric height. One might expect to obtain a similar value from the altitude curve. Since 1 cm. Hg means an increase in the effective depth of the atmosphere equivalent to 10.33/76 meter of water, the apparent absorption coefficient of the atmosphere at the earth's surface, as inferred above [Sec. 228(6)] from the observations of Millikan and Cameron, would lead us to expect a decrease per cm. Hg of $0.25 \times 10.33/76 = 3.1$ percent. This agrees only as to order of magnitude with the direct value of 1.6 percent.

Curve C in Fig. 224 reveals a maximum ionization of 240 ions per cm.³ per sec. (in standard air), as against about 2 at the earth's surface. The most interesting feature, however, is the *decrease* within the uppermost 5 percent of the atmosphere. This represents a real decrease in the ionizing power of the cosmic rays. Did the instruments, then, actually ascend above the *source* of the rays, which is, after all, in the atmosphere itself? A more probable explanation is believed to be the following. As the cosmic rays, whatever they are, enter the atmosphere, they produce secondary rays of some sort; the ionizing power of the secondary rays is greater than that of the primary, and hence, as the whole set of rays, primary and secondary, proceeds downward toward the earth, their combined ionizing power increases for a time. Eventually, however, the number of secondaries will reach an equilibrium value, and thereafter the ionization will decrease again, as all rays lose energy in collisions with air molecules or are stopped. What, then, are these mysterious rays that come out of the depths of space? And what is the nature of the secondaries that they produce? Are they photons? High-speed electrons? Protons? Or some new kind of particle, hitherto unknown to physics? Partial

¹ COMPTON and TURNER, *Phys. Rev.*, vol. 52, p. 799 (1937).

answers to these questions have been suggested by further observations now to be described.

231, Discovery of the Latitude Effect—In 1927, the Dutch physicist Clay made observations of cosmic-ray intensity during a voyage between the Netherlands and Java (7° S.) and observed that the intensity was distinctly less in equatorial regions than in Europe.¹ This variation, confirmed by him during later voyages and by other observers, is hard to explain on the assumption that the cosmic rays are photons; for there is no obvious reason why photons coming out of the depths of space should tend to avoid the equatorial regions of the earth. Just such a variation is to be expected, however, if the cosmic rays are *high-energy charged particles, perhaps electrons*. For a charged particle would be deflected by the magnetic field of the earth; if it were approaching the polar regions it would be moving more or less parallel to the magnetic lines and would undergo a minimum of deflection, but if its line of approach lay near the equatorial plane it would be deflected to a maximum extent and might be unable to reach the earth at all.

The general theory of the effect of the earth's magnetic field upon the motion of charged particles through it is complicated. For particles moving in the equatorial plane, however, the theory becomes very simple. Since the principal features of the motion are well illustrated in this special case, the appropriate theory may be developed here as an illustration. For this purpose, as is done also in treating the more general case, we shall idealize the earth's field somewhat by supposing it to be exactly like the field due to a magnetic dipole of a certain strength located at the center of the earth.

232. Motion of a Charged Particle in the Equatorial Plane of a Magnetic Dipole.—In the equatorial plane of a dipole, which is a plane drawn through the dipole and perpendicular to its magnetic axis, the magnetic lines are everywhere perpendicular to the plane, and the field strength is

$$H = \frac{M}{r^3} \text{ oersteds,}$$

where M is the magnetic moment of the dipole and r is the distance from its center. In this field a moving particle carrying a charge of e electromagnetic units, positive or negative, experiences a force

¹ CLAY, *Akad. Amsterdam, Proc.* vol. 30, p. 1115 (1927).

which deflects it sideways; to an observer looking in the direction of the magnetic lines (or *northward* at a point on the earth's equatorial plane) the deflection is *counterclockwise* for a *positive* particle, *clockwise* for a *negative* one. (Cf. Fig. 220, where the plane of the paper represents the equatorial plane, the *d* pole being indicated by *M*. The curves *AA*, *BB* illustrate possible paths which might be followed in the direction of the arrows by positive particles if the magnetic field is directed away from the reader, or by negative ones moving in the opposite direction, or moving with the arrows provided the field is directed toward the reader.) Being perpendicular to the velocity of the particle, the force does no work, and hence *the kinetic energy of the particle and its speed v remain constant*.

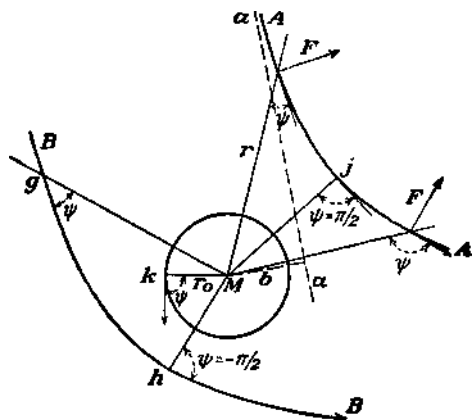


FIG. 220.—Diagram to illustrate the motion of a charged particle in the **equatorial plane** of a magnetic dipole *M*.

Let ψ denote the angle between the radius drawn from the instantaneous position of the particle to the dipole and the tangent to its path drawn in the direction of motion (Fig. 220); let this angle be measured positively in the usual counterclockwise direction from the former line to the latter (hence, negative in Fig 220 at *g* and *h*). The component of the momentum p of the particle perpendicular to the radius is $p \sin \psi$; and the moment of momentum of the particle about the center *M*, measured positively clockwise (contrary to the usual convention), is $rp \sin \psi$. Here at slow speeds we can write $p = mv$, but for energies of cosmic-ray magnitude it is necessary to use the relativistic formula. The component of the force F perpendicular to the radius, on the other hand, is

$$F \cos \psi = evH \cos \psi = \frac{evM}{r^3} \cos \psi.$$

This component of the force is to be positive when it is directed in the clockwise direction around M , and the sign of M is to be chosen to correspond to the other conventions (hence positive in Fig. 220 if the magnetic field is assumed to be directed away from the reader). Hence, by the usual principle connecting moment of momentum and moment of force,

$$\frac{d}{dt}(rp \sin \psi) = rF \cos \psi = \frac{evM}{r^2} \cos \psi.$$

But

$$\frac{dr}{dt} = -v \cos \psi. \quad (330)$$

Hence, by substituting in the preceding equation,

$$\frac{d}{dt}(rp \sin \psi) = -\frac{eM}{r^2} \frac{dr}{dt};$$

and, upon integrating, we have for the moment of momentum,

$$rp \sin \psi = \frac{eM}{r} + pb, \quad (331)$$

where we have written pb arbitrarily for the constant of integration. Thus we can write

$$\sin \psi = \frac{r_0^2}{r^2} + \frac{b}{r}, \quad r_0 = \sqrt{\frac{eM}{p}}. \quad (332a,b)$$

The reason for the choice of notation will appear presently.

Equation (332a) fixes the slope of the trajectory at every point in terms of r . Hence, in general, if one position of the particle is known, the trajectory can be traced. If we introduce a new variable, $p = r/r_0$, and a new constant, $p = b/r_0$, Eq. (332a) becomes $\sin \psi = p^{-2} + \beta p^{-1}$. It is thus made evident that a single set of curves for various values of p will serve to represent, in terms of r_0 as the unit of distance, the trajectories for particles of all masses and energies.

Ambiguity arises, however in tracing a trajectory past a point where $\sin \psi = \pm 1$, because there $\sin \psi$ ceases to vary with r and hence r is not given with sufficient completeness by Eq. (332a). In geometrical terms, it is evident from (332a) that, from any point where $\sin \psi = \pm 1$, either the trajectory may be continued with $\sin \psi$ returning toward zero, or, alternatively, the trajectory may be continued along a circle about M , with fixed r and with $\sin \psi$ fixed at one of the two values, $\sin \psi = \pm 1$. Thus through certain points two curves pass which are both solutions of Eq. (332a) for the same values of r_0 and

b. To ascertain how a moving particle goes past such a point, we may employ the time rate of change of ψ as obtained by differentiating Eq. (332a):

$$\cos \psi \frac{d\psi}{dt} = - \left(\frac{2r_0^2}{r^3} + \frac{b}{r^2} \right) \frac{dr}{dt},$$

whence, by (330),

$$\frac{d\psi}{dt} = v \left(\frac{2r_0^2}{r^3} + \frac{b}{r^2} \right). \quad (333)$$

The algebra does not actually give us this equation at an instant when $\cos \psi = 0$, but the equation holds both before and after such an instant, and all quantities involved in it are continuous functions of the time, hence it holds also when $\cos \psi = 0$. Thus the equation may be used to discover how ψ varies with the time as the particle passes a point where $\sin \psi = \pm 1$.

The circular paths or orbits, with center at M , represent "singular" solutions of the equations, which are possible when a certain relation exists between the constants r_0 and b . On a circular orbit $\psi = -\pi/2$, $\sin \psi = -1$, as at k in Fig. 219, and $d\psi/dt = 0$; by (332a) and (333), these conditions are satisfied if

$$-1 = \frac{r_0^2}{r^2} + \frac{b}{r}, \quad \frac{2r_0^2}{r^3} + \frac{b}{r^2} = 0,$$

or if

$$r = r_0, \quad b = -2r_0. \quad (334a,b)$$

(If $\psi = \pi/2$, r is imaginary.) The existence of circular orbits can also be deduced by elementary methods.

In the applications we shall be concerned with a set of orbits characterized by various values of b and a fixed value of r_0 , which is defined physically by (3336). For such a set of orbits, the circular one, occurring when $b = -2r_0$, may be called the *critical circle*. Its radius r_0 diminishes with increasing momentum or energy of the particle.

For the application to cosmic-ray particles, furthermore, we are interested principally in orbits that extend to infinity. A particle at infinity will be approaching along a certain straight line (such as aa in Fig. 220); and the distance of this line from M , multiplied by p , will represent the initial moment of momentum of the particle about M . If, however, we make $r =$ in Eq. (331), we find for this initial moment of momentum the value pb . Hence, the constant b represents the distance of the line of approach at infinity from the location of the dipole at M .

The possible orbits now fall into two distinct classes according as $b \gtrless -2r_0$.

1. $b > -2r_0$.—The properties of the orbits for $b > -2r_0$ differ according as $b > 0$ or $b < 0$. Start $r = \infty$, $\psi = 0$ all y. If $b > 0$, Eq. (332a) shows that $\sin \psi$ increases as r decreases; ψ continually increases, in fact, up to the value, $\psi = \pi/2$, at which point the path is perpendicular to the radius (e.g., at j in Fig. 220). The corresponding value of r , symbol r_1 , represents the distance of closest approach to the dipole, along this particular path, or the distance to the "perigee" of the orbit; its value may be found by putting $r = r_1$ and $\sin \psi = 1$ in Eq. (332a):

$$1 = \frac{r_0^2}{r_1^2} + \frac{b}{r_1}; \quad \therefore r_1 = \frac{1}{2}(b + \sqrt{b^2 + 4r_0^2}). \quad (335a,b)$$

(We reject the other root because it is negative.) At this point $d\psi/dt > 0$, as is evident from (333) and (335a); hence the orbit swings outward again and recedes back to infinity; in fact, it is symmetrical about the perigee. (Cf. orbit F in Fig. 221, where, as also in Fig. 220, the orbits are only very roughly correct.) Since, from (335b), for $b > 0$, $r_1 > r_0$, none of these orbits for $b > 0$ penetrates the critical circle. The orbits for $b = 0$ just touch it, with $r_1 = r_0$ (e.g., E in Fig. 221).

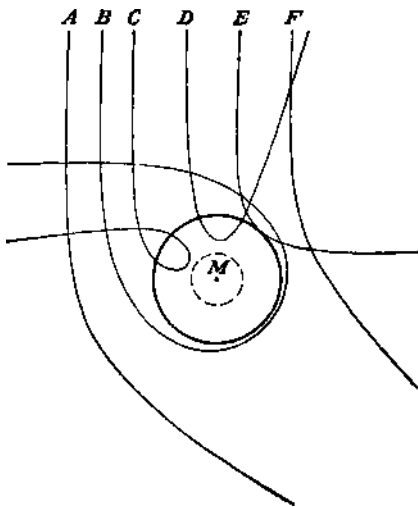


FIG. 221.—A few orbits (only roughly correct) for a charged particle in the equatorial plane of a magnetic dipole M .

If $-2r_0 < b < 0$, on the other hand, the term b/r in Eq. (332a) predominates at first over the term r_0^2/r^2 and $\sin \psi < 0$. Hence

decreases at first (cf. orbits C , D in Fig. 221), reaching a minimum value when $d\psi/dt = 0$ or, by (333),

$$0 = v \left(\frac{2r_0^2}{r^3} + \frac{b}{r^2} \right), \quad r = -\frac{2r_0^2}{b},$$

and, from (332a),

$$\sin \psi = -\frac{b^2}{4r_0^2}.$$

Since we are at present assuming that $b > -2r_0$, so that $|b| < 2r_0$, this last equation for $\sin \psi$ defines a real negative angle. Thereafter ψ increases again, and the orbit, after crossing the critical circle, passes perigee at $r = r_1$ as given by Eq. (335b) and then returns to infinity. If b is close to $-2r_0$, however, the minimum of ψ is close to $-\pi/2$, and at this point r is close to r_0 , and also dr/dt , which equals $-v \cos \psi$, is numerically much smaller than v ; hence the particle may revolve several or many times about the critical circle before plunging inward and reversing its direction of revolution.

2. $b < -2r_0$, $|b| > 2r_0$.— In this case r and ψ both decrease until ψ becomes $-\pi/2$, at which point, by (332a), $r = r_2$, where

$$-1 = \frac{r_0^2}{r_2^2} + \frac{b}{r_2}, \quad r_2 = \frac{1}{2}(-b + \sqrt{b^2 - 4r_0^2}).$$

Since $-b > 2r_0$, $r_2 > r_0$. Hence in (333), at $r = r_2$,

$$\frac{2r_0^2}{r_2^2} < 2r_0, \quad b < -2r_0, \quad \therefore \frac{d\psi}{dt} = \frac{v}{r_2^2} \left(\frac{2r_0^2}{r_2} + b \right) < 0.$$

Thus ψ continues to decrease. The orbit is, in fact, symmetrical about $r = r_2$ as perigee. Since $r_2 > r_0$, all orbits coming from infinity for $b < -2r_0$, like those for $b > 0$, lie entirely outside the critical circle (e.g., orbits *A*, *B* in Fig. 221). For these same values of b it can be shown that there exists another set of orbits which lie entirely inside the circle and are periodic; but with these we are not concerned.

For the applications to cosmic-ray particles, the feature of greatest interest about the orbits just described is the *least distance from the center*, $r_{\min}$, to which the particle can approach for a given value of r_0 but for any value whatever of b . This will be the smallest possible value of r_1 for $b > -2r_0$. From (335b), we find that

$$\frac{dr_1}{db} = \frac{1}{2} \left(1 + \frac{b}{\sqrt{b^2 + 4r_0^2}} \right) > 0 \quad (336)$$

for any value of b , since $b/\sqrt{b^2 + 4r_0^2} > -1$ even if $b < 0$. Hence the smallest value of r_1 occurs for the smallest admissible value of b , and, putting $b = -2r_0$ in (335b), we find

$$r_{\min} = (\sqrt{2} - 1)r_0 = 0.414r_0. \quad (337)$$

The actual orbit for $b = -2r_0$, to be sure, does not come so close as this to the dipole; it is a spiral wrapping endlessly **about** the critical circle **and** approaching it asymptotically **from** the outside. But, for

$b > -2r_0$, orbits can be found whose perigee distance lies as close as desired to

$r_{\min}$ as defined by Eq. (337).

A second point of physical interest is that, as stated above, *all orbits for* $-2r_0 < b < 0$ *cross the critical circle*; orbits of this type can be found, in fact, *crossing at all possible angles*. For, by (3356), $r_i = r_0$ at $b = 0$; and hence, in view of (330), $r_1 < r_0$ for any $b < 0$. The value of ψ as the orbit crosses, symbol ψ_0 , is found by putting $r = r_0$ in (332a):

$$\sin \psi_0 = 1 + \frac{b}{r_0}.$$

As b ranges from 0 to $-2r_0$, ψ_0 obviously ranges from $\pi/2$ to $-\pi/2$. At the same time, r_i ranges from r_0 down to $r_{\min}$ or $(\sqrt{2} - 1)r_0$.

233. Motion of Charged Particles in the Magnetic Field of the Earth.—With the mathematical results of the last section before us, we are prepared to discuss the conditions under which charged particles approaching the earth in its geomagnetic¹ equatorial plane will reach its surface. The results of the general theory will then be stated for any direction of approach, without proof. In all cases the earth's field will be imagined replaced by the most nearly equivalent dipole field.

(a) *Particles Approaching in the Geomagnetic Equatorial Plane.*—If the particles are moving too slowly, they cannot reach the surface of the earth at all. This happens when the momentum p of the particles is so small that $r_{\min}$, as given by (337) and (3326), is greater than the radius a of the earth. The value of $r_{\min}$ becomes just equal to the radius of the earth when the momentum p increases to such a value p_1 that

$$a = r_{\min} = (\sqrt{2} - 1)r_0 = (\sqrt{2} - 1) \sqrt{\frac{eM}{p_1}}, \quad (338)$$

$$p_1 = (\sqrt{2} - 1)^2 \frac{eM}{a^2}. \quad (339)$$

Here M is the earth's dipole moment and e is the charge on the particle, both in electromagnetic units; p_1 , e , and M may be assured to stand for numerical values.

As p increases above p_1 , r_0 and $r_{\min}$ decrease further and particles begin to strike the surface at small glancing angles. They will come from the west if they are positively charged, from the east if negatively charged (positive ones moving counterclockwise in Figs. 219 and 220, if the geographic north pole is below the plane of the paper). When p

¹ The geomagnetic equatorial plane of the earth is a plane perpendicular to the axis of that dipole which, placed at its center, best reproduces its magnetic field.

has increased to the value

$$p_2 = \frac{eM}{a^2}, \quad (340)$$

r_0 has shrunk to the value $r_0 = a$, as is evident from Eq. (3326). All orbits crossing the critical circle are now intercepted by the earth; since we saw the above that orbits cross this circle at all angles, the particles may now strike the earth at all angles above the horizon. The same remains true at all higher values of the momentum. Thus there are two critical values of the momentum p ; if $p < p_h$ no particles reach the earth; if $p > p_2$, they may strike it at any angle.

If we wish to speak in terms of the velocity v of the particle in place of its momentum p , we may calculate v from the relativistic formula, Eq. (GO) in Sec. 63,

$$p = \frac{mv}{(1 - v^2/c^2)^{1/2}},$$

m being the rest mass.

The velocity of cosmic-ray particles is nearly that of light, however, and hence varies little. It is more useful to introduce the kinetic energy E . According to Eq. (296) in Sec. 184

$$E = mc^2 \left[\left(1 + \frac{p^2}{m^2 c^2} \right)^{1/2} - 1 \right]. \quad (341)$$

At small energies, *i.e.*, if $p \ll mc$, this reduces to $E = p^2/2m = mv^2/2$. If $p \gg mc$, or $E \gg mc^2$, the approximation

$$E = pc - mc^2 \quad (342)$$

is useful.¹ The kinetic energies E_w and E_{20} that correspond to the critical momenta p_1 and p_2 can be found by substituting for p in Eq. (341) the values of p_1 and p_2 as given by Eqs. (339) and (340), *viz.*:

$$\text{For } E_{10}, \quad p = (\sqrt{2} - 1)^2 \frac{eM}{a^2}; \quad \text{for } E_{20}, \quad p = \frac{eM}{a^2}.$$

If $E < E_{10}$, no particles can reach the earth; if $E > E_{20}$, they can fall upon it at any angle to its surface.

If we insert in these formulas, for the earth,²

$$M = 8.04 \times 10^{25} \text{ e.m.u.}, \quad a = 6,370 \text{ km.},$$

¹ Equation (342) is obtained by expanding thus:

$$\left(1 + \frac{p^2}{m^2 c^2} \right)^{1/2} = \frac{p}{mc} \left(1 + \frac{m^2 c^2}{p^2} \right)^{1/2} = \frac{p}{mc} + \frac{mc}{2p} + \dots$$

² "Smithsonian Physical Tables," 1034.

and also $c = 2.998 \times 10^{10}$, $e = 4.803 \times 10^{-10}/2.998 \times 10^{10}$ e.m.u., $m = 0.9107 \times 10^{-27}$ gram for electrons or $m = 1.0076/(6.023 \times 10^{23})$ grams for protons, we obtain:

For electrons, $E_{10} = 10.2 \times 10^9$ ev; $E_{20} = 59.4 \times 10^9$ ev.

For protons, $E_{10} = 9.3 \times 10^9$ ev; $E_{20} = 58.5 \times 10^9$ ev.

These are enormous energies; and it is notable that at such high energies the difference in mass between proton and electron makes very little difference in the deflection of the particles by a magnetic field.

(b) *Particles Approaching from any Direction.*—For the theory of the general motion of electrons in the earth's field, there are available the studies made long ago by Stormer,¹ who ascribed the aurora to excitation of the earth's atmosphere by electrons coming from the sun. The application of the theory to cosmic rays was discussed by Epstein,² who pointed out that the variation of inertial mass with velocity does not cause complications, and this application was developed in detail by Lemaître and Vallarta³ and others.⁴ The results of the rather complicated analysis can easily be summarized.

It will be assumed in the remainder of this section that the cosmic-ray particles of a given energy approach the earth equally in all directions. Then, as Lemaître and Vallarta point out, they remain equally distributed in direction in spite of their deflection by the magnetic field.

If the earth's field is that of a magnetic dipole, there exists a family of periodic orbits in which electrons might be trapped permanently; one of these periodic orbits is the critical circle in the equatorial plane as described above. Most of the orbits, however, extend to infinity.⁶ Critical energies E_1 and E_2 occur in all cases, corresponding to those described above, but these energies decrease with increasing geomagnetic latitude⁶ X on the earth. At the poles, $E_1 = E_2 = 0$. That

¹ Cf. STORMER, *Terr. Mag.*, vol. 22, pp. 23, 97 (1917); vol. 35, p. 193 (1930); also *Nat. Acad. Set. Proc.*, vol. 17, p. 62 (1931), where references to the original publications are given.

² EPSTEIN, *Nat. Acad. Sci. Proc.*, vol. 16, p. 658 (1930); vol. 17, p. 160 (1931).

³ LEMAITRE and VALLARTA, *Phys. Rev.*, vol. 43, p. 87 (1933); vol. 47, p. 434 (1935); vol. 49, p. 719 (1936); vol. 50, p. 493 (1936).

⁴ For references see Johnson, *Rev. Modern Physics*, vol. 10, p. 193 (1938).

⁶ Some interesting figures and photographs of electron paths are given by Brüche, *Zeits. f. Physik*, vol. 31, p. 1011 (1930).

⁸ Geomagnetic latitude is measured from a plane perpendicular to the axis of that magnetic dipole, situated at the center of the earth, whose magnetic field most nearly approximates that of the earth. A better approximation can be obtained by displacing the dipole some 200 miles toward the western equatorial Pacific

$E_i = 0$ is evident without calculation; for particles of any energy can easily reach the poles, if the field is exactly that of a dipole—they have only to follow the axial line of force, thus undergoing no deflection. At points on the equator ($\lambda = 0$), the values of E_1 and E_2 are a maximum and are equal to E_{10} and E_{20} as given above.

At a given point on the earth's surface, in geomagnetic latitude λ , particles with energy less than the value of E_1 proper to that latitude cannot arrive at all. As the energy is raised above E_1 , particles begin to arrive at the top of the atmosphere within a certain cone of directions near the horizon; this cone opens from the west if the particles carry a positive charge, from the east if they are negative. As the energy rises, the cone enlarges. It consists of a principal part and a number of narrow bands; within it the rays arrive with the full intensity with which they approach from the depths of space. When the energy reaches the second critical value, E_2 , the cone fills the whole sky; particles now arrive with equal intensity from all directions, and with the same intensity as they would if the magnetic field were absent. No further change in the arriving particles then occurs with further increase in the energy.

Or, viewing the distribution of the particles over the earth, we may summarize as follows. Charged particles with energy less than a lower limit E_{10} cannot reach the earth's surface on the geomagnetic equator, although they may reach it elsewhere; $E_{10} = 10.2 \times 10^9$ ev (electron-volts) for electrons, 9.3×10^9 ev for protons. Particles with energy exceeding a second critical value E_{20} arrive at all parts of the earth's surface from all directions and hence with full intensity, just as if the magnetic field were absent; $E_{20} = 59.4 \times 10^9$ ev for electrons, 58.5×10^9 ev for protons. As we move north or south from the equator, the minimum energy for arrival, E_1 , and the minimum energy for full intensity, E_2 , both decrease; between latitudes 30° and 50° , E_2 decreases especially rapidly, until at magnetic latitudes above 60° even electrons or protons of 2 billion electron-volts will arrive from all directions without diminution by the magnetic field.

Further details concerning the intensity of the arriving particles are furnished by Fig. 222, copied from one of the papers by Lemaitre and Vallarta. The abscissa λ represents the geomagnetic latitude, measured from the geomagnetic equator. The ordinate represents

Ocean. The actual observed magnetic poles, where the earth's field at its surface is vertical, differ considerably in position from the locations of the geomagnetic poles, or those points at which the geomagnetic latitude is 90° ; e.g., the *observed* north magnetic pole is located at 70° N., 96° W., whereas the *geomagnetic* north pole is at 78° N., 69° W.

the total intensity of the arriving cosmic-ray stream, expressed as percent of the maximum possible intensity; it also represents the fraction of the sky from which the rays come, the number per unit solid angle from this part of the sky being the same as it is outside of the earth's field. The curves drawn in the figure refer to particles having given values of $p = a/r_0$, as indicated near the center of each curve [a — radius of the earth, r_0 = radius of the critical circle, defined by Eq. (3326)]; the corresponding energies in terms of billions of electron-volts are given on each curve for two kinds of particles, those for electrons preceded by the letter *e* and those for protons by *p*.

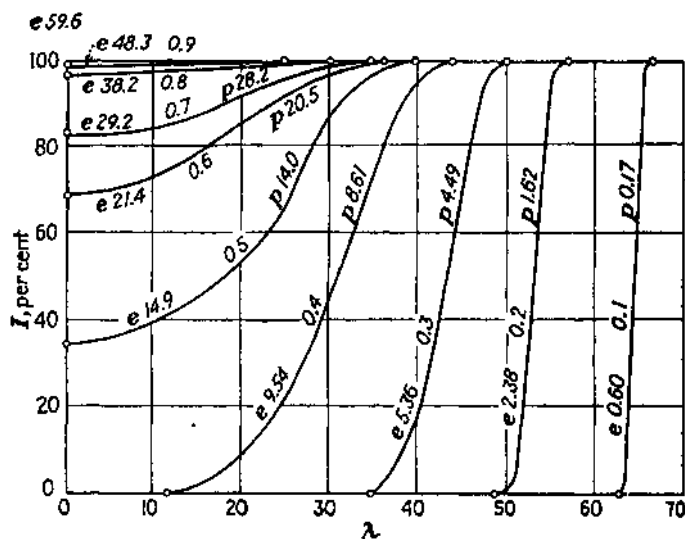


FIG. 222.—Dependence of cosmic-ray intensity on magnetic latitude. (For explanation see text.)

In using such curves in the interpretation of the observed ionization due to cosmic rays, however, it must be remembered that rays entering the earth's atmosphere obliquely have farther to go through it before reaching a given altitude above the earth, and thus will be weakened by absorption much more than are rays that enter vertically. The centers of the curves in Fig. 222, at $I = 60$ percent, should give a fair idea of the latitude at which particles of given energy produce half of their maximum possible effect; but the increase of the effect due to such particles with increase in the latitude will probably occur mostly between the latitudes corresponding to $J = 25$ and $I = 75$ percent.

In these calculations any possible deflection of arriving particles due to interference by the earth's atmosphere has, of course, been ignored. The theoretical results on the motion of charged particles

in the magnetic field of the earth may now be compared with the distribution of cosmic rays over the earth as observed.

234. The Variation of Cosmic-ray Intensity with Latitude.¹—In 1930, A. H. Compton, Bennett, and Stearns organized an extensive survey of cosmic-ray intensity at various locations on the earth. Observations were made at 69 stations by a number of cooperating observers, all using similar instruments calibrated on the spot by means of the ionization produced by a standard radium capsule. These observations confirmed the existence of the equatorial drop in intensity, as reported by Clay.

Compton pointed out, however, that his results correlated much better with *geomagnetic* latitude than with *geographic* latitude, and this has been confirmed by later work, especially by the extensive sea-level observations of Millikan and Neher. The latter observers, and Clay independently, discovered in 1934 that even the slight variation in the strength of the earth's magnetic field along the equator is reflected in the cosmic-ray intensity, which is some 4 percent lower in the East Indies and the Indian Ocean, where the magnetic field is strongest, than it is on the opposite side of the earth, in equatorial South America, near which the field is weakest.

All observations indicate, on the other hand, that the cosmic-ray intensity does not vary much from one location to another above geomagnetic latitude 40°, N. or S., which means in the United States above, roughly, 30° in geographic latitude. From a geomagnetic latitude of 40° N. or 40° S. to the equator, the drop in intensity amounts to about 10 percent on the average, rising to 12 percent in eastern Asia and sinking to 8 percent in the longitude of equatorial South America. The drop in intensity begins rather suddenly as the limiting latitude is passed.

The most accurate study of the equatorial drop is probably that made by Compton and Turner. They mounted a recording electro-scope on a steamship of the Canadian Australasian Steamship Company plying regularly between Vancouver and Auckland (New Zealand) and Sydney (Australia), the route crossing the geomagnetic equator almost where the latter crosses the geographic equator. The observations were extended over 10 months.

A decided *seasonal effect* was noted in these observations, especially at the extreme north and south ends of the range, where the cosmic-

¹ Cf. COMPTON, *Phys. Rev.*, vol. 43, p. 387 (1933); MILLIKAN and NEHER, *Phys. Rev.*, vol. 47, p. 205 (1935), vol. 50, p. 15 (1936); COMPTON and TURNER, *Phys. Rev.*, vol. 52, p. 799 (1937). Other references are given by JOHNSON, *Rev. Modern Phys.*, vol. 10, p. 193 (1938).

ray intensity during the warm months averaged less than during the cold months. In Fig. 223 are shown the curves obtained by Compton and Turner for the four seasons, ordinates representing average cosmic-ray intensity and abscissas the geomagnetic latitude. The seasonal variation, which has also been recorded by others, is believed to be correlated with atmospheric temperature; such an interpretation is supported by observations in Europe showing a variation of the same order of magnitude with temperature at a given station. (Temperature effects on the apparatus itself were shown not to occur.) From

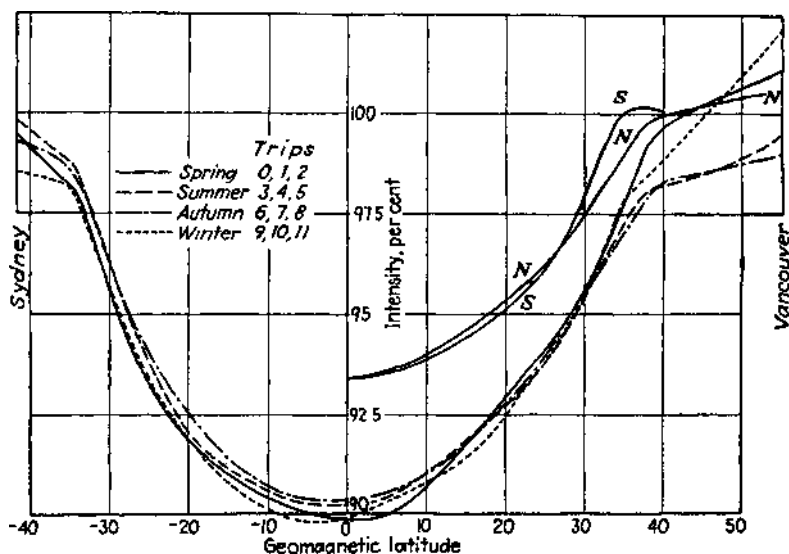


FIG. 223.—Variation of cosmic-ray ionization with latitude over the Pacific Ocean. The value called "100 percent" is chosen arbitrarily (and differently for the curves NN, SS). (See text.)

their data, Compton and Turner inferred a decrease of about 0.18 percent per degree centigrade of rise of temperature.

The temperature effect was believed by Compton and Turner to account for perhaps 3 percent of the observed decrease of 10 percent from high latitudes to the equator. Correcting for it, they found as the true geomagnetic effect the curve labeled NS in Fig. 223, representing ionization corrected to a temperature of 21°C. In drawing this curve, the southern half SS has been folded back over the northern half NNN in order to show how nearly they agree.

The close correspondence thus revealed between variations in cosmic-ray intensity and the features of the magnetic field at the surface of the earth constitutes convincing evidence that at least

part of the primary cosmic rays must consist, not of photons, but of charged particles of some sort. The effect of the magnetic field upon the particles must be produced before they enter the earth's atmosphere; for the atmosphere is so thin relative to the earth's dimensions that any effect produced within it could have little influence on the large-scale distribution of cosmic-ray intensity over the earth. If all of the primary cosmic rays are charged particles, however, those which cause the ionization near the equator must have energies exceeding about 35 billion electron-volts (or $p = 0.75$), for otherwise, as is evident from Fig. 222, the latitude effect would be larger than it actually is. An alternative possibility, of course, would be that several different kinds of cosmic rays are striking the earth, the equatorial effect being due to an agency (photons?) which, for some reason or other, is little affected by the earth's field, whereas the latitude-sensitive part of the cosmic-ray ionization is due to charged particles of much lower energy.

From the curves of Fig. 222 it appears that particles of about $p = 0.37$, or electrons of 8 billion or protons of 7 billion electron-volt energy, would show little increase in their effect above latitude 40° . They should also be ineffective below 25° , however, so that the further drop in the ionization from that point to the equator must be due to the presence of particles of still higher energy. Thus we may conclude that the primary cosmic rays, whatever their nature otherwise, must include charged particles of energies ranging broadly upward from 7 or 8 billion electron-volts.

At higher altitudes than sea level, the latitude effect is observed to be much greater. As an illustration, altitude curves obtained in different latitudes by Bowen, Millikan, and Neher, with the use of sounding balloons, are shown in Fig. 224. The abscissa represents depth below the atmosphere in meters of water (1 atmosphere = 10.3 meters). We note that the maximum ionization decreases by 73 percent in going from high latitudes to the equator (as against a decrease of 8 to 12 percent at sea level). The decrease is about one half if the comparison is made at altitudes a third of the way down from the top of the atmosphere. The critical geomagnetic latitude, above which variation with latitude ceases, seems to be 50 to 55° in the upper half of the atmosphere, as against 40° at sea level.

The higher value of the critical latitude at higher altitudes indicates the presence at such altitudes of particles that can reach the earth at higher latitudes and hence have less energy than those that are responsible for the sea-level effect. Not much energy can be coming in carried by particles of less than 2 billion ev ($p = 0.2 \pm$ in Fig. 222),

however, otherwise the critical latitude would be higher still. The primary cosmic-ray spectrum must, therefore, have an effective lower limit in energy at about 2×10^9 electron-volts.

A quantitative analysis made by Johnson¹ points toward a broad energy distribution of the incident particles with a maximum around 10 billion volts and reaching well up toward 50 billion. The number

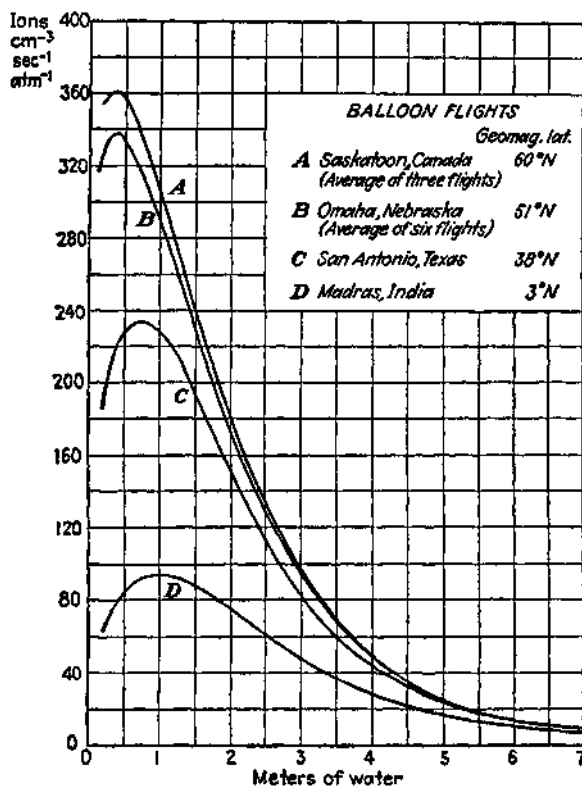


FIG. 224.—Cosmic-ray ionization as a function of altitude in four different latitudes. (Bowen, Millikan, and Neher.)

of particles, however, measured per unit of the energy axis, increases continuously with decreasing energy. The curve of total incoming energy falls off below 10 billion volts only because of the smaller energy carried per particle. Such calculations, however, must be regarded as tentative.

235. Observations on Single Cosmic-ray Particles.—The investigations described in preceding sections had to do with the average ionization produced by cosmic rays. Methods have also been invented

¹ JOHNSON, *Rev. Modern Phys.*, vol. 10, p. 193 (1938), especially p. 236.

for the observation of single particles, and such observations might be expected to furnish direct evidence concerning the nature of the rays. Even if some of the primary rays are high-energy photons, the ionization resulting from them will nearly all be produced by Compton electrons that have been ejected by the photons from the walls of the apparatus or from molecules of the air, and it should be possible to observe these secondary electrons.

Cloud-chamber observations of cosmic-ray particles were first made, accidentally, by Skobelzyn in 1927.¹ (For a description of cloud chambers see Sec. 203.) While studying the secondary β -rays produced by γ -rays from radioactive material, he noticed occasional tracks which were hardly curved at all in his magnetic field of 1,500 gauss. The ionization along these tracks was about as dense as along the tracks of the β -rays; therefore, he concluded that they also were made by electrons. From the smallness of the curvature of the tracks he calculated that the energy of the particles must be above 15×10^6 ev, which exceeds any energies encountered in radioactivity. In direction the tracks were strongly concentrated toward the vertical. Skobelzyn concluded, therefore, that these tracks were those of the particles that are responsible for the ionization ascribed to cosmic rays.

From the frequency of the occurrence of such tracks in his apparatus, Skobelzyn estimated that about 1.2 per minute crossed each square centimeter of a horizontal plane. Assuming a specific ionization of 40 ion pairs per cm. of track, he concluded that the particles would cause a total ionization of $1.2 \times \frac{4}{100} = 0.8$ pairs of ions per second in each cubic centimeter of air. This is of the same order of magnitude as the observed cosmic-ray ionization, 1.5 to 2 ion pairs per cm.³ per sec. A better modern estimate would be an average of perhaps 70 ion pairs per cm. of track [Sec. 236(6)] and 1.7 particles per cm.² per minute; these data give about 2 ion pairs per cm.³ per sec.

Counter observations constitute a second method for the detection of individual charged particles and have been widely adopted in the study of cosmic rays. Whenever a counter is used for any purpose, a slow background rate of counting is always evident. Part of this is doubtless due to radioactive contamination in the walls of the counter, for the background count varies with the material of the wall; but part of it is undoubtedly caused by cosmic-ray particles.

In order to eliminate effects due to rays from the walls of the counters, which are not very penetrating, Bothe and Kolhorster introduced the device of using *two counters* and recording only instances

¹ SKOBELZYN, *Zeits. f. Physik*, vol. 43, p. 354 (1927); vol. 54, p. 686 (1929).

in which both counters discharged at the same moment, which they called *coincidences*.¹ In their work the counter discharges were recorded on a moving photographic film running 1 cm. per sec, and a coincidence was considered to have occurred when two counts were recorded within 0.01 sec. of each other. It will occasionally happen, to be sure, that a coincidence occurs owing to the independent passage of two particles, one through each counter; but the number of such chance occurrences to be expected in a given time can be calculated from the total counting rates of the individual counters. The number of chance coincidences so determined was subtracted from the observed number as a correction.

The arrangement just described was used by Bothe and Kolhorster in an attempt to measure directly the absorbability of the cosmic-ray particles. Their apparatus is sketched in Fig. 225. The circles *C* represent in cross section two tubular counter chambers 5 cm. in diameter and 10 cm. long. Between these could be placed at will a (borrowed) block of gold 4.1 cm. thick to act as an absorber. The whole was surrounded by iron 5 cm. thick and also by lead 6 cm. thick, in order to screen off all local radiations and thereby permit only the highly penetrating cosmic-ray particles to activate the counters. Two blocks *Pb*, *Pb* of lead were placed close alongside the counters in an effort to stop particles that might pass through the upper one obliquely and then be scattered into the lower.

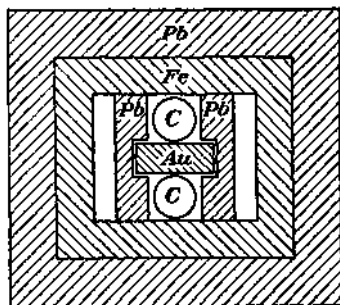


FIG. 225.—Apparatus of Bothe and Kolhorster for measuring the absorption of cosmic-ray particles in gold.

When observations were made with this apparatus near the roof of the building, about 3 coincidences per minute were observed, and this number decreased by 25 percent when the gold absorber was put in place. From this result, Bothe and Kolhorster deduced the value, $\mu/\rho = 3.5 \times 10^{-3} \text{ cm.}^2 \text{ gram}^{-1}$, for the mass absorption of the cosmic-ray particles. A similar experiment performed later by Rossi,² in which the cosmic rays were filtered through a layer of lead 5 cm. thick placed above the counters, gave $\mu/\rho = 1.6 \times 10^{-3}$ in a lead absorber 9.7 cm. thick. For comparison, we may cite the value deduced from Millikan and Cameron's curve for the (apparent) mass absorption coefficient of the cosmic rays in air at sea level, $\mu/\rho = 2.5 \times 10^{-3}$.

¹ BOTHE and KOLHORSTER, *Zeits. f. Physik*, vol. 56, p. 751 (1929).

² Rossi, *Zeits. f. Physik*, vol. 68, p. 64 (1931).

This is of the same order of magnitude as that found for the particles in the counter observations. The value deduced from the depth curve should be higher, as it actually is, partly because few of the rays were filtered out in making the measurements of the ionization, and partly because many of them would be traveling obliquely and so should decrease relatively more rapidly as the thickness of the absorbing layer of atmosphere is increased.

The counter method was improved later by Rossi in two respects. First, he replaced the photographic film by an electromechanical device for recording the discharges of the counters.¹ This was accomplished by connecting each counter to the grid of a three-electrode tube in such a way that the discharge of a counter momentarily lowers the potential of the grid and stops the current flowing through it. The tubes are connected in parallel and function as comparatively low resistances in a circuit that contains also a detector tube. When passage of electricity through the counter tubes is completely stopped by their functioning simultaneously, the grid of the detector tube rises momentarily in potential and so causes a pulse of current in its anode circuit, which causes in turn an audible click in a telephone, or may be made to actuate a mechanical counter of some sort. Such a counter circuit can be used to record coincidences between any number of counters.

The second improvement introduced by Rossi was the use of *three* counters in a row,² an observation being recorded only when all three counters respond simultaneously (or at least within a certain very small fraction of a second of each other). In this way coincidences due to secondary electrons ejected from the surroundings are effectively eliminated. When a primary particle passes through the first counter, only a secondary produced by it in the walls of this counter itself could pass through the other two counters; but the counter walls are very thin, and hence such secondaries should be very few in number. Equally rare should be the occasions when two secondaries are produced in the surroundings by the same primary particle and one of them passes through each of the other two counters. Two secondaries produced by a high-speed photon will also seldom cause a coincidence. A single secondary passing through the entire row of counters, however, could not be distinguished from a primary particle. Thus the arrangement selects charged particles moving within a narrow solid angle

¹ Rossi, *Nature*, vol. 125, p. 636 (1930). A good circuit for use with a single counter is described by Neher and Harper, *Phys. Rev.*, vol. 49, p. 940 (1936).

² Rossi, *Zeits. f. Physik*, vol. 82, p. 151 (1933). Cf. also TUVE, *Phys. Rev.*, vol. 35, p. 651 (1930).

of directions. Such an arrangement of counters is sometimes called a *counter telescope*.

Rossi also avoided screening the apparatus closely with lead all round, because, as his investigations showed, secondary particles are produced in such screens in large numbers and may seriously falsify the results.

Using three counters (Fig. 226) in a vertical line, Rossi¹ studied the absorption in lead of cosmic-ray particles which had already been filtered through 7 cm. of lead (placed above the counters). Moving one counter out of line was found to reduce the frequency of coincidences to that which was to be expected as a result of chance (1 in 26

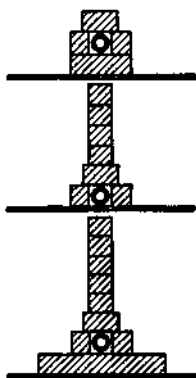


FIG. 226.—A "counter telescope" used by Rossi in studying the absorptibility of cosmic-ray particles.

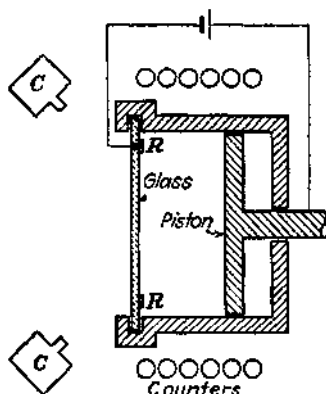


FIG. 227.—A counter-controlled cloud chamber, with stereoscopic cameras CC. An electric field is momentarily applied between the ring R and the piston to sweep the space clear of ions. [After Stearns and Froman, *Am. Phys. Teacher*, vol. 7, p. 79 (1939).]

hr. as against 1 to 2 per hour with the counters in line). He found that 46 ± 5 percent of the coincidences remained when 101 cm. of lead were placed between the counters, concluding that nearly half of the particles can penetrate such a thickness of lead. If we assume for purposes of calculation that the absorption is exponential through each successive layer of lead, the following values are found for the mean mass absorption coefficient: in the first 10 cm of lead, $\mu/\rho = 1.8 \times 10^{-3}$ (total absorption, 19 percent); in the next 15 cm., $\mu/\rho = 0.5 \times 10^{-3}$; and in the last 76 cm., $\mu/\rho = 0.55 \times 10^{-3}$.

These figures show unmistakably that the cosmic rays include charged particles of enormous penetrating power.

In another investigation, the two methods of observing cosmic-

¹ *Ibid.*

ray particles that have been described were brought into relation with each other in 1931 by Mott-Smith and Locher, who placed a cloud chamber between two counters and noted that tracks in the cloud chamber and coincident responses of the counters occurred together. This indicated that the two effects must be due to the same type of charged particle. In 1932, Blackett and Occhialini introduced the system, now generally employed, of allowing the coincident discharge of the counters to release the expansion of the cloud chamber. Such an arrangement is sketched in Fig. 227. The counters can be connected so that a particle passing through any chosen combination of them will release the mechanism that withdraws the piston and operates the cameras. In this way, much useless photographing is avoided (otherwise, it requires 10 to 20 random expansions to obtain a track), and only tracks having a suitable position and direction are photographed.

From 1930 on, many investigations of cosmic-ray particles were made by various observers. The interpretation of the results of such experiments presents special difficulties because of the enormous penetrating power of the particles and because of their powerful effects upon matter, and it is often hard to harmonize the results of different investigations. We shall discuss only certain points on which general agreement seems to have been reached, referring the student for further information to recent summarizing articles, such as appear from time to time in the *Reviews of Modern Physics* and elsewhere,¹ or to the literature.

The interpretation of recent work has often depended upon a precise knowledge of the behavior of high-speed particles. It will be convenient to summarize at this point what is known in regard to certain of their properties.

236. Energy, Mass, and Specific Ionization of Charged Particles.

(a) *Measurement of Charge, Mass, and Energy.*—The standard method of crossed electric and magnetic fields (Sec. 41) yields, for a charged particle, its velocity and the ratio of its charge to its mass. Additional observations of some sort are needed in order to find e and m separately; from m and the velocity, the energy can then be calculated.

When the velocity of the particle is almost equal to the velocity of light, however, the velocity becomes difficult to measure with sufficient accuracy to make the measurement significant. There is then no point in using an electric field to deflect the particle; the easier measurement with a magnetic field yields all the information that can readily be obtained.

¹ Cf. *Am. Phys. Teacher*, vol. 7, p. 79.

Formulas for the rapid motion of a charged particle in a uniform magnetic field were obtained in Sec. 184. If M is the mass of the particle, q its numerical charge in *electrostatic units*, and p the radius of the circle which it describes in a magnetic field of H oersteds, then, from Eqs. (294) and (297), the momentum p and kinetic energy E of the particle are given by the equations¹

$$p = \frac{qH\rho}{c}, \quad H\rho = \frac{cp}{q}, \quad (343a,b)$$

$$E = Mc^2 \left[\left(1 + \frac{(pc)^2}{(Mc^2)^2} \right)^{1/2} - 1 \right] = Mc^2 \left[\left(1 + \frac{q^2 H^2 \rho^2}{M^2 c^4} \right)^{1/2} - 1 \right]. \quad (344)$$

The quantity pc has the dimensions of an energy and may be expressed in electron-volts; if pc is given in c.g.s. units, its value in ev is obtained by dividing by the number of ergs in an electron-volt or 1.601×10^{-12} . The value of pc in ev is also the value of p itself expressed in ev/c as a unit. For a charge n times as large as the electronic charge, pc expressed in ev equals $(300n)\text{Hp}$.

If $E \ll Mc^2$, expansion gives, since $(1+x)^{1/2} = 1 + x/2 - \dots$,

$$E \ll Mc^2: \quad E = \frac{p^2}{2M} - \dots = \frac{q^2}{2Mc^2} H^2 \rho^2 - \dots \quad (345)$$

If we write E_{ev} for E expressed in electron-volts and m for the electronic mass, we have from the last equation, for a particle of mass M , numerical charge equal to that of the electron, and kinetic energy $E \ll Mc^2$, approximately,

$$E_{\text{ev}} = 0.088 \frac{m}{M} (H\rho)^2, \quad (346)$$

within 1 percent if $E < Mc^2/50$. If $E \gg Mc^2$, we have

$$E \gg Mc^2: \quad E = pc - Mc^2 + \dots = qH\rho - Mc^2 + \dots, \quad (347)$$

or, if q is n times as large as the electronic charge,

$$E_{\text{ev}} = 300nH\rho - (Mc^2)_{\text{ev}}, \quad (348)$$

within 1 percent if $E > 10Mc^2$. The coefficient 0.088 represents q^2/mc^2 divided by the value of the electron-volt or $\text{g}/300$, with $q = 4.803 \times 10^{-10}$ e.s.u., $m = 0.9107 \times 10^{-27}$ g, $c = 2.998 \times 10^{10}$. For the intermediate region, the plot shown in Fig. 228 may be of use in passing from one to another of the three quantities, velocity, $H\rho$, and E . Scales, logarithmic in both directions, are shown for both electrons and protons.

An important conclusion from these formulas is that the curvature

¹ Calculations from such formulas are most easily made in the following way:
 $E = Mc^2 [(\cos \tan^{-1} qH\rho/mc^2)^{-1} - 1]$.

of the path of a charged particle moving with energy greatly exceeding its rest energy ($E \gg Mc^2$) is almost independent of its mass, since $p = E/qH$ nearly. The energy of such a particle can be estimated, therefore, from the curvature of its track in a cloud chamber without knowing its mass. Partly for this reason, and partly because the particles were long supposed always to be electrons, the "energies" of cosmic-ray particles as determined from cloud-chamber observations are commonly reported on the assumption of an electronic mass. An

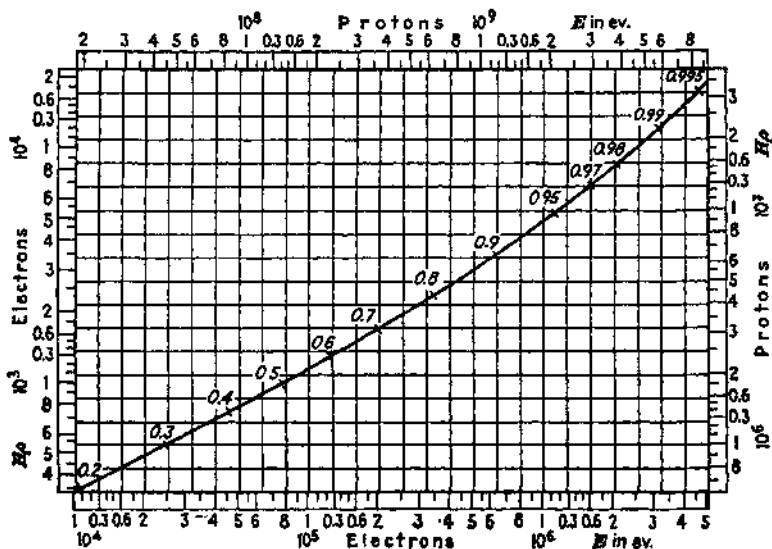


FIG. 228.—Relation between kinetic energy E in electron-volts and $H\rho$ (p — radius of curvature in centimeters in a field of H oersteds), for electrons and for protons, on logarithmic scales. Numbers on curve refer to the cross lines and give ratio of velocity of particle to that of light. For a particle k times as heavy as an electron, multiply $H\rho$ and E for an electron by k ; or, multiply by 100 and read on the scales n cross-section spaces to the right of or higher than the point on the curve, where for $K = 126, 158, 200, 251, 316$, respectively, $n = 1, 2, 3, 4, 5$.

energy so calculated might be called an "equivalent electron energy." A more precise method of reporting such data, sometimes employed, is to report the values of $H\rho$ or of p .

The curvature of the tracks in a cloud chamber thus gives primarily a measure of the momentum of the particle, from which a rough estimate of the energy can be made provided the energy is high enough. The mass must then be inferred in some other way. The principal source of evidence as to the mass, and also as to the magnitude of the charge, has been the amount of ionization caused by the particle in the gas in the cloud chamber. The sign of the charge can be inferred from the direction of curvature of the path, provided the

direction of motion of the particle is known. In the case of cosmic-ray particles, it may reasonably be assumed that tracks near the vertical were traversed in a downward direction. Or, if a track passes through a slab of absorbing material, it may safely be assumed, as in Anderson's discovery of the positron, that the particle traveled toward the side on which greater curvature of the track gives evidence of lower energy.

(b) *The Specific Ionization in Air.*—The ionization produced by electrons in air has been studied experimentally over a wide range of energies.^x The ionization produced by a cathode ray has been found to increase with increasing energy of the ray, up to a maximum of several thousand ion pairs per centimeter of path at an energy of 1,000 volts or so, above which it decreases again. It was found by W. Wilson² that for β -rays of energies from 6×10^4 to 1.7×10^6 ev the total ionization produced in standard air was almost inversely proportional to the square of the speed of the β particle, decreasing from about 214 ion pairs per centimeter of path at 6×10^4 ev to about 46 ion pairs at 1.7×10^6 ev.

The ions measured in these observations include those produced by secondary electrons ejected from molecules by the β particles with sufficient energy to produce additional ions. The *primary* ionization produced directly by the β particle itself can be determined separately by counting the drops of water along its path in a cloud chamber (C. T. R. Wilson). The path of most of the secondary electrons is so short that usually all of the ions formed directly or indirectly as the result of a particular primary ionization lie at first very close together, and, if the expansion is performed immediately after the passage of the β -ray, a single droplet of water is condensed on the entire cluster. The number of primary ions as determined in this way is about half as great as the total number of ions, at electron energies around 10^6 ev. On the other hand, the total ionization can be determined in the cloud chamber by delaying the expansion a fraction of a second (*e.g.*, 0.2 sec.) so as to give time for the ions to spread out by diffusion. If the expansion is great enough, each ion then gives rise to a separate droplet; or, alternatively, the expansion can also be so adjusted that droplets form only on the positive ions. Another scheme is to separate the positive and negative ions by means of an electrostatic field.³

As regards the appearance of tracks in the cloud chamber, however,

¹ See E. RUTHERFORD, CHADWICK, and ELLIS, "Radiations from Radioactive Substances," pp. 444ff., 1930.

² W. WILSON, *Roy. Soc. Proc.*, vol. 85, p. 240 (1911).

³ HAZEN, *Phys. Rev.*, vol. 65, p. 259 (1944).

a complication is presented by the fact that occasionally a secondary electron of exceptionally high energy makes what appears to be a branch track. The appearance of the tracks as a whole is determined by the average amount of ionization exclusive of that on such branch tracks; this ionization is called the *probable* ionization.

The term "specific ionization" is applied in a general way to any number representing ions or ion pairs per centimeter.

Approximate theoretical formulas for the ionization produced by charged particles of high energy have been derived from wave mechanics. The way in which the ionization is found to vary with the energy of the particle is illustrated by certain of the curves in Fig. 237 below [Sec. 238(c)]. Curves *e-cc* and *ep-ep* serve to illustrate the total and the probable ionization, respectively, produced by an electron, as a function of its energy, whereas the curves labeled *p* and *ml60* illustrate the total ionization produced, respectively, by a proton and by a "meson" (Sec. 240) 160 times as heavy as an electron, the medium being in all cases air of standard density. The quantity actually plotted is the rate of loss of energy per cm., but the number of ion pairs produced is nearly proportional to this, being about equal to the energy loss in ev per cm. divided by 32 (for air). References are given under the figure. The theoretical formulas seem to agree with observation, but they have not been checked thoroughly for electrons.

As shown by the curves in Fig. 237, the theoretical formulas predict a decrease in the ionization, due to the decrease in the time during which the particle can act on an atom, as its speed increases, up to approximately the speed of light, when the energy of the particle becomes of the order of its rest energy. Above this point the ionization should increase again as the energy of the particle increases, but very slowly, about as the logarithm of the energy; from 10^6 to 10^{10} electron-volts, the probable ionization should about double. The increase results from the increasing contraction of the electric field of the particle toward a plane perpendicular to its motion (Sec. 34), which increases the sharpness of the impulse given to atomic electrons by particles passing at some distance from the atomic center. A further factor, causing an increase in that part of the total ionization which is produced by secondaries of high energy, is the increase in the maximum energy which can be transferred to an atomic electron in a close collision.

Observations by Corson and Brode¹ on cosmic-ray particles,

¹ CORSON and BRODE, *Phys. Rev.* 53, p. 773 (1936). For a correction see Brode, *Rev. Modern Phys.*, vol. 11, p. 222 (1939), especially p. 228.

presumably electrons, with energies up to 30 Mev, gave evidence of **the** predicted increase. They found a (probable) specific ionization **of about** 50 ion pairs per cm. of path in standard air at the minimum (**just** above 10^6 ev). A similar value has been found for mesons.¹ The total ionization will then be somewhat greater. The theoretical formula indicates a minimum of about 50 ion pairs per cm. for the probable and 65 for the total ionization.

The most important feature of the theoretical formulas in their application to cosmic-ray observations is the way in which the ionization should vary with the kind of particle. The ionization should be roughly *proportional to the square of the charge*; it should thus be about the same for positive and negative particles. Otherwise it should depend, to a close approximation, only on the velocity of the particle. Since, at given velocity, kinetic energy and rest mass are proportional to each other, even in relativistic mechanics, it follows that *particles of equal charge but different masses should produce about the same ionization at equal velocities, or at kinetic energies proportional to their masses*.

Herein lies the reason for the relatively heavy ionization ordinarily produced by protons, which are usually moving comparatively slowly, with an energy of such magnitude that slower velocity means heavier ionization. During the last few millimeters of its path through standard air, a proton produces roughly 30,000 ion pairs per centimeter.² According to theory, however, a proton track, though clearly distinguishable from an electron track of the same energy up to energies of about 200 Mev, should actually be less dense than an electron track at energies exceeding 2,000 Mev. In general, a singly charged particle k times as heavy as an electron should produce about the same ionization as an electron moving at the same velocity and hence with an energy $1/k$ times as great as that of the given particle. Because of this feature, corresponding curves for particles of different mass in Fig. 237 are almost exactly alike in shape but are displaced relatively to each other along the axis representing the logarithm of the energy. A cloud-chamber photograph showing a meson and a proton track is reproduced in Fig. 229.

Since we have seen that the curvature of the tracks in a cloud chamber gives at once an approximate value of the energy of the particle, it is clear that, over a considerable range of energies, the density of the ionization along the track furnishes information in

¹ See DUNLAP, *Phys. Rev.*, vol. 67, p. 67 (1945); HAZEN, *Phys. Rev.*, vol. 67, p. 269 (1945).

² FEATHER, *Nature*, vol. 147, p. 510 (1941).

regard to the mass, provided the magnitude of the charge can be inferred in some way with sufficient certainty.

In *different materials*, theory indicates, with some support by experiment, that the total ionization should be roughly proportional to the total number of atomic electrons per cubic centimeter.

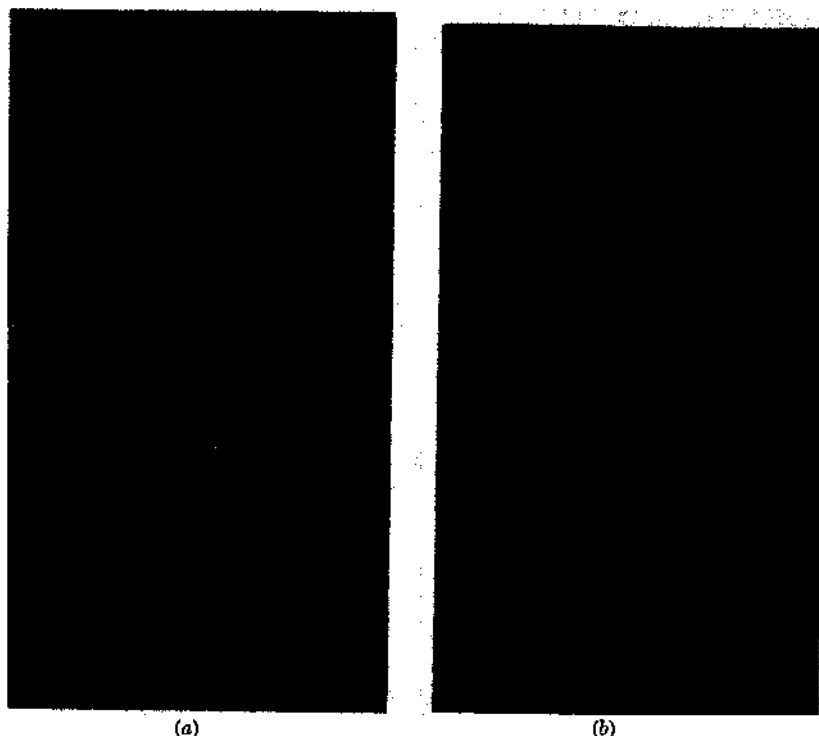


FIG. 229.—Photographs of cloud-chamber tracks in a magnetic field, made by (a) a positive meson having a mass roughly 250 times the electronic mass and a kinetic energy of about 4 Mev, and (b) a proton of energy about 3 Mev. [Photographs courtesy of Dr. D. J. Hughes. See *Phys. Rev.*, vol. 69, p. 371 (1946).]

237. Showers and Bursts. (a) *Showers.*—When cosmic-ray tracks are observed in a cloud chamber, it frequently happens that two or more tracks are seen which appear to have been produced simultaneously; often they diverge from a common center lying in the walls or outside of the chamber. The first observation of this sort was made by Skobelzyn, in the course of the work described above (1927; Sec. 235). If there is a solid obstacle in the chamber itself, such as a sheet of lead, groups of tracks may often be seen diverging from a point in this obstacle; frequently a single cosmic-ray track coming from above seems to end at the initial point of a group of tracks. (Cf. Fig. 231.)

Such groups of tracks are believed to be made by showers of (negative) electrons and positrons created as the result of the impact of a high-speed particle or photon upon matter. From a statistical study of the curvature of the tracks it is inferred that positive and negative particles occur about equally often, and with energies ranging from 1 Mev up to 500 Mev, 5 to 20 Mev being the most common energy. The distribution curve obtained by Anderson and Neddermeyer is shown in Fig. 230.¹

Interesting statistical data on the occurrence of *showers of various sizes* were also obtained by Anderson and Neddermeyer.² They used a counter-controlled cloud chamber divided midway by a horizontal lead plate 3.5 mm. thick, placed in a horizontal magnetic field of 7,900 oersteds and viewed horizontally. At Pasadena, near

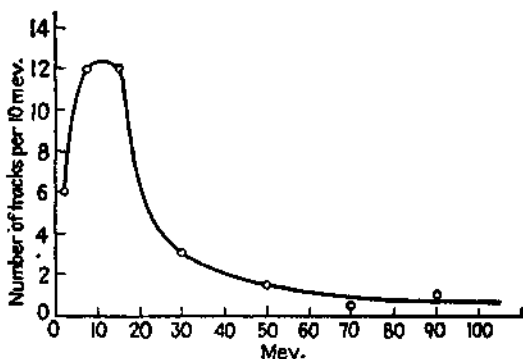


FIG. 230.—Energy distribution of positive and negative electrons occurring as showers in a cloud chamber. (Anderson and Neddermeyer.)

sea level, out of 2,684 photographs of cosmic-ray tracks, 383 or 14 percent showed showers of two or more particles. On Pikes Peak (elevation, 4,300 meters above sea level), out of 1,775 photographs, 752 or 42 percent showed showers. Two-particle showers were the commonest; but in both places the next commonest size was a shower of 6 to 10 particles. One photograph obtained on Pikes Peak showed more than 300 tracks of electrons and positrons, their total energy being estimated to exceed 15,000 Mev; four others showed showers of at least 100 particles. Often it was evident that the number of particles in a shower increased as the shower passed downward through the lead sheet. Three of their photographs are reproduced in Fig. 231.

The *increased frequency of showers higher elevations* is an impor-

¹ "International Conference on Physics/" p. 174, University Press (John Wilson & Son, Inc.), Cambridge, Mass., 1935.

² ANDERSON and NEDDERMEYER, *Phys. Rev.*, vol. 50, p. 263 (1936).

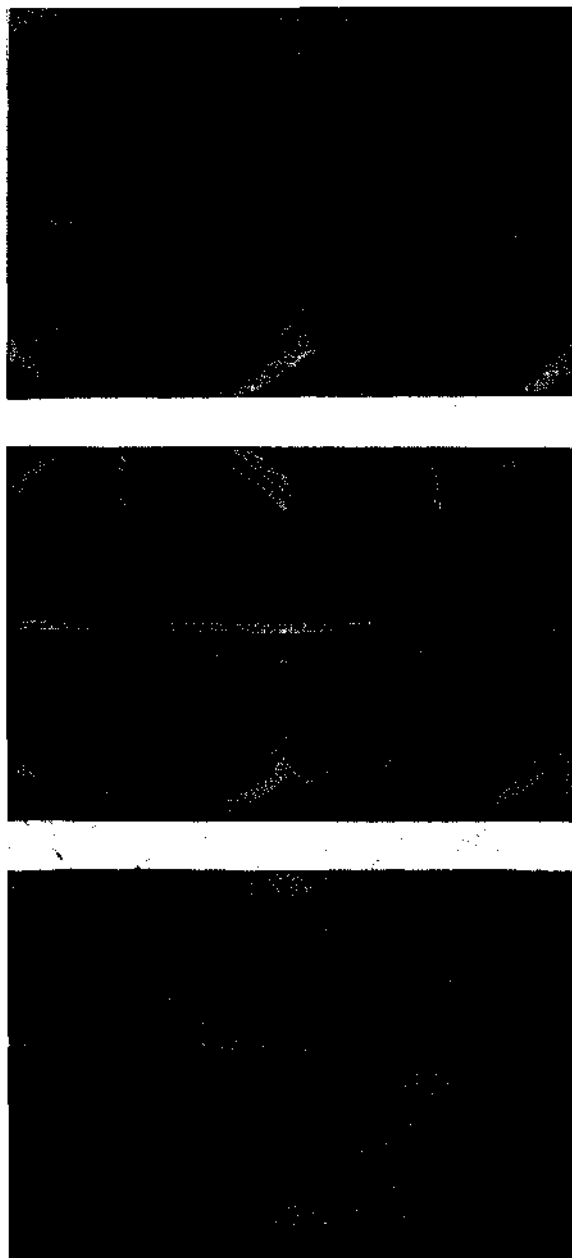


FIG. 231.—Reproduction of three stereoscopic cloud-chamber photographs showing showers, two of them exhibiting multiplication in a lead plate. [Photographs kindly lent by Professor C. D. Anderson. See Anderson and Neddermeyer, *Phys. Rev.*, vol. 50, p. 263 (1936).]

tant feature of these observations. An increase in the frequency of all tracks would be expected, since the cosmic-ray ionization itself increases with altitude; but *the showers increase notch faster than the single particles*, especially the larger ones. Frequencies of occurrence were calculated by dividing the number of photographs showing showers by the time taken to obtain them, exclusive of the 15-second intervals required after each expansion to reset the cloud chamber. The rate was thus found to be 34 photographs showing tracks per hour in Pasadena and 120 per hour on Pikes Peak. The ratio of the number of photographs per hour showing n tracks on Pikes Peak to the corresponding number at Pasadena, for various values of n , was calculated to be as follows:

n	1	2	3	4	5	6-10	11-20	21-100
Ratio	2.7	7.5	0.9	17	24	18	33	24

The increase in the frequency of large showers at the higher elevation is very striking.

The statistics of showers as observed depend greatly, however, upon the experimental arrangement. Hazen, using the apparatus described in Sec. 238(d) at an altitude of 10,000 feet, studied showers either originating or culminating within the cloud chamber.¹ He found that their frequency decreased by a factor of about 10 as the maximum number of particles in the shower increased from 5 to 12, or the total energy of the shower from 200 to 550 Mev, and that showers of more than 21 particles or showers of total energy exceeding 1,000 Mev were relatively rare.

(b) *Bursts of Ionization*.—In 1927, the German physicist Hoffmann, while studying the ionization caused by cosmic rays in an ionization chamber, noticed that occasionally a large deflection of his electroscope occurred.² He convinced himself that these deflections were not caused by some defect in the apparatus and indicated, therefore, the sudden production of a large "burst" of ionization. The phenomenon has been studied extensively, especially since 1932. Bursts are observed to vary greatly in size, their frequency decreasing with increasing size. Apparently no limit can be set as yet to the possible size of a burst. While using an ionization chamber 36 cm. in diameter filled with argon at 50 atmospheres and shielded by 12 cm. of lead, at Huancayo, Peru (altitude, 3,350 meters), Schein and Gill observed a burst of some 10^9 ion pairs, which they ascribed to about 10,000

¹ HAZEN, *Phys. Rev.*, vol. 65, p. 67 (1944).

² Cf. "International Conference on Physics," p. 226.

cosmic-ray particles having a total energy of some 10^{12} ev; and they say that even larger bursts occurred.¹

No further details concerning observations on bursts will be given here, however, chiefly because it seems to be well established that most bursts are caused by the same showers of particles that are observed in cloud chambers. This conclusion has been supported by observations in which a cloud chamber and an ionization chamber were placed one above the other and coincidences between showers in the former and bursts in the latter were studied. Burst observations thus constitute an alternative method of studying showers.² A few bursts, however, are the result of nuclear disintegrations caused by the cosmic rays.

(c) *Observations on the Production of Showers.*—Still a third method

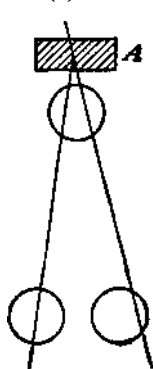


FIG. 232.—Sketch of three noncollinear counters for detection of showers generated in A.

of observing showers of charged particles is by means of three or more coincidence counters *not* placed in a row (if placed in line, they count both showers and single particles).³ In 1933, Rossi used the arrangement shown in Fig. 232. A coincident discharge of the three counters can only be caused by a shower which contains at least two particles, including the incident one (or, very rarely, by two independent particles passing through the apparatus almost simultaneously).

With the apparatus unshielded, Rossi observed 6.75 triple coincidences per hour. The number of chance coincidences, calculated from the rates of the counters when counting separately, he estimated at 3 per hour. A lead plate 1 cm. thick, placed above the upper counter at A_f increased the coincidences to 14.1 per hour. This increase is ascribed to the production of showers in the lead plate.

The number of showers as a function of the thickness of the shower-producing matter was studied, using a similar arrangement, by varying the thickness of the block A. When the counting rate with the block A removed was subtracted as a correction, the number of the remaining coincidences, plotted against the thickness of the lead block A as abscissa, gave a curve like the left-hand third of the lower one shown in Fig. 233. Many similar observations have been made subsequently by others, the curve being thereby extended to greater thicknesses.

¹SCHEIN and GILL, *Rev. Modern Phys.*, vol. 11, p. 267 (1939).

²For a summary of burst observations see Froman and Stearns, *Rev. Modern Phys.*, vol. 10, p. 133 (1938).

³Rossi, *Zeits. f. Physik*, vol. 82, p. 151 (1933).

The figure shows the data obtained by several observers. A considerable discrepancy is to be noted for the greater thicknesses.

The maximum in the number of showers under lead occurs at a thickness of about 1.3 cm. or at 14 grams per cm.²; this maximum is further discussed in Sec. 238(e). Some observers, but not all, report a second slight maximum at 15 to 20 cm.¹ Similar results have been obtained using iron or aluminum. Such curves, or similar ones in which the integrated total number of showers is plotted, are called (Rossi) "transition" curves.

It seems clear that the showers observed in this way have their origin, at least partly, in the lead. That their number should *increase*

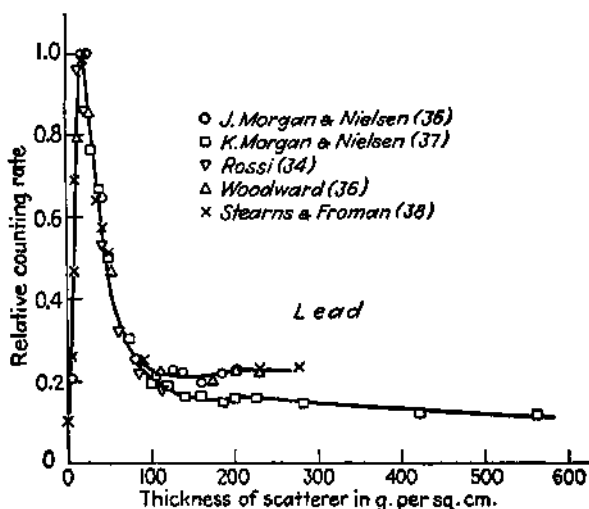


FIG. 233.—Relative numbers of showers observed under various thicknesses of lead.

at first with increasing thickness of lead is understandable. The subsequent decrease in their number, however, as the thickness is increased further, can only be due to absorption of the primary shower-producing agency. For an increase in the thickness is equivalent to the addition of a fresh layer of lead on top. This added layer will (1) itself give rise to shower particles, some of which may penetrate the lower layers of lead, and (2) absorb some of the shower-producing particles, so that fewer showers will be generated in the underlying lead. If the lead already in place is thick enough so that few of the shower particles produced in the added layer can get through it, the absorption

¹Cf. BROUSSARD and GRAVES, *Phys. Rev.*, vol. 00, p. 413 (1941); NBRESON, *Phys. Rev.*, vol. 61, p. 111 (1942).

effect will predominate; the number of showers observed below the lead will then be decreased when more lead is added.

The initial rapid decrease of the showers up to 10 cm. of lead, followed by a very slow continued decrease, must mean, then, that the shower-producing agency, whether it be particles or photons, consists roughly of two parts, a more highly absorbable or "soft" component and a very penetrating or "hard" component. The rate of the continued slow decrease, corresponding to

$$\mu/\rho = 0.6 \times 10^{-3},^*$$

agrees with the value $\mu/p = 0.55 \times 10^{-3}$ found by Rossi for the absorption in lead of cosmic-ray particles filtered through 25 cm. of lead, which would certainly represent the penetrating component. The more absorbable rays which, in the absorption measurements, were removed by 10 to 20 cm. of lead would then represent the soft component.

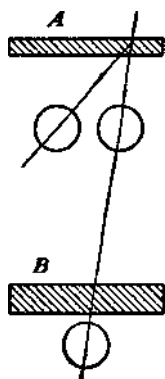


FIG. 234.—Arrangement for measuring the absorptivity of shower particles.

In order to explain the occurrence of a maximum in the transition curve at less than 2 cm. of lead, we must suppose that the shower particles themselves are easily absorbed. This hypothesis can be tested directly by measuring the *penetrating power of the shower particles*. This is easily done by placing an absorbing screen above the lowest of the three counters (e.g., at B in Fig. 234). It is found that most of the shower particles are absorbed by a few centimeters of lead.

In view of these results, it is a reasonable hypothesis, now generally held, that the soft component of the primary cosmic rays consists of shower particles that have been produced, directly or indirectly, by the penetrating component in the overlying atmosphere or in other material bodies. The nature of the hard component then remains to be determined. The mode of initiation and growth of showers in matter thus becomes a theoretical problem of great importance. We shall devote the next section to a discussion of this problem.

238. Theory of the Shower Phenomenon, (a) *The Shower Process.* Until 1934, the origin of showers of cosmic-ray particles was generally held to be a mystery. The materials for a theoretical understanding of them had already been accumulated, however, and during the next 3 years an apparently satisfactory theory for the common type of shower was developed.

* MORGAN and NIELSEN, *Phys. Rev.*, vol. 51, p. 689 (1937).

Several physicists, noting that shower tracks usually do not all diverge from a common point, suggested that they may originate in cascade fashion.¹ In 1935, Auger pointed to the production of electron-positron pairs by photons, which was described in Sec. 212, as a possible step in the process. These ideas led to the following picture of the life history of a shower. A high-energy photon becomes converted, in the field of a nucleus, into an electron and a positron [*cf.* Fig. 235(a)]. These particles constitute a shower of two rays; or, if the photon itself had been generated by a high-speed particle, we have at this stage a shower of three rays. The particles, accelerated in the fields of other nuclei, then radiate part of their energy in the form of fresh photons, this process constituting the extension to high energies of the process by which the continuous X-ray spectrum is produced [Fig. 235(b)]. These photons in turn disappear in giving rise to additional pairs of electrons and positrons; these produce fresh photons; and so on. At such high energies, it can be shown that the particles and photons should usually move almost in the same direction as the ray by which they are produced, so that the whole group of particles and photons will move on in close array.

The number of photons and particles in the shower may continue to increase for some distance. Eventually, however, the energy of a particle becomes so low that it loses more energy in the production of ions than in the production of photons; its remaining energy is then rapidly frittered away and the particle comes to rest. Similarly, the photons come finally to lose energy principally by the ejection of Compton electrons from the atoms, and soon after that they become too weak to produce pairs. Because of these factors, the number of rays (particles and photons) in a shower increases to a maximum and then decreases again. In the end the positrons that have been produced all combine with atomic electrons, and the total number of charged particles in the world is thereby restored to its initial value. The entire history of the shower occupies, of course, only a minute fraction of a second, since the particles are moving almost with the speed of light.

To develop this hypothesis concerning the growth and decay of showers in quantitative form, it is necessary to know the probabilities

¹ *Cf.* MONTGOMERY, *Phys. Rev.*, vol. 45, p. 62 (1934.)

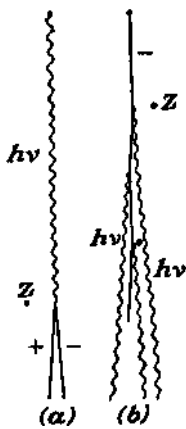


FIG. 235.—
Diagrams illustrating the processes by which showers are built up.

for the various elementary processes that are involved. We shall consider these in turn.

(b) *The Laws of Pair Production by Photons.*—The theory of the conversion of photons into pairs is given in Heitler's book on radiation.¹ The probability for the occurrence of such a conversion can conveniently be stated in terms of a cross section σ_p for pair production. If a uniformly distributed beam of photons, each of energy $h\nu$, passes through matter containing n nuclei per cm^3 , as many photons will be converted into pairs per centimeter of matter traversed as cross an area $n\sigma_p$ perpendicular to the beam. Or, the mean distance L that

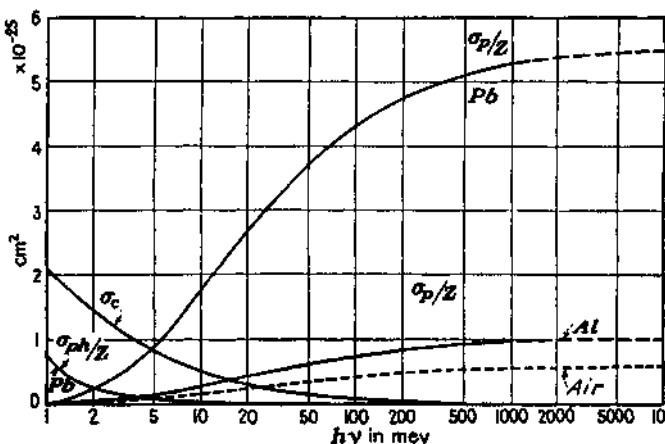


FIG. 236.—Approximate plot of the theoretical cross sections for the action of high-energy photons upon atoms of lead, aluminum, or air: σ_p/Z is the cross section σ_p for absorption of the photon with production of an electron-positron pair divided by the atomic number Z of the atom ($Z = 7.26$ for air); σ_p is the mean cross section for the Compton ejection of each electron from any atom (the cross section for the entire atom being $Z\sigma_p$); σ_{ph}/Z is the cross section for photoelectric absorption of the photon divided by Z (shown for lead only, inappreciable in aluminum or air for $h\nu > 1$ Mev). The photon energy $h\nu$ is plotted logarithmically. (From data in Heitler, "Quantum Theory of Radiation" 1936.)

a photon goes before being converted into a pair is $L = 1/(n\sigma_p)$. Theory indicates that σ_p should vary rapidly with the atomic number Z of the nuclei, in fact almost as Z^2 , especially below $h\nu = 25$ Mev. The theoretical values of σ_p for three materials, divided by Z , are shown in Fig. 236. For comparison, the cross section per electron for Compton scattering, σ_c , is also shown ($nZ\sigma_c$ being, therefore, the number of Compton electrons produced per photon per centimeter of path), and also the cross section σ_{ph} for photoelectric absorption of a photon in lead, in which photoelectric absorption is still appreciable up to 5

¹ HEITLER, W., "Quantum Theory of Radiation," 1936; 2d ed., 1945.

Mev. Some values of L , the mean distance for pair production, in centimeters are as follows:

Energy $h\nu$, Mev	In standard air	In Al	In Pb
25	9.8 X 10*	27.4	1.25
100	5.9 X 10*	17.0	.86
1,000	4.5 X 10*	13.2	.70

When a photon turns into a pair, part of its energy becomes the proper energy of the electron and the positron that are created; this part amounts to $2mc^2$ or 1.02 Mev. The remainder takes the form of kinetic energy. The latter may be divided in any ratio between the electron and the positron; and the probability for either particle to receive a given fraction of the available kinetic energy is, very roughly, the same for all fractions.

(c) *The Loss of Energy by Charged Particles.*—A charged particle passing through matter may lose energy by any one of the following processes (at least):

- (a) ionization and excitation of atoms;
- (β) emission of radiation due to acceleration in the atomic fields;
- (7) collision with nuclei resulting in nuclear transmutations.

In the case of particles of very high energy, (7) is known to occur but is rare. The principal causes of energy loss are therefore (α) and (β).

The loss due to ionization and excitation can be calculated from the observed ionization, provided we know the mean loss of energy per ion produced. Experiments with β-rays indicate that in air the mean loss is about 32 electron-volts per ion pair, not varying much with the speed of the electron. The observed minimum value of the "probable" ionization, 50 ion pairs per cm., thus implies an energy loss of at least $50 \times 32 = 1,600$ ev per cm. of path in standard air.

Theoretical formulas for the energy loss have been developed and appear to be in agreement with such experimental data as exist. Some curves plotted from the theoretical formulas are shown in Fig. 237, as is explained further below, or under the figure. The theoretical minimum loss due to ionization and excitation by electrons is about 11 Mev per cm. in lead, or 2,100 ev per cm. in standard air, at an electron energy of about 10^6 ev; of the minimum loss in air, about 1,680 ev per cm. corresponds to probable ionization. The curves for heavier singly charged particles, such as protons, are nearly the same but with all energies increased in the ratio of the mass. Thus the minimum loss for a proton should occur at an energy of about **1,800 Mev.**

The energy lost in radiation, owing to acceleration of the charged particle in the fields of the nuclei (partly screened by the atomic electrons), has been studied in a number of theoretical papers, cul-

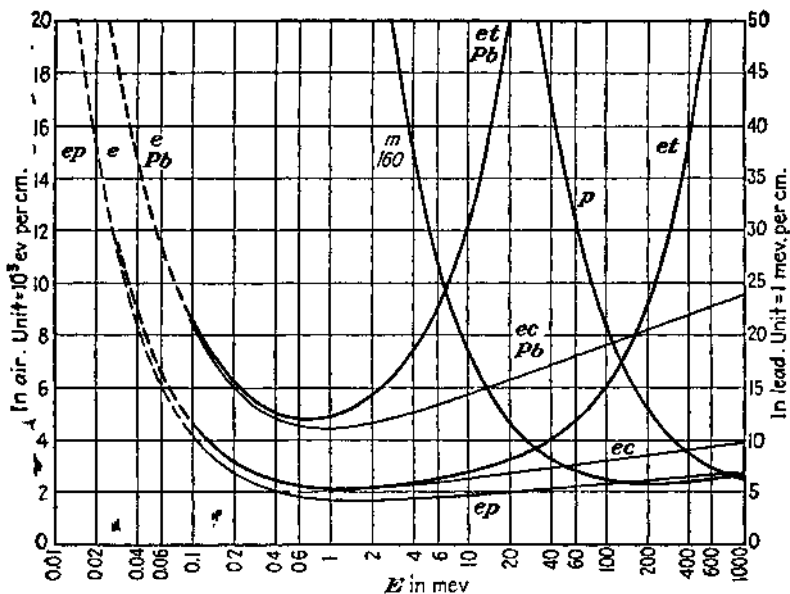


FIG. 237.—Some curves showing the approximate energy loss by charged particles passing through matter, as calculated from wave mechanics. Abscissa: kinetic energy E of the particle on a logarithmic scale. Ordinate: expectation ($-dE/dx$) of loss of energy by a particle per cm. (TV particles all of energy E lose $N(-dE/dx)dx$ units of energy in going a distance dx). Curve $ePb - ecPb$: loss due to ionization and excitation of atoms by an electron in lead. $ePb - etPb$: total loss when the direct emission of radiation (due to acceleration of the incident electron) is included, the difference in ordinates between this curve and the preceding representing the loss due to radiation alone. Curve $ep - ep$: "probable" loss in standard air by an electron due to ionization and excitation, omitting single collisions in which more than 10^4 ev is lost, $e - ec$, the same loss when all collisions are included, the fastest electron that leaves the scene of an ionizing collision being regarded as the incident one. $e - et$, loss by the same electron when the direct emission of radiation is included. Curves $m/160$ and p : total loss in standard air by a meson 160 times as heavy as an electron, and by a proton, respectively, the radiation loss being negligible. Broken parts of curves are less certain.*

* For mesons and protons, and for the "probable" loss by an electron, Bloch's formula [*Zeits. f. Physik*, vol. 81, p. 363 (1933)] was used, as modified by Babha for particles of spin $\frac{1}{2}$, [*Roy. Soc. Proc.*, vol. 164, p. 257 (1938)], viz.:

$$(-dE/dx) = (2\pi e^4 N Z / \mu \beta^2) \{ \log [2\mu \beta^2 W / (1 - \beta^2) I^2 Z^2] - 2\beta^2 \},$$

where e = electronic charge, N = number of nuclei per unit volume and Z is their atomic number ($Z = 7.26$ for air), $\mu = mc^2$ = rest energy of an electron, β = ratio of velocity of the incident particle to that of light, $I = 13$ ev, W = maximum allowed loss in a collision [$W = 10^4$ ev for the "probable" loss, otherwise $W = 2\mu(2M + E)E / [(M + \mu)^2 + 2\mu E]$ where M is the rest energy and E the

minating in the treatment given by Bethe and Heitler in 1934.¹ The radiant energy emitted per centimeter by a particle of mass M carrying a charge ze (e = electronic charge), moving with kinetic energy E through material composed of atoms of atomic number Z , is found to be represented by an expression of the type

$$\frac{z^4 Z^2}{M^2} E f\left(\frac{E}{M}\right), \quad (349)$$

where f denotes a function that is the same for all particles and materials. At high energies, f tends to become constant; the rate of emission of radiant energy is then nearly proportional to the energy of the particle. The loss of energy by radiation decreases rapidly with increase in the mass M of the particle; hence, it is usually important only in the case of electrons. Because of the factor Z^2 , the rate of loss by radiation is much larger in matter composed of heavy atoms, such as lead, than in matter composed of light atoms.

For our purpose it will be convenient to express the rate of loss of energy in terms of an equivalent mean free path for radiation ξ . This is a distance such that, if the loss by radiation were to remain constant, the particle would lose all of its kinetic energy in going a distance ξ ; or, in going a distance dx at initial energy E , the expectation of loss of energy by emission of radiation is $E \, dx/\xi$. According to Eq. (349), we can write

$$\xi = \frac{M^2}{z^4 Z^2} F\left(\frac{E}{M}\right), \quad (350)$$

where $F = 1/f$.

An individual particle will not radiate all of its kinetic energy in the distance ξ , of course, because, as the energy decreases, the rate of emission decreases. Furthermore, the radiation is actually emitted in quanta or photons; the statements just made have reference to the

kinetic energy of the incident particle}. For electron collisions with all losses included, Möller's expression [*Ann. d. Physik*, vol. 14, p. 531 (1932)] was integrated up to $W = E/2$, giving

$$(-dE/dx) = (2\pi e^4 N Z / \mu \beta^2) \{ \log [\mu \beta^2 E / (1 - \beta^2) I^2 Z^2] + 1 - \beta^2 - [2 - (1 - 1/\gamma)^2] \log 2 + (1 - 1/\gamma)^2/8 \},$$

$\gamma = (1 - \beta^2)^{-1/2}$; below $E = 0.1$ Mev, however, the curve in Fig. 237 was merely continued so as to join the probable-loss curve at 10^4 ev per cm.

¹ BETHE and HEITLER, *Roy. Soc. Proc.*, vol. 146, p. 83 (1934). Cf. also HEITLER, *LOC.cit.*

mean expectation of the emission of photons, not to a continuous emission of radiation as in classical theory. A photon may contain any fraction, from 0 to almost 1, of the kinetic energy possessed by the particle before emission of the photon. According to the theory, slightly more energy is radiated, on the average, in the form of photons containing less than half of the kinetic energy than in photons each containing more than half. If we *count* the photons emitted by a group of particles, however, we shall find more of the small ones than of the large ones; the mean number of photons per unit of energy varies a little faster than the reciprocal of $h\nu$.

In Fig. 237, the energy lost per cm. of path, according to the theoretical formulas, is plotted against the logarithm of the energy E of the particle. If y is the ordinate of the curve, the expectation of loss of energy by a particle moving with energy E in going a distance dx is ydx ; or N such particles lose as a group $Nydx$ units of energy. As explained under the figure, some of the curves refer to the total loss including the loss due to radiation, others to the loss due to ionization and excitation of atoms alone; and, for electrons in air, a curve is shown for the loss of ionization and excitation when those collisions are omitted in which the electron loses more than 10^4 ev in a single collision, the loss so defined corresponding to the "probable" ionization as defined in Sec. 236(6). Curves are shown for electrons in lead, and, in standard air, for electrons, for protons, and for mesons (Sees. 239 and 240) 160 times as heavy as an electron, the radiative energy loss for the latter two kinds of particle being negligible over the range of energies considered.

The curves illustrate strikingly the fact that the radiation loss predominates over all other losses at high energies of the moving particle (for electrons, above 100 Mev in air or 10 Mev in lead). It may be added that, at energies exceeding the limits for the curves shown, an electron whose kinetic energy is $E \times 10^6$ ev radiates on the average about $30 E$ ev per cm. in standard air, or $1.95 E \times 10^6$ ev per cm. in lead.

By integrating the area under the curve that gives the total energy loss, we can find the range of an electron in a given material as a function of its initial kinetic energy, or the total distance that it will travel before coming to rest. Heitler¹ gives the following values of the range in centimeters for an electron moving through standard air or through lead, the initial kinetic energy E being expressed in terms of mc^2 , the electronic rest energy, as a unit ($mc^2 = 0.511$ Mev):

¹ HEITLER, *loc. cit.*

E	0.1	1	10	100	1,000 mc ²
Air.....	4.4	160	20×10^3	17×10^3	63×10^3
Pb.....	9×10^{-4}	3.7×10^{-3}	0.30	1.25	2.5

(d) *The Growth and Decay of Showers.*—The application of the theoretical results just described to the growth and decay of showers is complicated by the great *straggling* of the elementary processes. A photon may travel a relatively short distance or a long distance before turning into a pair; and charged particles may generate photons early or late in their career. Furthermore, the energy of individual particles or photons may vary over a wide range. For this reason, to follow the course of a shower mathematically presents a difficult problem which has been solved only roughly.¹ We can cite only the results of the approximate analysis. Before doing this, however, it may be instructive to describe a simplified model of the shower process which succeeds in predicting correctly many of its principal features.

For this purpose, let us assemble rough values of a few of the quantities that are important in shower production, as follows:

	Standard air	Al	Pb
L (for $h\nu = 1,000$ Mev).....	4.5×10^4	13	0.7 cm.
ξ (for radiation).....	3.3×10^4	10	0.5 cm.
E_i (ionization = radiation).....	120	65	9 Mev
R_i	280×10^3	8	0.5 cm.
E_{cp}	25	16	5 Mev

Here L is the mean free path for pair production by photons, and ξ is the corresponding quantity for the emission of radiation by an electron or positron, as just defined. E_i is the critical kinetic energy of an electron or positron at which its mean loss of energy due to inelastic collisions (ionization and excitation) equals that due to the emission of radiation; and R_i is the approximate range of such a particle in the material in question, or the distance it will go before stopping if it starts with energy E_i . E_{cp} is the photon energy at which the probability of loss of energy from the photon by production of a Compton electron equals the probability of loss due to pair production.

¹ CARLSON and OPPENHEIMER, *Phys. Rev.*, vol. 51, p. 220 (1937); BHABHA and HEITLER, *Roy. Soc. Proc.* vol. 159, p. 1 (1937); SNYDER, *Phys. Rev.*, vol. 53, p. 960 (1938); BHABHA and CHAKRABARTY, *Indian Acad. Sci., Proc.* vol. A 15, p. 464 (1942); *Roy. Soc. Proc.* vol. A181, p. 267 (1943).

To develop a simplified theory, let us now assume, arbitrarily, that each charged particle produced as one member of a pair goes a distance ξ and then emits all of its kinetic energy in 1 or more photons. Since the average frequency at which radiation is actually emitted, as stated above, is about $\nu/2$, where $h\nu = E$, the kinetic energy of the particle, we might expect about 2 photons from each particle. The number of photons increases rapidly, however, as the frequency is decreased; a given amount of energy at a frequency $\nu/4$, *e.g.*, would make 4 photons, as against $4/3$ at frequency $3\nu/4$. A better estimate of the number, therefore, should be 3 photons per particle. Each photon, we will assume, goes a distance exactly equal to L and then turns into two particles, whose energies we suppose to be equal, as they actually are on the average. The net result, on this simplified picture, will then be that in a distance $\xi + L$ one particle is replaced $3 \times 2 = 6$ particles. This is equivalent to an average rate of multiplication in proportion to $e^{\gamma x}$, where x is the distance covered from the beginning of the shower and $e^{\gamma(\xi+L)} = 6$; since, as appears from the data given above, $L = 1.4 \xi$ roughly, this makes $e^{2.4\gamma\xi} = G$ and

$$\gamma = \frac{\log 6}{2.4\xi} = \frac{1}{1.34\xi}.$$

In an actual shower, of course, photons and particles will coexist. But photons are produced more rapidly than particles, and small ; photons are produced especially rapidly; hence, we might expect to find about twice as many photons as particles in a shower.

To allow roughly for the effects of ionization, an attractive hypothesis is the simple assumption that the shower stops when the kinetic energy of each particle equals E_i , the value at which losses by ionization become equal to losses by radiation. Assuming that particles and photons have equal energies, we should thus obtain for the maximum number of particles in a shower $(1/3)(E_0/E_i)$, E_0 being the initial total energy. The distance covered up to this point would be x_i , where

$$e^{\gamma x_i} = e^{\gamma(1.34\xi)} = \frac{E_0}{E_i},$$

or $x_i = 1.34 \xi \log (E_0/E_i)$, E_0/E_i representing the total number of particles and photons. Actually, however, ionization losses interfere continually with the growth of the shower from the start, so that its maximum size should be reached somewhat before it has traveled a distance x_i . A further feature tending to decrease the number of particles is that, in the production of each pair, energy equal to 2rac^2

or about 1 Mev is lost in the rest energy of the pair particles. On the other hand, because of straggling (in time) in the elementary processes, many particles will travel beyond the point x_i . We may try to make a first correction for straggling by assuming that at the point X_i only half of the particles have sunk below E_i in energy and have been stopped. We are thus led to put down as rough estimates of the maximum number N_m of charged particles in the shower and of the distance x_m from the starting-point at which the maximum number occurs the simple expressions

$$N_m = \frac{1}{6} \frac{E_0}{E_i}, \quad x_m = \xi \log \frac{E_0}{E_i}, \quad (351a,b)$$

The corresponding values derived from the mathematical analysis, according to Serber,¹ are well represented, for $E_0/E_i > 20$, by the equations

$$N_m = 0.4 \frac{E_0}{E_i} \left[1 + 1.6 \left(\log \frac{E_0}{E_i} - 1 \right) \right]^{-1/2}, \quad (352)$$

$$x_m = \xi \left[\log \frac{E_0}{E_i} - \frac{0.8 \log \frac{E_0}{E_i} + 0.5}{\log \frac{E_0}{E_i} - 1} \right]. \quad (353)$$

A few values of N_m and x_m calculated from these equations, for selected values of the ratio E_0/E_i , are:

$E_0/E_i, \dots$	25	50	100	250	500	1,000
$\log (E_0/E_i) \dots$	3.22	3.91	4.61	5.52	6.21	6.91
$N_m, \dots$	4.7	8.4	15	35	65	124
$x_m/\xi \dots$	1.8	2.7	3.4	4.4	5.2	5.9

These values are of the same order of magnitude as the rough estimates given by the easily remembered Eqs. (351a,6).

Additional features more or less predicted by our simplified theory and confirmed by the more exact analysis are the following:

1. The general course of the shower is the same whether it is started by a particle or by a photon.

2. Showers occurring in different materials differ chiefly in the spatial scale of the phenomenon. A shower attains its maximum number of particles, in lead, after traveling from 1 to 3 cm., depending on the energy of the initiating ray, in iron after traveling four times as great a distance, and in standard air after $\frac{2}{3}$ to 2 kilometers. The

¹ SERBER, *Phys. Rev.*, vol. 54, p. 317 (1938).

whole range of the shower will be several times as great as the distance to the maximum number of particles.

3. The number of photons in a shower should be nearly double the number of particles.

4. The maximum number of particles should increase linearly with E_0 (actually a little more slowly); and it may attain any magnitude if the initial energy is sufficiently large.

5. The density of the shower at a given level should be greatest in the center and should decrease rapidly toward the sides. This feature results from the crisscrossing of the slightly diverging paths, which produces a distribution somewhat analogous to that of gaseous molecules diffusing from a common initial position.

One other interesting feature, predicted by the more exact theory and caused in large part by the relatively rapid production of small photons, is:

6. The particles composing the shower soon come to be distributed in energy roughly in proportion to $1/E^2$.

(e) *Experimental Test of Shower Theory.*—The predictions of the theory of cascade showers are probably in as good agreement with observation as can be expected in view of the many approximations that are necessary in order to make the calculations tractable.

The cascade process in lead can be studied by introducing into a cloud chamber parallel sheets of lead separated by a few centimeters. The gas between the plates has so little effect upon the shower particles that the cascade process proceeds as it would in a solid mass of lead; the only difference is that the shower is drawn out lengthwise and its lateral spread is increased, while at intervals the shower particles are revealed to view in gas-filled gaps.

Observations of the sort just described were made by Hazen,¹ using eight plates, and by Nassar and Hazen,² using 4 lead plates 0.7 cm. thick separated by spaces 5.3 cm. wide. The cloud chamber used by Nassar and Hazen was 30 cm. in diameter and 45 cm. deep, and was placed in a magnetic field of 1,100 gauss in order to measure the energies of the particles. The chamber was controlled by a three-counter telescope above it (Sec. 235). From the observations it was possible to test three features of the shower process, as follows:

1. The number of charged particles present at various distances along the path of the shower was studied in 17 showers, each of about 230 Mev total energy, as estimated from the total energy of all particles. Fair agreement was found with the theoretical curve.

¹ HAZEN, *Phys. Rev.*, vol. 66, p. 254 (1944).

² NASSAR and HAZEN, *Phys. Rev.*, vol. 69, p. 298 (1946).

2. The ratio of the maximum number of particles N_m to the total number produced in the shower, and also the total number of particles in the shower, as observed, somewhat exceeded the theoretical predictions.

3. The distribution of energy among the particles when their number is at a maximum was studied in 17 showers containing 6 to 20 particles at maximum and representing total energies of 200 to 1,000 Mev. Good agreement with the theory was found when particles of less energy than 5 Mev were excluded, for which the approximations made in the theoretical calculations are particularly poor.

Large fluctuations occur in all features from one shower to the next; and the error in even the latest theoretical calculations was estimated at 20 to 30 percent. It was concluded, therefore, that the basic assumptions of the cascade theory are probably correct.

(/) *Final Remarks Concerning Showers.*—It may be mentioned that the maximum of the air-lead transition curve that was described in Sec. 237(c) occurs in the lead about 0.3 cm. later for a photon-induced shower than for one induced by an electron. Its position varies greatly, also, with the size of the shower, from about 0.9 cm. for small showers to 1.7 cm. for large ones.¹ Electron-induced showers can be distinguished through the presence at their head of a single electron track, whereas a photon-induced shower begins necessarily with two tracks.

Occasionally a charged particle appears to eject an atomic electron from an atom with very high energy; thus one particle may be seen to enter a plate of material and two to leave it. In such cases it may be impossible to tell whether the emergent pair of particles arose in the manner described or represents a true shower of two particles, the incident particle having stopped in the plate after emitting a high-energy photon.

Our discussion has been limited to showers of the ordinary type, in which all tracks are "well collimated" or not too far from parallel (an average semiangle of divergence of 10 to 15°). Occasionally showers are observed in the cloud chamber in which the tracks do not seem to be collimated at all, and are unusually heavy. Such showers will be discussed in the next section under the heading "Stars,"

239. Neutrons, Protons, and Stars.—(a) Free *neutrons* have been detected in the atmosphere by various investigators.² They may be detected with a counter filled with BF₃ (Sec. 216); the walls of the

¹ NERESON, *Phys. Rev.*, vol. 61, p. 111 (1942).

² Cf. KORFF, *Rev. Modern Phys.*, vol. 11, p. 211 (1939); BETHE, KORFF, and PLACZEK, *Phys. Rev.*, vol. 57, p. 573 (1940).

counter may be made rather thick in order to exclude all slow electrons. In later investigations a sleeve of boron carbide was used, which could be slipped over the counter, in order to absorb neutrons entering from the outside; the difference in the counting rates with and without the sleeve is taken as a measure of the neutron intensity, the residual counts due to other causes being thus eliminated.

In the observations of Korff and Hamermesh¹ the counter was sent aloft carried by balloons. The sleeve was automatically slipped into place or removed at intervals of 2 min., the counts being transmitted to the observer on the ground by means of short-wave radio²; and the observations were interrupted occasionally while the temperature and pressure of the air were transmitted to the ground. The counter was 100 cm. long and 4.9 cm. in diameter, and was operated at 2,000 volts; it was insensitive to single fast electrons or mesons, but would respond to protons or the cores of dense showers. The surrounding sleeve of boron carbide was G mm. thick.

Using a tube containing boron, whose cross section for neutron absorption varies as $1/v$ up to neutron energies of hundreds of volts, the counting rate will be proportional to the density of the neutrons in the surrounding space, independently of their velocity v ; faster neutrons traverse the counter more frequently than slower ones having the same spatial density, but they have a proportionately smaller chance of being captured and detected. Relying upon this fact, Korff and Hamermesh conclude from their observations that at an altitude where the air pressure is equivalent to 2 meters of water there is present 1 neutron in about 20 cubic meters of air. The neutrons must be continually undergoing capture by the nitrogen of the air; from the known cross section for this process they estimate that, at the altitude in question, in order to maintain the observed density, about 1 neutron per second must be set free somehow in each G cubic meters of air. As the altitude decreases, both the neutron density and their rate of production decrease, being less than half the values cited at an altitude of 10,000 feet. At sea level there is present about 1 neutron per 200 cubic meters.

(6) *Protons* moving at high energy are believed to constitute a very small fraction, perhaps a few percent at high altitudes, of the penetrating rays in the atmosphere. Korff believed that he detected them by means of counters carried aloft by balloons, the counts being transmitted to ground as usual by radio.³ For this observation the

¹ KORFF and HAMERMESH, *Phys. Rev.*, vol. 69, p. 155 (1946).

² Cf. PICKERING, *Rev. Set. Instruments*, vol. 14, p. 171 (1943).

³ KORFF, *Phys. Rev.*, vol. 59, p. 949 (1941).

counters contained the usual argon-alcohol filling and were adjusted so as to be insensitive to single electrons or to γ -rays, but so as to respond to protons having an energy between 0.05 and 5 Me v. Estimates were made of the rate of production of such protons in the atmosphere by allowing for the rate at which they are slowed down and rendered inactive by the air. The rate of production as inferred in this manner was about 15 protons per second per cubic meter of air, at an altitude where the pressure is equivalent to 2 meters of water, or 4 per second per cubic meter at an altitude of 10,000 ft.

Powell, on 22,400 cloud-chamber photographs taken at an altitude of 10,000 ft., identified as protons about 4 percent of the rays that were observed to penetrate at least 1 cm. of lead.¹ He suggests that many of the protons *as observed* may be recoil protons generated in hydrogen-containing material near by, perhaps even in the body of an observer!

(c) *Stars*.²—A cloud-chamber photograph occasionally shows two or more heavy tracks, often short, diverging widely from a point lying either within the cloud chamber or outside of it. These appearances have been called "stars."

Analogous stars, much smaller in size, are found in photographic plates that have stood for some time without exposure to light and are then developed in the usual way. Stars were studied by Shapiro in a plate with emulsion 43 μ thick, left for 239 days on Mt. Evans at an altitude of 4,300 meters. The individual star tracks were 10-50/* long, consisting of developed grains spaced 3 or 4 μ apart. Identification was attempted by comparing the tracks with those made by fast protons or those made by α particles generated in the laboratory. The emulsion was not appreciably sensitive to fast electrons or to fast mesons.

The number of particles in a given star ranges with rapidly decreasing frequency from 2 to a maximum at least as high as 14. As regards their direction, the tracks have often been reported to be distributed at random, but Hazen found a preponderance of downward directions.³ As the altitude is increased, the number of particles per star increases only slightly, but the average energy of the particles increases rapidly. At 10,000 ft, individual energies of 10 or 20 Mev are common, and total star energies of the order of 50 Mev; but individual energies may

¹ POWELL, *Phys. Rev.*, vol. 69, p. 385 (1946).

² For early references and discussion, see SHAPIRO, *Phys. Rev.*, vol. 61, p. 115 (1941); Chadwick, May, Pickavance, and Powell, *Roy. Soc. Proc.*, vol. 183, p. 1 (1944).

³ HAZEN, *Phys. Rev.*, vol. 65, p. 67 (1944).

occasionally be as high as 100 Mev, and total star energies as high as 900 Mev.¹

The nature of the particles themselves is subject to some uncertainty. Most of them are believed to be protons, although some may be mesons; α particles also occur, apparently in about one-sixth of the stars studied by Shapiro.

Stars are observed to occur much less frequently than do single cosmic-ray particles. At 10,000 ft., for example, Hazen found, in 8,500 cloud-chamber photographs, 8,678 sharp tracks made by penetrating particles and only 58 stars. The frequency of stars increases very rapidly with altitude, about in proportion to the number of large showers or bursts. Shdanow obtained in photographic plates 8 stars per square meter per hour at sea level, but 400 at an altitude of 5,000 meters.²

Star particles are presumed to represent the products of nuclear disintegrations that are produced in some way by the cosmic rays. Presumably the disintegrations result also in the emission of neutrons, which of course are not observed either in cloud-chamber photographs or in photographic films. It is possible, in fact, that all the free neutrons found in the atmosphere owe their origin to nuclear disintegrations induced in some manner by the cosmic rays; and such disintegrations may also give rise, directly or indirectly, to some or all of the protons.

240. Mesons.—After the discovery of the latitude effect on cosmic-ray ionization, the belief became general for a time that the primary cosmic rays consist, at least in large part, of (negative) electrons. It was also believed that the cosmic-ray tracks seen in cloud chambers are mostly made either by these primary electrons or by secondary electrons or positrons ejected by them from atoms. By the end of 1934, however, several difficulties with this hypothesis had emerged.

Let us consider what energy an electron would have to have in order to penetrate the whole depth of the atmosphere. If we multiply an ionization loss of about 2,500 ev per cm. by the thickness of the atmosphere in terms of standard air, or 8×10^6 cm., we obtain 2×10^9 ev for the total loss. This is less than the energies with which most of the primary rays enter the atmosphere, as inferred from the latitude effect (up to 10^{10} ev or more—see Sec. 234).

So far, therefore, the assumption that the primary cosmic rays are electrons meets with no difficulty. The picture is changed completely,

² SHDANOW, *Acad. Set. U.S.S.R.*, vol. 4, p. 266 (1940).

however, when losses by radiation are included. The total loss becomes then, roughly, $2,500 + 3 \times 10^5 E$ ev cm.⁻¹, E denoting the energy of the electron; the second term represents the mean loss of energy due to radiation according to the Bethe-Heitler formula [Sec. 238(c)]. Thus in a distance dx the mean loss is

$$dE = -(a + bE) dx,$$

where $a = 2,500$, $b = 3 \times 10^{-5}$; whence, upon integrating,

$$E = \frac{a}{b} \left[\left(1 + \frac{bE_0}{a} \right) e^{-bx} - 1 \right],$$

E_0 being the initial energy in ev at $x = 0$. Thus E becomes zero, and the electron stops moving, when

$$1 + \frac{bE_0}{a} = e^{bx}, \quad x = \frac{1}{b} \log \left(1 + \frac{bE_0}{a} \right),$$

or

$$x = 3.3 \times 10^4 \log(1 + 1.2 \times 10^{-8} E_0) \text{ cm.}$$

To make $x = 8 \times 10^5$ cm., so that the particle can just reach the earth, we should have to have

$$\log(1 + 1.2 \times 10^{-8} E_0) = \frac{80}{3.3} = 24, \quad E_0 = 2.2 \times 10^{18} \text{ ev.}$$

If E_0 were only 10^{11} ev, which is already far above the upper limit of 6×10^9 at which a latitude effect is possible, the range in standard air would be only 2.4 km. Thus electrons or positrons with energy of ordinary cosmic-ray magnitude could not possibly get through the whole thickness of the atmosphere.

Similar difficulties were revealed by laboratory experiments. As we have seen in Sec. 235, Rossi detected charged particles capable of passing through a meter of lead, which is equivalent, as regards theoretical energy loss, to more than the whole atmosphere. In 1934, Auger and Ehrenfest, using four counters and a cloud chamber in a vertical row, with 50 cm. of lead between the two pairs of counters, observed single tracks in the cloud chamber which were accompanied by a simultaneous discharge of all four counters.¹ They concluded that such tracks were made by particles capable of traversing 50 cm. of lead. Highly penetrating particles were also observed by Crussard and Leprince-Ringuet.²

¹ AUGER and EHRENFEST, *Comptes Rendus*, vol. 199, p. 1609 (1934).

² CRUSSARD and LEPRINCE-RINGUET, *J. de Physique et le Radium*, vol. 8, p. 213 (1937).

The difficulty became accentuated when direct measurements were attempted of the loss of energy by an electron in passing through a layer of matter. Such measurements can easily be made in a cloud chamber placed in a magnetic field; it is only necessary to measure the difference in the curvatures of the tracks above and below the layer of matter, from which the change in energy can be calculated [with a possible error of a few percent in case the mass of the particle is uncertain—see Sec. 236(a)]. Extensive measurements of this sort were begun by Anderson in 1932, at Millikan's suggestion, and were continued by Anderson and Neddermeyer¹ and others.² In such observations it was found very difficult to secure consistent results. In some cases, energy losses were observed that agreed well with the Bethe-Heitler theory of radiation, which had been brought out in 1934. In other cases, especially for high-energy particles, the loss was certainly much less.

For a few years, it was believed by some that all of the known facts taken together pointed toward a breakdown of the wave-mechanical theory of radiation for electron energies above 10^8 electron-volts. The development of the theory of showers was probably delayed for a time by this erroneous conclusion, since it seemed to render hopeless any attempt to construct a theory on the basis of accepted principles.

As the experimental work progressed, however, it gradually became clear that not all cosmic-ray particles are of the same nature. In 1935, Auger suggested that the more penetrating rays might be protons. In 1936, Anderson and Neddermeyer reported observations on the amount of energy lost by cosmic-ray particles in passing through a lead plate 3.5 mm. thick placed across a cloud chamber. They found that the *average loss by particles forming part of showers* was in approximate agreement with the theory,³ although the individual variations were enormous (presumably due to straggling in the emission of photons). For a further test, Neddermeyer and Anderson replaced the lead plate by one of platinum 1 cm. thick, equivalent to nearly 2 cm. of lead.⁴ It was found, on the one hand, that even shower particles of energy below 500 Mev could get through such a thick plate in small numbers, but they emerged with almost a total loss of their energy. This can be understood if these particles are electrons or positrons and are occasionally able to traverse the plate only because their

¹ ANDERSON, *Phys. Rev.*, vol. 41, p. 405 (1932); and later issues of the *Phys. Rev.*

² For references see *Rev. Modern Phys.*, vol. 10, p. 174 (1938).

³ ANDERSON and NEDDERMEYER, *Phys. Rev.*, vol. 50, p. 263 (1936).

⁴ NEDDERMEYER and ANDERSON, *Phys. (IRev.)*, vol. 51, p. 884 (1937).

radiative energy losses happen to be abnormally low. The particles occurring singly, on the other hand, lost, as a rule, less than half of their energy in traversing the platinum.

The data in question and some later ones are shown in Fig. 238. The abscissa represents the energy with which a particle entered the cloud chamber, the ordinate, its loss of energy in passing through 1 cm. of platinum. As we have seen (Sec. 236), the kinetic energy of a very high-energy particle can be inferred to a good approximation

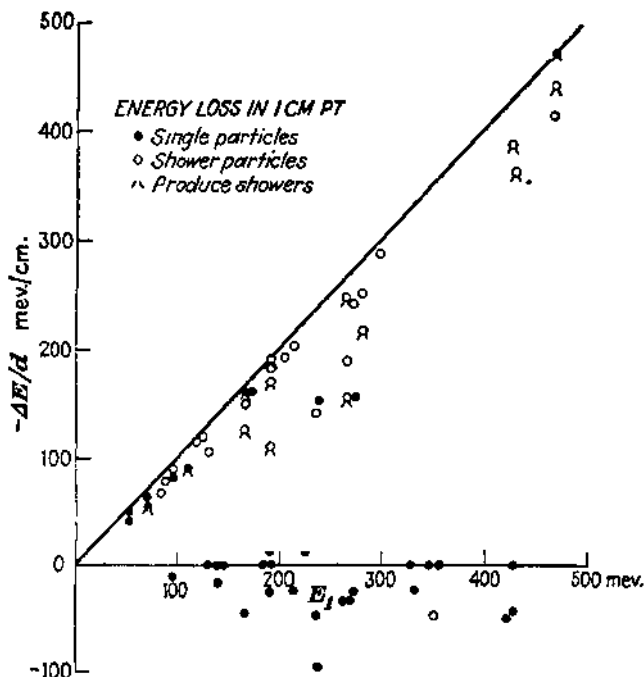


FIG. 238.—Energy loss in 1 cm. of platinum as function of incident energy. (Energy calculated on assumption of electronic mass.) (Anderson and Neddermeyer.)

from the curvature of its track without knowledge of its mass. Circles refer in the figure to particles forming part of a shower, solid dots to particles that occurred alone. All points referring to particles that traverse the platinum must necessarily lie below the line drawn in the diagram, since this line represents total loss of energy in the plate. The plot furnishes clear evidence for the conclusion that at least two different kinds of cosmic-ray particles occur in cloud chambers, characterized by wide differences in penetrating power. The occurrence of cases of an *apparent gain* of energy, however, well illustrates the difficulty of making accurate measurements of this type.

Highly penetrating particles, with energies up to 10^9 ev, were reported also by Crussard and Leprince-Ringuet.¹

All of the facts cited can be brought into harmony with the wave-mechanical theory of radiation if we assume that, unlike the shower particles, the exceptionally *penetrating particles* have a *much smaller ratio of charge to mass* than has an electron, so that the radiative energy losses are much less. As we have seen, a satisfactory theory of *showers* can be developed, using wave mechanics, on the assumption that showers consist of ordinary electrons, positrons, and photons. The soft or nonpenetrating component of the cosmic rays, defined as that part of the rays which is absorbed by about 10 cm. of lead [Sec. 237(c)], may likewise consist, at least chiefly, of electrons and positrons. The *penetrating component* of the cosmic rays is then presumably to be identified with the penetrating particles that are observed in cloud chambers. What is the nature of these particles?

Can they be protons? A proton of cosmic-ray energy would suffer no appreciable loss by radiation in passing through matter, losing energy only by ionization, and so would be much more penetrating than an electron. The following objections can be raised against the assumption that the particles in question are protons:

1. Since more than a third of them carry a negative charge, as is shown by the direction of curvature of their tracks in the cloud chamber, it would be necessary to admit the existence of negative protons, for whose existence there is no other evidence.

2. The tracks of the penetrating particles are scarcely distinguishable from electron tracks, whereas a proton of energy below 2×10^8 ev should make a track many times heavier. If the penetrating particles are protons, they ought often to be seen close to the end of their range, moving slowly enough to cause dense tracks. Careful search by several investigators has failed to reveal an appreciable number of such tracks in cloud chambers.²

3. The distribution in energy of the high-speed electrons ejected from atoms by penetrating particles does not agree well with that obtained from theory on the assumption that these particles are protons, but would be consistent with a mass several times smaller.⁸

In view of these considerations, it was concluded that the penetrating cosmic-ray particles are probably a new type of particle, with a mass intermediate between the masses of the electron and the proton.

¹ CEUSSARD and LBPBINC-RINGUET, *J. de Physique et le Radium*, vol. 8, p. 213 (1937).

² Cf. *Rev. Modern Phys.*, vol. 10, p. 180 (1938).

Ibid., vol 11, p. 203 (1939).

From Eq. (349), it is evident that the mass would only have to be 10 to 20 times the electronic mass in order to reduce the radiation loss to small proportions and so make the particles highly penetrating. Various names were proposed for these new particles; they are commonly called *mesons* or *mesotrons*.¹

241. Properties of Mesons. (a) *Mass*.—The mass of rapidly moving mesons is not easy to determine with precision. The curvature of the track of a meson in a magnetic field gives at once its momentum (Sec. 236), provided its charge is assumed to be of electronic magnitude. To find the mass, either the speed or the energy must then also be known. The speed can be inferred, usually with low accuracy, from the degree of ionization produced (Sec. 236); the energy can be inferred from the total range in the gas, provided this can be determined. An alternative method of finding the mass is to study those close collisions of mesons with atomic electrons in which the electron is ejected with high velocity.²

One of the first good observations by the ionization method was made by Street and Stevenson,³ who obtained a track in a cloud chamber under 11 cm. of lead with $H^p = 3.6 \times 10^4$ oersted-cm. and a specific ionization about six times as great as in normal thin tracks. If the particle entered from about its sign was negative. The heavy ionization would be explained' .⁵ particle had a mass 130 ± 30 times the electronic m a was not a proton, for, if it had been, its energy v , only 4.5×10^6 ev [by Eq. (346) in Sec. 236] and a π , ^energy is known to have a range of only 1 cm. in standard air, eas the track in question was clearly visible over a length of 7 cm. A thousand counter-controlled expansions were necessary to obtain this single useful track.

Values obtained for the meson mass have ranged from 30 to 1,000 but are usually between 150 and 250, more often below 200 than above it.⁴ A favorite guess has been 200. It may be that the mass is not the same for all mesons. Perhaps this question cannot be settled until mesons can be produced in the laboratory.

(b) *Mean Life of Mesons*.—Since mesons apparently occur on the earth only as cosmic-ray particles, those that enter the lower part of

¹ *Meson* is the neuter form of the Greek adjective "mesos," meaning middle; it should be pronounced "meezon." In *mesotron*, apparently, a meaningless fragment of the word, electron, has been added.

² For a discussion of methods see Hughes, *Phys. Rev.*, vol. 69, p. 371 (1946).

³ STREET and STEVENSON, *Phys. Rev.*, vol. 52, p. 1003 (1937).

⁴ Cf. BETHE, *Phys. Rev.*, vol. 70, p. 821 (1946); HUGHES, *Phys. Rev.*, vol. 71, p. 387 (1947).

the atmosphere must somehow be destroyed. Perhaps to some extent they are absorbed in some manner by atomic nuclei. At least three lines of evidence, on the other hand, indicate that mesons may decay spontaneously. To conserve electrical charge, it is assumed that in such cases a positive or negative electron replaces the meson; and, to preserve the conservation of both energy and momentum, it is assumed, in analogy with the theory of β -decay, that a neutrino is also emitted, the electron and the neutrino carrying off between them the momentum of the meson, and also its excess rest energy in addition to its kinetic energy.

1. For one thing, such a spontaneous decay of the mesons would explain the slight decrease in the cosmic-ray intensity with rise of atmospheric temperature. For, at given barometric pressure, a higher temperature means that the atmosphere extends farther upward from the earth, so that the mesons, which appear to be the chief carriers of cosmic-ray effects downward, have farther to go and more of them perish on the way.¹ The mean life of a meson, when at rest, has been estimated from this effect as 3.4×10^{-6} sec.

2. Data of another sort pointing to the existence of mesotron decay were obtained by Rossi, $\sim 10^{11}$ γ , and Hoag.² Using a vertical counter telescope, with ~ 2.7 cm^2 $\times$ between the counters in order to stop the soft component γ -ray particles, these experimenters compared the meson γ -ray at different elevations near latitude 50° , viz., at $\sim 1,616$ and stations at 1,616, 3,240, and 4,300 meters. The observed increase of meson intensity with altitude was in good agreement with the combination of absorption and disintegration of the mesons in the intervening vertical air mass. At each station except Chicago, the absorption in a thickness of graphite equal to 87 grams cm^{-2} was also determined. Carbon has the same ratio of electrons to mass as have nitrogen and oxygen; therefore, the true absorption of mesons in the two mediums should be about the same. Nevertheless, only about half as many mesons were found to be stopped by the graphite as by an equal mass of atmospheric air. The greater apparent absorption in the air may be due to spontaneous disintegration of the mesons during the longer time it takes them to traverse the air.

On this assumption, it was calculated from the data that a meson of cosmic-ray energy would travel in a vacuum an average distance of 9.5 km., occupying a time of 3.2×10^{-6} sec, before disintegrating. For a meson at rest, the mean life would then be about 2×10^{-6} sec,

¹ BLACKETT, *Phys. Rev.*, vol. 54, p. 973 (1938).

² Rossi, HILBERRY, and HOAG, *Phys. Rev.*, vol. 57, p. 461 (1940).

the longer time for the moving one being due to the relativistic effect on time scales (Sec. 61); the mean life of a meson has a fixed value in that frame of reference in which it is at rest.

It may be remarked that the data just mentioned gave for the apparent mass absorption coefficient of the mesons in air, near sea level (70 cm. Hg), $\mu/\rho = 0.8 \times 10^{-3}$, increasing to 2.4×10^{-3} at 48 cm. Hg. In carbon, the data gave $\mu/\rho = 0.4 \times 10^{-3}$ near sea level. This latter figure is in rough agreement with the value of 0.55×10^{-3} found by Rossi for the penetrating component of the cosmic rays in lead near sea level (Sec. 235). The two measurements are probably not strictly

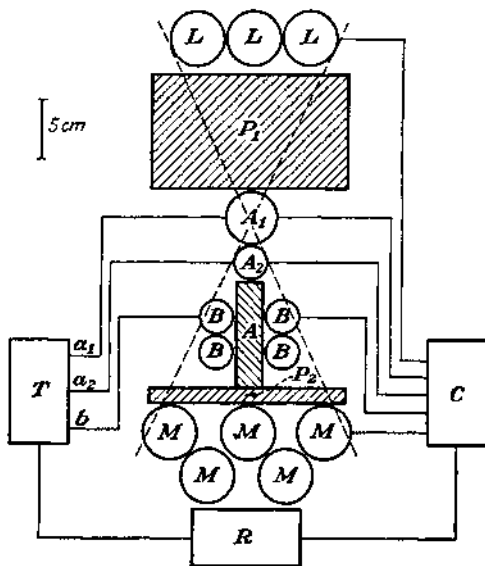


FIG. 239.—Diagram of the experimental arrangement for the study of meson decay.

comparable because of differences in the loss of intensity due to scattering in the two cases.

3. A third and much better method for the measurement of meson lifetimes is based on the idea of catching them and holding them at rest until the decay electron is emitted.¹ The experimental arrangement is interesting and is diagrammed in Fig. 239. P_1 is a lead block thick enough to transmit only mesons. Some of the mesons are stopped in the brass block A , where some or all of them decay, emitting an electron, which is very likely to pass through one of the counters B . Counters L , connected in parallel, counters A_1 and A_2 , and count-

¹ NERESON and Rossi, *Phys. Rev.*, vol. 64, p. 199 (1943); see also vol. 62, p. 417 • (1942).

ers B , connected in parallel, all feed in coincidence into circuit C . Counters M_9 connected in parallel, also feed into C but in anticoincidence with the other counters; i.e., C responds only when counters LA_1A_2B all act within a millisecond or so of each other whereas all of the counters M do *not* act. In such cases it is clear that no particle passed through M , hence it is practically certain that LA_1A_2B were not activated by a single meson passing through all of them, which would also have passed through one of the counters M . B might have been activated by a meson scattered widely in the absorber A , or by an electron ejected from A , but such occurrences will be very rare. It is thus practically certain that C responds only when a meson has stopped and decayed in A , with emission of an electron that activated one of the counters B .

To measure the time spent by the meson in A before decaying, counters A_1 , A_2 , and B are also connected to a time circuit T in such a way that, when a coincidence A_1A_2 is followed after a certain interval of time r by a pulse from B , T emits a pulse whose amplitude is proportional to r . Finally, the recorder R records with a pen the magnitude of the pulse received from T , but only in case a pulse is simultaneously received from circuit C , indicating that an event of the desired kind has occurred.

It was assumed that the meson decay followed the exponential law that is familiar in radioactive decay; thus, life periods of all possible magnitudes will be obtained from individual mesons, and their average will be the mean life, independently of the time already spent by the mesons before entering the absorber. The exponential law of decay was verified from the observations; and the mean life of a meson at rest was found from the decay curve to be $2.15 \pm 0.07 \times 10^{-6}$ sec.

If the principle that underlies the experiment just described is correct, the tracks of decay electrons ought frequently to be seen in cloud chambers. In a few cases a rather heavy track, presumably made by a slowly moving meson, has been seen to be continued with an abrupt change of direction by a thin track, presumed to be that of the decay electron.¹ But more frequently, apparently, heavy tracks have been seen to stop in the gas without any continuation. Final clarification of the situation must probably await laboratory experimentation.

(c) *Origin of Mesons*.—If mesons are subject to rapid spontaneous decay, they cannot possibly be themselves the primary cosmic-ray particles; they must be produced somehow in the upper atmosphere,

¹ Cf. *Phys. Rev.*, vol. 62, p. 552 (1942).

and for the most part, in view of their distribution in altitude, at great heights above the earth.

At this point the speculation published by Yukawa in 1935 is of interest.¹ He suggests that the attractive field which holds nucleons (protons and neutrons) together in nuclei may well involve the existence of "quanta" of some sort, in analogy with the photons that appear in the electromagnetic field in association with the forces between electrical charges. Following up this analogy and speculating as to the probable form of the correct mathematical theory, he concludes that the new quanta should carry an electrical charge, positive or negative, equal in magnitude to that on an electron, but should have a mass about 200 times as great. This speculation may well constitute a long step toward the correct theory of the mesons. If so, it may be concluded that mesons are generated in close collisions between nucleons. It is thus an attractive hypothesis that the primary cosmic-ray particles are protons and that they generate the mesons in impacts upon molecules of air. This hypothesis, however, still awaits confirmation (in 1946).

On the other hand, many observers have obtained evidence that at high altitudes mesons may be generated to a small extent in solid bodies upon which no charged particle is incident. The generating agent in such cases is presumed to be either photons or neutrons. To provide the rest energy of the meson, the photon energy would have to be of the order of 100 Mev. The effect increases rapidly with altitude.

In the investigation of Tabin² an elaborate arrangement of counters was employed in order to select cases in which only mesons were present, natural or generated, and a large cap of closely spaced counters was placed over the generating body in order to detect and eliminate all cases in which a charged particle was incident on this body from above. The body generating mesons was a block of paraffin, iron, or lead of variable thickness up to a maximum of 11.4 cm., or up to 57.2 cm. of paraffin. Observations were made at Echo Lake (altitude, 10,600 ft.) and on Mt. Evans (14,200 ft.).

With paraffin as the generating agent, saturation in the production of mesons was not reached even at the maximum thickness of 57.2 cm. With lead, on the other hand, fewer mesons were obtained from a layer 11.4 cm. thick than from a layer only a few centimeters thick. These facts are in agreement with the assumption of *photons* as the causative agent, since it should take a meter or so of paraffin to absorb

¹ YUKAWA, *Phys. Math. Soc. Japan*, Proc. vol. 17, p. 48 (1935). (In English.)

² TABIN, *Phys. Rev.*, vol. 66, p. 86 (1944).

the photons (mainly by pair production), whereas all photons would be absorbed by a few centimeters of lead, and an additional thickness of lead would then merely stop some of the mesons that had already been generated. As was also found previously by Schein, Jesse, and Wollan, the production of mesons seemed to increase about in proportion to the intensity of the soft component of the cosmic rays, which is believed to contain an appreciable number of photons of energy exceeding 100 Mev. If the mesons are generated by the impact of these photons upon the atomic nuclei, the cross section for this process, to account for the observations, would be $4 \text{ or } 5 \times 10^{-26} \text{ cm.}^2$ in paraffin or iron and $1 \times 10^{-24} \text{ cm.}^2$ in lead, and thus would be comparable with the cross-sectional area of the nucleus itself.

The study of mesons is obviously in its infancy. The student should consult the literature from 1947 on for further developments.

242. Mesons and Electrons in the Atmosphere.—Many observations of the principal cosmic-ray particles as they occur in the atmos-

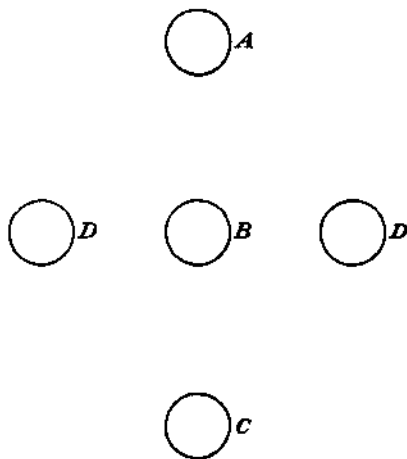


FIG. 240.—Diagram of a counter arrangement for the study of ionizing particles.

phere have been reported, but it is not easy to bring them all into harmony. The discussion will be limited here chiefly to conclusions that appear to be well established.

The experimental arrangements vary greatly. Some observers use cloud chambers, with or without a magnetic field, others use large ionization chambers for the study of bursts, and others use Geiger counters in various and often elaborate combinations. Sometimes Geiger counters are connected in "anticoincidence." For example, an arrangement such as that sketched in Fig. 240 may be used to

count only particles that do not form part of a shower. For this purpose, counters *D* are connected in parallel with each other, and this pair is then connected, by means of suitable electric circuits, in "anti-coincidence" with *ABC*, which are in coincidence with each other; i.e., it is arranged that a "count" shall be registered only when, simultaneously *A*, *B*, and *C* all act and neither of counters *D* acts. Thus a count is registered only if a particle passes through *A*, *B*, and *C* but no other particle passes simultaneously through either of the counters *D*. Another example of anticoincidence connections was described in Sec. 241(6). Alternatively, the shower-producing power of particles traversing *A* can be determined by inserting shower-causing material between *A* and *B* and connecting all counters in coincidence. Then, to be counted, a particle must pass through *ABC* and also give rise to at least two shower particles which pass one through each of counters *D*. In either case, if at least 10 cm. of lead is inserted between *B* and *C*, the count will be restricted to mesons as the incident particle.

Identification of particles within the cloud chamber is often not easy.¹ A particle that passes through several centimeters of lead without initiating a shower is almost certainly not an electron. If it is a proton, it will ionize a gas about twice as heavily as a meson unless its energy is extremely high (above 300 Mev). Because of its large mass, a proton is capable of producing in a gas many 5-rays or secondary electrons of high energy, whose tracks are visible as short sidetracks along the almost undeviated track of the proton. Tracks of slow mesons are hard to distinguish from those of electrons, except at very low speed, where the ionization is exceptionally heavy over a considerable range.

(a) *Count of the Particles.*—Excellent statistics on the number of particles occurring naturally in the atmosphere were reported by Greisen.² He used a telescope of six counters of the usual cylindrical form with their axes 7.3 cm. apart; the counters were made of brass 1 mm. thick, of internal diameter 4.24 cm., and contained collecting wires 20 cm. long. The telescope could be turned so as to count particles traveling at various inclinations to the vertical and also in various azimuths around the vertical. The following counts were made:

1. Mesons with momentum above 3.0×10^8 ev/c and energy, therefore, roughly above 350 Mev, obtained by counting coinci-

¹ Cf. HAZEN, *Phys. Rev.*, vol. 67, p. 67 (1944); POWELL, *Phys. Rev.*, vol. 69, p. 385 (1945).

² GREISEN, *Phys. Rev.*, vol. 61, p. 212 (1942).

dences between all six counters with 13 cm. of lead interposed in the form of 2-cm. blocks between the successive counters and 3 cm. above the top one.

2. Electrons with energies above about 20 Mev, and also slow mesons stopped by 13 cm. of lead, by repeating the count with all lead removed and noting the increase in the number of particles as observed.
3. Electrons having energies roughly between 4 and 20 Mev, obtained as the difference between the number of twofold coincidences between the two extreme counters, (a) when all lead was removed, and (b) when the intervening counters and their supporting wooden shelves were also removed.

Observations were made at 4 stations above 50° north magnetic latitude, viz., at Ithaca, New York (elevation, 259 meters), and in Colorado at Denver (1,616 m.), Echo Lake (3,240 m.), and Mt. Evans (4,300 m.). In all locations many obliquely moving particles were found; at 45° to the vertical, the numbers ran mostly from a third to a half of those at 90°. By integrating the curve obtained as a function of zenith angle, the total number of particles arriving at all angles was found. The results are summarized in Table I.

TABLE I.—COSMIC-RAY INTENSITIES IN THE ATMOSPHERE

Elevation	259 meters		1,616 meters		3,240 meters		4,300 meters	
	(v)*	(t)†	(v)	(t)	(v)	(t)	(v)	(t)
Fast mesons	0.44	0.68	0.56	0.86	0.81	1.21	1.01	1.62
Electrons > 20 Mev, } and slow mesons }	0.12	0.14	0.21	0.27	0.47	0.52	0.78	0.91
Electrons, 4–20 Mev..	0.11	0.23	0.20	0.33	0.50	0.74	1.05	1.47
Total particles	0.67	1.05	0.97	1.46	1.78	2.47	2.84	4.00

* Number of particles moving nearly vertically and crossing 1 cm.² of horizontal area per minute per unit of solid angle.

† Total number of particles crossing 1 cm.² of horizontal area per minute in all directions.

It will be noted from Table I that near sea level the fast mesons are twice as numerous as the electrons, whereas at 4,300 meters electrons predominate. Below the surface of the earth, the soft component of the cosmic rays becomes further reduced with increasing depth below the surface, but it is never less than 10 to 20 percent of the hard component.¹

¹ Cf. WILSON and HUGHES, *Phys. Rev.*, vol. 63, p. 161 (1943).

At still higher altitudes than those shown in the table, on the other hand, the predominance of electrons has been found to be much greater. In Fig. 241 is reproduced a curve from a paper by Bhabha and others¹ in which are plotted, in arbitrary units, as a function of the atmospheric pressure, the vertical intensity of (A) mesons at 3.3° north magnetic latitude (Bangalore, India), as observed by Bhabha, Aiya, Hoteko, and Saxena; (B) mesons at 52.5° N, as inferred from the data of

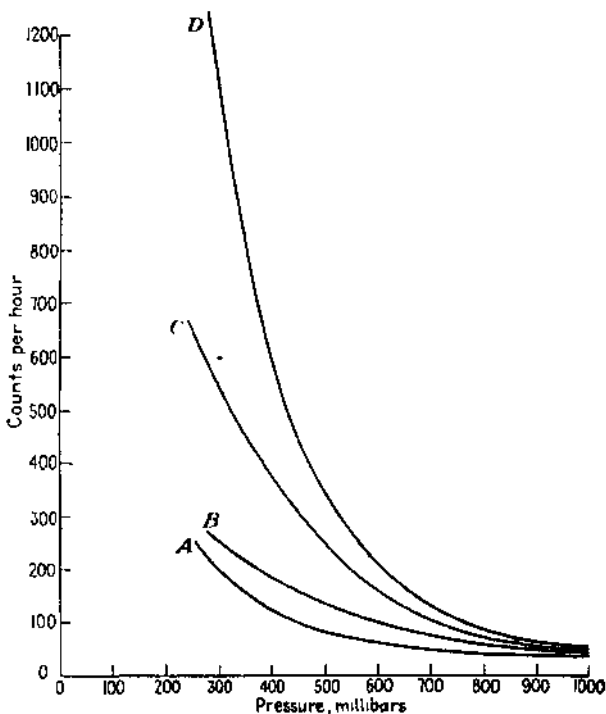


FIG. 241.—Curves showing variation with altitude of (A) mesons at 3.3° N, (B) mesons at 52.5° N, (C) mesons and electrons at 3.3° N, (D) mesons and electrons at 49° N.

Schein, Jesse, and Wolian;² (C) mesons and electrons at 3.3° N., as inferred from the data of Neher and Pichering;³ and (D) mesons and electrons at 49° N. as inferred from Pfotzer's observations.⁴ From these curves it is clear that at very high altitudes the electrons increase in number with geomagnetic latitude much more than do the mesons.

¹ BHABHA, AIYA, HOTEKO, and SAXENA, *Phys. Rev.*, vol. 68, p. 147 (1945).

² SCHEIN, JESSE, and WOLLAN, *Phys. Rev.*, vol. 47, p. 207 (1935).

³ NEHER and PICKERING, *Phys. Rev.*, vol. 61, p. 207 (1942).

⁴ PFOTZER, *Zeits. f. Physik*, vol. 102, p. 23 (1936).

Apparently there are 20 percent more positive mesons than negative. The reason for this difference is not known.

High-frequency photons should also be present in the atmosphere, in association with atmospheric showers; but these photons are hard to count. At an altitude of 10,000 ft., Hazen concluded that only 70 percent of the showers produced in lead plates were initiated by charged particles, whose tracks could be seen above the head of the shower; presumably the remaining 30 percent were initiated by photons.¹

(b) *Energy Spectra*.—The *electrons* found in the atmosphere have energies ranging with decreasing frequency from a few Mev up at least to 500 Mev. At sea level, Lombardo and Hazen find, roughly, the following relative numbers of electrons in the energy ranges stated:

Energy:	9—20—50—100—200—500
Number:	100 50 40 32 24

The mesons, on the other hand, have energies ranging from 200 to 15,000 Mev (0.2 to 15 Bev). There is a pronounced peak in the spectrum which has its maximum at about 1,000 Mev or 1 Bev, decreases to about half height at 2 Bev, and continues into a tail extending to 10 Bev at sea level,² or, with greater intensity, to 15 Bev at an altitude of 4,350 meters.³

(c) *Meson Showers*.—As was stated in Sec. 237(a), cosmic-ray showers appear ordinarily to consist entirely of electrons, together with the invisible associated photons. Many observers, however, have reported small showers composed of mesons, identified as such by their ability to penetrate 10 to 15 cm. of lead. In the cloud chamber it can also be verified that the particles composing these showers do not themselves generate showers in a lead plate. Meson showers are especially apt to be observed at high altitudes, where they appear to occur at the rate of one for every 3,000 or 4,000 photographs of single tracks. It is noted that the paths of the particles tend to diverge less widely than do those of electron showers.

Shutt gives a list of references for meson showers⁴ and reports further observations of small showers generated in heavy material above or within the cloud chamber. Most of the apparent showers

¹ HAZEN, *Phys. Rev.*, vol. 65, p. 67 (1945).

² HUGHES, *Phys. Rev.*, vol. 57, p. 592 (1940).

³ HALL, *Phys. Rev.*, vol. 66, p. 32 (1942).

⁴ SHUTT, *Phys. Rev.*, vol. 69, p. 261 (1946); see also CLAY, *Physica*, vol. 9, p. 897 (1942).

contained only 2 mesons, and of these it was estimated that all in which the divergence of the paths exceeded 6° could have represented merely the chance association of 2 independent mesons. Angles of divergence under 2° were the commonest. The mode of production of¹ such showers remains obscure.

(d) *Auger Showers*.—So far nothing has been said about the lateral extent of showers as they occur in the atmosphere. Their breadth may be considerable.¹ In 1938 Auger obtained coincidences between two or three counters placed as much as 10 meters apart. About 60 such showers seemed to arrive per hour at sea level, but 600 per hour on the Jungfrauoch (3,500 meters). At the higher "altitude coincidences were found even up to distances of 300 meters. The data indicated that the total number of particles in a shower might occasionally exceed a million, and that the primary particle to which the entire shower presumably owes its origin must often have an energy of the order of 10^{16} ev.

Such showers have come to be called "Auger" showers. In a cloud chamber the particles produce parallel tracks resembling rain. Several investigations of the phenomenon have been made using various arrangements of counters, or of ionization chambers for the observation of bursts, but the data are not easy to analyze conclusively because of the threefold variation of (1) the particle density over the cross section of a shower, (2) the sizes of individual showers, and (3) the position of a shower relative to the apparatus.

Interesting data are reported by Lewis, and by Kingshill and Lewis.² Simultaneous or "coincident" bursts were observed in two ionization chambers having steel walls only 0.03 cm. thick, in the form of spheres 35 cm. in diameter. The frequency of bursts in these chambers was found to decrease rapidly with n , the number of particles entering the chamber and causing the burst; and their frequency was also much less near sea level than at higher altitudes. At Echo Lake (Colorado; 10,650 ft.) 150 bursts, each caused by 80 or more particles, occurred per hour in each chamber. Simultaneous bursts in both chambers occurred at the rate of 15 per hour when they were 1 meter apart; this number decreased smoothly with increasing separation of the chambers until it was only 4 or 5 per hour at a separation of 10 meters. Of larger bursts caused by 240 or more particles there were 9 per hour in each chamber; 0.6 per hour occurred simul-

¹ AUGER, *Comptes Rendus*, vol. 206, p. 1721; vol. 207, p. 228 (1938); *Jour. de Physique et le Radium*, vol. 10, p. 39 (1939); *Rev. Modern Phys.*, vol. 11, p. 288.

² LEWIS, *Phys. Rev.*, vol. 67, p. 228 (1945); KINGSHILL and LEWIS, *Phys. Rev.*, vol. 69, p. 159 (1946). Cf. also COCCONI, *Phys. Rev.*, vol. 70, p. 975 (1946).

aneously in both chambers at 1 meter of separation, and roughly 0.13 per hour at 10 meters. In the largest showers, at least 6,000 particles must have struck each square meter over an area of 100 square meters, so that the total number of particles must have been close to a million. At Chicago (altitude, 259 meters) the bursts were much less frequent; the ratio of the numbers at the two places varied with the size of the burst nearly as $n^{2.7}$, increasing from 8 for $n = 50$ to 400 for $n = 200$.

Presumably such showers are initiated by a single particle high up in the atmosphere. If this is so, and if the cascade theory of electron showers is applicable, more recent observations indicate that the energy of the initiating particle must occasionally range up to as much as 10^{17} ev or 10^8 Bev. The cascade theory predicts the presence in a shower of a high-density core, from which the density of the particles decreases outward. This conclusion from the theory appears to be borne out by the observations.¹

There is some reason to believe, however, that mesons may be an important constituent of Auger showers.² Clay, using several clusters of counters spaced a meter or two apart, looked for a meson component by dividing each cluster into two decks, one above the other, with 10 cm. of lead between them; only coincidences between two or more clusters and also between the two decks of each cluster were recorded. He concluded that about a seventh of the particles that arrived in wide showers were mesons. Thus it is not impossible that Auger showers differ in nature from the showers of diverging particles that are more commonly seen in cloud chambers.

(e) *Relation between Electrons and Mesons in the Atmosphere.*— In 1941 Schein, Jesse, and Wollan sent a counter telescope carried by balloons into the stratosphere and found that the vertical intensity was little affected by increasing the thickness of the absorbing lead from 4 to 18 cm.; also, by establishing the absence of coincidences with side counters, it was shown that very few showers were produced in the lead.³ The conclusion was drawn that at great heights the cosmic rays consist almost exclusively of mesons; at least, very few electrons of energy between 10^9 and 10^{12} ev are present. It was accordingly suggested that all the cosmic-ray electrons that are found in the atmosphere owe their origin ultimately to the mesons, being either generated by meson decay or ejected as negative "knock-on" electrons

¹ LAPP, *Phys. Rev.*, vol. 64, p. 129 (1943).

² CLAY, *Phytica*, vol. 9, p. 897 (1942); ROGOZINSKI, *Phys. Rev.*, vol. 65, p. 291, (1944).

³ SCHEIN, JESSE, and WOLLAN, *Phys. Rev.*, vol. 59, p. 615 (1941).

from atoms, and then multiplied to a limited extent by the cascade process.

Fair success has attended efforts to verify this hypothesis by calculating the number of electrons of various energies that should accompany the observed distribution of mesons in their passage downward through the atmosphere. The total number of electrons and slow mesons taken together was calculated by Rossi and Greisen and was found to agree roughly with the observed number of "soft" particles at low altitudes, but not at high.¹ Stanton, however, finds fair agreement for electrons of various energies above 50 Mev at elevations of 3,000 to 4,000 meters.² Lombardo and Hazen extended the agreement at sea level to the energy spectrum.³

The contribution of the knock-on electrons is relatively rather small, especially at higher altitudes in the atmosphere. Furthermore, multiplication in cascade showers is relatively unimportant; at sea level, shower electrons constitute probably not more than one-thirtieth of all electrons present. Thus, it appears that cosmic-ray electrons probably arise chiefly from decay of the mesons.

243. Conclusion.—In this chapter, as proposed at the beginning, we have discussed only the most important topics in the field of cosmic rays, and especially those concerning which definite conclusions appear to have been reached. In the literature the student will find information concerning other minor features, and many speculations in regard to the nature of the basic particles themselves.

For example, small *variations in cosmic-ray intensity* at a given station have been detected, in addition to the variation with temperature; but the significance of these variations is uncertain.⁴ There appears to be a diurnal variation with a maximum at noon and a minimum at midnight, the extreme range being roughly 0.4 percent. A slight variation in step with the sunspot cycle has also been detected.⁵ Such a variation may be caused either by the presence of cosmic rays coming directly from the sun or by the deflecting effects of the sun's magnetic field upon particles approaching from outside the solar system.

These variations are obviously important in connection with the problem of the *nature and origin of the primary cosmic rays*. What are

¹ Rossi and GBBISEN, *Phys. Rev.*, vol. 61, p. 121 (1942).

² STANTON, *Phys. Rev.*, vol. 66, p. 48 (1944).

³ LOMBARDO and HAZEN, *Phys. Rev.*, vol. 68, p. 59 (1945).

⁴ See *Rev. Modern Phys.*, vol. 11, pp. 153-190 (1939); WAFFLER, *Helv. Phys. Acta*, vol. 14, p. 215 (1941); DITPERIER, *Nature*, vol. 158, p. 196 (1946).

⁵ BROXON, *Phys. Rev.*, vol. 62, p. 508 (1942).

they, where do they come from, and how do they acquire their enormous energies?

It was shown in the balloon observations of Schein, Jesse, and Wollan that electrons "are not present in appreciable numbers at high altitudes,"¹ hence the primary particles cannot be electrons. Mesons are ruled out by their short finite life; they must be secondaries produced somehow by the primary particles. In low latitudes, more mesons are observed to arrive from directions west of the zenith than from directions east of it. Shower particles do not show this effect.² Now, as was explained in Sec. 233, the theory of the motion of charged particles through the earth's magnetic field indicates that such particles, if their energy lies in a certain range, should arrive at the top of the earth's atmosphere predominantly from the west or from the east according as they carry positive or negative electrical charges. Johnson concluded in 1938, from an analysis of all the data, that the part of the cosmic rays that causes the variation of ionization with latitude must consist entirely of *positive* particles.³ He favors the hypothesis put forward 4 years earlier by Compton and Bethe⁴ and mentioned in Sec. 241(c) that the primary particles are protons.

Mention should be made, however, of the remarkable speculation published by Millikan, Neher, and Pickering in 1942.⁵ They suppose that "in the depths of space" it is possible for a light nucleus to become converted into a pair of rather light charged particles (or perhaps photons), each of which carries as kinetic energy half of the rest energy of the original nucleus. From spectroscopic evidence it appears that in ring nebulae by far the commonest elements are hydrogen, helium, carbon, nitrogen, oxygen, and silicon; it may be inferred that these atoms are also common among those scattered through interstellar space. Conversion of the nuclei of these atoms into pairs would provide five bands of cosmic-ray particles capable of reaching the earth in spite of the sun's magnetic field. The energies of these bands, in billions of ev, would be as follows: He, 1.9; C, 5.6; N, 6.6; O, 7.5; Si, 13.2. The silicon rays could reach even the equator; the other four would form plateaus in the cosmic-ray spectrum, terminating at

¹ SCHEIN, JESSE, and WOLLAN, *Phys. Rev.*, vol. 59, p. 615 (1941).

² Cf. *Rev. Modern Phys.*, vol. 10, p. 228 (1938).

³ JOHNSON, *Rev. Modern Phys.*, vol. 10, p. 232 (1938). See also SWANN, *Frank. Inst. J.*, vol. 226, p. 757 (1938).

⁴ COMPTON and BETHE, *Nature*, vol. 134, p. 734 (1934).

⁵ MILLIKAN, NEHER, and PICKERING, *Phys. Rev.*, vol. 61, p. 397 (1942); vol. 63, p. 234 (1943); vol. 66, p. 295 (1944); *Nature*, vol. 151, p. 663 (1943). See also KUSAKA, *Phys. Rev.*, vol. 66, p. 354 (1944); WARREN, *Phys. Rev.*, vol. 66, p. 252 (1944).

various distances from the equator. Millikan, Neher, and Pickering believe that the existing data actually indicate the existence of such plateaus.

Perhaps the riddle of the primary cosmic rays will not be solved until observations can be extended with the aid of rockets to much higher altitudes than have yet been reached.

In the meantime, there are signs that the center of greatest interest in experimental physics may now be shifting back into the laboratory. Plans are on foot to produce ever larger voltages in particle accelerators of various sorts. An era of high-speed particles seems to be dawning in the laboratory; and also an era of big physics, with millions of dollars spent on a single instrument. It will be a dramatic day when mesons are for the first time generated in the laboratory. Further surprises may also be in store for the physicist as the available energy of bombarding particles rises in the laboratory to a billion volts and more.

APPENDIX I

ISOTOPIC CONSTITUTION OF THE ELEMENTS

The data in this Appendix have been taken from the sources listed below, chemical atomic weights being taken from the first paper listed and relative abundances of isotopes from the last.

HAHN, FLAGGE, and MATTAUCH, *Phys. Zeds.* vol. 41, p. 1 (1940); BARKAS, *Phys. Rev.* 55, p. 691 (1939); DEMPSTER, *Phys. Rev.*, vol. 53, pp. 74, 869 (1938); ASTON, *Roy. Soc. Proc.* vol. 163, p. 391 (1937); POLLARD, *Phys. Rev.*, vol. 57, p. 1186 (1940); BIRGE, *Rev. Modern Phys.*, vol. 13, p. 233 (1941); SEABORG, *Rev. Modern Phys.*, vol. 16, p. 1 (1944).

In computing the chemical atomic weight from the isotopic weights and the data on relative abundance, the factor $16/16.0044 = 1/1.000275$ was used to convert from a.m.u. to chemical units. The masses include the circumnuclear electrons.

Element Z	Mass number A of isotopes	Atomic mass, a. m. u. (O ¹⁶ = 16)	Packmg fraction × 10 ⁴	Relative abundance, percent	Chemical atomic weight (O = 16)	
					Computed from relative abundance	Observed
H 1	1	1.00813	81.3	99.98	1.0080	1.0080
He 2	2	2.01473	73.6	0.02		
Li 3	4	4.00889	9.7	100		
Be 4	6	6.0109	28.2	7.5		
	7	7.0180	25.7	92.5	6.943	6.940
	9	9.0150	16.7	100	9.013	9.02
B 5	10	10.0160	16.0	18.4	10.829	10.82
	11	11.0129	11.7	81.6		
C 6	12	12.0040	3.3	98.9	12.012	12.010
	13	13.0077	5.9	1.1		
N 7	14	14.00750	5.4	99.62	14.007	14.008
	15	15.0049	3.3	0.38		
O 8	16	16.0000	0	99.76	16.0000	16.0000
	17	17.0045	2.7	0.04		
	18	18.005	3	0.20		
F 9	19	19.0045	2.4	100	18.9993	19.00
Ne 10	20	19.9988	-0.6	90.00	20.191	20.183
	21	20.9997	-0.1	0.27		
	22	21.9986	-0.6	9.73		
Na 11	23	22.9891	-1.7	100	22.990	22.987
Mg 12	24	23.9824	-3.2	77.4	24.323	24.32
	25	24.9838	-2.5	11.5		
	26	25.9808	-3.9	11.1		
Al 13	27	26.980	-3.7	100	26.983	26.97
Si 14	28	27.987	-4.6	89.6	28.125	28.06
	29	28.987	-4.5	6.2		
	30	29.983	-5.7	4.2		
P 15	31	30.984	-5.2	100	30.977	30.98
S 16	32	31.9823	-5.5	95.0	32.065	32.06
	33	32.983	-5.2	0.74		
	34	33.978	-6.5	4.2		
	36	35.978	-6.1	0.016		
Cl 17	35	34.980	-5.7	75.4	35.462	35.467
	37	36.978	-5.9	24.6		
A 18	36	35.978	-6.1	0.31	39.926	39.944
	38	37.974	-6.8	0.06		
	40	39.9750	-6.2	99.63		
K 19	39	38.975	-6.4	93.38	39.097	39.096
	40*	39.975	-6.5	0.012		
	41	40.974	-6.4	6.61		
Ca 20	40	39.974	-6.5	96.96	40.08	40.08
	42	41.971	-6.9	0.64		
	43	42.972	-6.5	0.15		
	44			2.06		
	46			0.003		
	48			0.19		

Element Z	Mass number A of isotopes	Atomic mass, a.m.u. (O ¹⁶ = 16)	Packing fraction × 10 ⁴	Relative abundance, percent	Chemical atomic weight (O = 16)	
					Computed from relative abundance	Observed
Sc 21 Ti 22	45	44.969	-6.9	100	44.96	45.10
	46	45.968	-7.0	7.95		
	47	...	...	7.75		
	48	47.965	-7.2	73.45		
	49	48.966	-7.2	5.51		
	50	49.963	-7.4	5.34	47.88	47.90
V 23 Cr 24	51	50.960	-7.9	100	50.95	50.95
	50	...	...	4.49		
	52	51.959	-7.9	83.78		
	53	52.957	-8.1	9.43		
	54	...	...	2.30	52.00	52.01
Mn 25 Fe 26	55	54.965	-8.1	100	54.94	54.93
	54	53.960	-7.4	6.04		
	56	55.961	-7.0	91.57		
	57	56.961	-6.9	2.11		
	58	...	...	0.28	55.85	55.84
Co 27 Ni 28	59	...	...	100		58.94
	58	...	...	67.4		
	60	59.959	-6.8	26.7		
	61	...	...	1.2		
	62	...	...	3.8		58.89
	64	...	...	0.88		
Cu 29	63	62.956	mean, -6.9	70.13	63.58	63.57
	65	64.955		29.87		
Zn 30	64	63.956	-6.9	50.9	65.33	65.38
	66	65.952	-7.4	27.3		
	67	...	...	3.9		
	68	67.956	-6.5	17.4		
	70	69.954	-6.5	0.5		
Ga 31	69	68.955	-6.5	61.2	69.70	69.72
	71	70.953	-6.6	38.8		
Ge 32	70	...	...	21.2		72.60
	72	...	...	27.3		
	73	...	...	7.9		
	74	...	...	37.1		
	76	...	...	6.5		
As 33 Se 34	75	...	...	100	...	74.91
	74	...	...	0.9		
	76	...	...	9.5		
	77	...	...	8.3		
	78	...	...	24.0		78.96
Br 35	80	...	...	48.0		79.916
	82	...	...	9.3		
	79	...	...	50.6		
	81	...	...	49.4		
	...	...	...	...		
Kr 36	78	77.945	-7.0	0.35	83.82	83.7
	80	...	...	2.01		
	82	81.939	-7.5	11.53		
	83	...	...	11.53		
	84	83.938	-7.3	57.11		
	86	85.939	-7.1	17.47		
Rb 37	85	...	...	72.8		85.48
	87*	...	...	27.2		
Sr 38	84	...	...	0.56		87.63
	86	...	...	9.86		
	87	...	...	7.02		
	88	...	...	82.56		
Y 39 Zr 40	89	...	...	100		88.92
	90	...	...	48.0		
	91	...	...	11.5		
	92	...	...	22		
	94	...	...	17		91.22
	96	...	...	1.5		

Element Z	Mass number A of isotopes	Atomic mass, a.m.u. (O ¹⁶ = 16)	Packing fraction × 10 ⁴	Relative abundance, percent	Chemical atomic weight (O = 16)	
					Computed from relative abundance	Observed
Cb 41	93			100		92.91
Mo 42	92			14.9		
	94			9.4		
	95	94.945	-5.8	18.1		
	96	95.946	-5.6	16.6		95.95
	97	96.945	-5.8	9.65		
	98	97.944	-5.7	24.1		
	100			9.25		
43						
Ru 44	96	95.946	-5.7	5.68		
	98			2.22		
	99	98.944	-5.7	12.81		
	100			12.70		101.7
	101			16.98		
	102			31.34		
	104			18.27		
Rh 45	103			100		102.91
Pd 46	102			8		
	104	103.946	-5.2	9.3	106.55	
	105	104.945	-5.2	22.6		
	106	105.945	-5.2	27.2		106.7
	108	107.943	-5.2	26.8		
	110	109.942	-5.2	13.5		
Ag 47	107	106.948	-4.8	51.9	107.87	107.880
	109	108.947	-4.8	48.1		
Cd 48	106			1.4		
	108			1.0		
	110			12.8		
	111			13.0		
	112			24.2		112.41
	113			12.3		
	114			28.0		
	116			7.3		
In 49	113			4.5		114.76
	115			95.5		
Sn 50	112			1.1		
	114			0.8		
	115			0.4		
	116	115.942	-5.0	15.5	118.70	
	117			9.1		
	118	117.939	-5.1	22.5		118.70
	119	118.938	-5.2	9.8		
	120	119.94	-6	28.5		
	122	121.944	-4.6	5.5		
	124	123.943	-4.6	6.8		
Sb 51	121			56		121.76
	123			44		
Te 52	120			<1		
	122			2.9		
	123			1.6		
	124			4.5		
	125			6.0		127.61
	126			19.0		
	128			32.8		
	130			33.1		
	127			100		126.92
	124			0.094		
	126			0.088		
	128			1.90		
	129	128.946	-4.2	26.23		
I 53 Xe 54	130			4.07		
	131			21.17		
	132	131.946	-4.4	26.96	(131.3)	131.3
	134			10.54		
	136			8.95		

Element Z	Mass number A of isotopes	Atomic mass, a.m.u. (O ¹⁶ = 16)	Packing fraction × 10 ⁴	Relative abundance, percent	Chemical atomic weight (O = 16)	
					Computed from relative abundance	Observed
Cs 55	133			100		132.91
Ba 56	130			0.101		
	132			0.097		
	134			2.42		
	135			6.59		
	136			7.81		
	137			11.32		
	138			71.66		137.36
La 57	139	138.955	-3.2	100	138.92	138.92
Ce 58	136			<1		
	138			<1		
	140			89		140.13
	142			11		
Pr 59	141			100		140.92
Nd 60	142			25.95		
	143			13.0		
	144			22.0		
	145			9.2		
	146	145.960	-2.8	16.5		144.27
	148	147.961	-2.7	6.8		
	150	149.967	-2.2	5.95		
Bi 61						
Sm 62	144			3		
	147			17		
	148*			14		
	149			15		
	150			5		150.43
	152			26		
	154			20		
Eu 63	151			49.1		152.0
	153			50.9		
Gd 64	152			0.2		
	154			2.80		
	155	154.977	-1.5	15.61		
	156	155.976	-1.5	20.59		156.9
	157	156.976	-1.5	16.42		
	158			23.45		
	160			20.87		
Th 65	159			100		159.2
Dy 66	158			0.1		
	160			1.5		
	161			22		162.46
	162			24		
	163			24		
	164			28		
Ho 67	165			100		164.935
Er 68	162			0.1		
	164			1.5		
	166			32.9		167.2
	167			24.4		
	168			26.9		
	170			14.2		
Tm 69	169			100		169.4
Yb 70	168			0.06		
	170			4.21		
	171			14.26		
	172			21.49		173.04
	173			17.02		
	174			29.58		
	176			13.38		
Lu 71	175			97.5		174.99
	176*			2.5		

Element <i>Z</i>	Mass number <i>A</i> of isotopes	Atomic mass, a.m.u. (O ¹⁶ = 16)	Packing fraction × 10 ⁴	Relative abundance, percent	Chemical atomic weight (O = 16)	
					Computed from relative abundance	Observed
Hf 72	174	..		0.18		178.6
	176			5.30		
	177			18.47		
	178			27.18		
	179			13.86		
	180			35.14		
Ta 73 W 74	181			100	..	180.86
	180			0.2 ±		
	182			22.6		
	183			17.3		
	184			30.1		
	186			29.8		
Re 75	185			38.2		186.31
	187			61.8		
Os 76	184			0.018		190.2
	186			1.59		
	187			1.64		
	188			13.3		
	189			16.1		
	190			26.4		
	192			41.0		
		190.038 192.038	+2.0 2.0			
Ir 77	191	191.040 193.041	2.1 2.1	38.5	192.22	193.1
	193			61.5		
Pt 78	192	194.040 195.040 196.039 198.044	2.0 2.0 2.0 2.2	0.8	195.14	195.23
	194			30.2		
	195			35.3		
	196			26.6		
	198			7.2		
Au 79	197	197.039	2.0	100	196.99	197.2
Hg 80	196	200.028	1.4	0.15		200.61
	198			10.1		
	199			17.0		
	200			23.3		
	201			13.2		
	202			29.6		
	204			6.7		
Tl 81	203	203.057 205.057	2.8 2.8	29.1	204.42	204.39
	205			70.9		
Pb 82	204	204.058 207 208.057	2.8 2.7	1.5	..	207.21
	206			23.6		
	207			22.6		
	208			52.3		
Bi 83	209	209.055	2.6	100	209.00	209.00
Po 84	222*	..	..	..	..	222
Rn 86	222*	..	..	..	..	222
.. 87	222*	..	..	..	..	222
Ra 88	223* 224* 226* 227*	..	..	..	..	226.05
Ac 89	223* 224* 226* 227*	..	..	..	..	226.05
Th 90	232*	232.12	5.2	100	232.06	232.12
Pa 91	231*	..	..	..	..	..
U 92	234*	238.14	5.6	0.006	238.06	238.07
	235*			0.71		
	238*			99.28		

* Radioactive.

APPENDIX II

THE PERIODIC TABLE

Periods	I	II	III	IV	V	VI	VII	VIII	
I	1 H 1, 2							2 He 4	(a) 40, 44, 42, 48, 43, 46 (b) 80, 78, 76, 82, 77, 74 (c) 84, 86, 82, 83, 80, 78 (d) 98, 96, 95, 92, 94, 97, 100 (e) 114, 112, 111, 110, 113, 116, 106, 108 (f) 120, 118, 116, 119, 117, 124, 122, 112, 114, 115 (g) 130, 128, 126, 125, 124, 122, 123, 120 (h) 132, 129, 131, 134, 136, 130, 128, 124, 126 (i) 138, 137, 136, 135, 134, 130, 132 (j) 180, 178, 177, 179, 176, 174 (k) 202, 200, 199, 201, 198, 204, 196
II	3 Li 7, 6	4 Be 9	5 B 11, 10	6 C 12, 13	7 N 14, 15	8 O 16, 18, 17	9 F 19	10 Ne 20, 22, 21	
III	11 Na 23	12 Mg 24, 25, 26	13 Al 27	14 Si 28, 29, 30	15 P 31	16 S 32, 34, 33, 36	17 Cl 35, 37	18 Ar 40, 36, 38	
IV	19 K 39, 41, 40	20 Ca (a)	21 Sc 45	22 Ti 47, 49, 50	23 V 51	24 Cr 52, 53, 50, 54	25 Mn 55	26 Fe 56, 54, 57, 58	27 Co 59
	29 Cu 63, 65	30 Zn 64, 66, 68, 67, 70	31 Ga 69, 71	32 Ge 72, 74, 73, 76	33 As 75	34 Se (b)	35 Br 79, 81	36 Kr (c)	
V	37 Rb 85, 87	38 Sr 86, 87, 84	39 Y 89	40 Zr 92, 94, 91, 96	41 Nb 93	42 Mo (d)	43 Tc (e)	44 Ru 102, 101, 104, 100, 99, 96	45 Rh 103
	47 Ag 107, 109	48 Cd (e)	49 In 115, 113	50 Sn (f)	51 Sb 121, 123	52 Te (g)	53 I 127	54 Xe (h)	
VI	55 Cs 133	56 Ba (i)	57-71 * (j)	72 Hf (j)	73 Ta 181	74 W 182, 183, 180	75 Re 187, 185	76 Os 192, 190, 189, 188, 187, 186, 184	77 Ir 193, 191
	79 Au 197	80 Hg (k)	81 Tl 205, 203	82 Pb 206, 207, 204	83 Bi 209	84 Po 209	85 At 209	86 Rn 222	87 La 138, 139, 137, 136, 135, 134, 133, 132, 131, 130, 129, 128, 127, 126, 125, 124, 123, 122, 121, 120, 119, 118, 117, 116, 115, 114, 113, 112, 111, 110, 109, 108, 107, 106, 105, 104, 103, 102, 101, 100, 99, 98, 97, 96, 95, 94, 93, 92, 91, 90, 89, 88, 87, 86, 85, 84, 83, 82, 81, 80, 79, 78, 77, 76, 75, 74, 73, 72, 71, 70, 69, 68, 67, 66, 65, 64, 63, 62, 61, 60, 59, 58, 57, 56, 55, 54, 53, 52, 51, 50, 49, 48, 47, 46, 45, 44, 43, 42, 41, 40, 39, 38, 37, 36, 35, 34, 33, 32, 31, 30, 29, 28, 27, 26, 25, 24, 23, 22, 21, 20, 19, 18, 17, 16, 15, 14, 13, 12, 11, 10, 9, 8, 7, 6, 5, 4, 3, 2, 1
VII	87 Ra (223, 224, 226)	88 Ac 227	89 Th 232	90 Pa 231	91 U 238, 235, 234	92 Np 237	93 Pu 239	94 Am 241	95 Cm 247

Explanation.—The Roman numerals in the top row designate the "columns" of the periodic table. The Roman numerals at the left give the several "periods" according to Bohr's arrangement (Fig. 93), and the italic numerals in the upper right-hand corners of the several spaces give the ordinal numbers of the elements in these periods. Atomic numbers are indicated by bold-faced type in the upper left-hand corners. Mass numbers of the constituent isotopes are given by the numbers at the bottoms of the spaces, in the order of decreasing abundance (except for Ra).

APPENDIX III

FIRST IONIZATION POTENTIALS (WHERE KNOWN), LOWEST SPECTRAL TERMS, AND ELECTRON CONFIGURATION OF THE ELEMENTS

X-ray symbol n, l Electron wave function			→ K 1, 0 1s	→ L I 2, 0 2s	L II, III 2, 1 2p	M I 3, 0 3s	M II, III 3, 1 3p	M IV, V 3, 2 3d	N I 4, 0 4s	N II, III 4, 1 4p	N IV, V 4, 2 4d	N VI, VII 4, 3 4f	O I 5, 0 5s	O II, III 5, 1 5p	O IV, V 5, 2 5d	P I 6, 0 6s	P II, III 6, 1 6p	P IV, V 6, 2 6d	Q I 7, 0 7s
Element		First ionization potential	Lowest term																
H 1		13 575	¹ S _{1/2}		1														
He 2		24 544	¹ S ₀		2														
Li 3		5.383	² S _{1/2}		2	1													
Be 4		9 207	¹ S ₀		2	2													
B 5		8 30	² P _{1/2}		2	2	1												
C 6		11 249	² P ₀		2	2	2												
N 7		14 52	⁴ S _{3/2}		2	2	3												
O 8		13 595	² P ₂		2	2	4												
F 9		18.7	² P _{3/2}		2	2	5												
Ne 10		21 529	¹ S ₀		2	2	6												
Na 11		5 136	² S _{1/2}		2	2	6	1											
Mg 12		7 634	¹ S ₀					2											
Al 13		5 976	² P _{1/2}					2	1										
Si 14		8.140	² P ₀					2	2										
P 15			⁴ S _{3/2}					2	3										
S 16		10 3	² P ₂					2	4										
Cl 17		13 00	² P _{3/2}					2	5										
A 18		15.733	¹ S ₀					2	6										
K 19		4 333	² S _{1/2}		2	2	6	2	6	1									
Ca 20		6 103	¹ S ₀							2									
Sc 21		6 7	² D _{3/2}							2									
Ti 22		6 82	² F _{7/2}							2									
V 23		6.78	⁴ F _{3/2}							3									
Cr 24		6.75	² S _{1/2}							5	1								
Mn 25		7 419	⁶ S _{5/2}							5	2								
Fe 26		7.85	¹ D ₄							6	2								
Co 27		8.5	⁴ F _{9/2}							7	2								
Ni 28		7.62	² F ₄							8	2								

FIRST IONIZATION POTENTIALS (WHERE KNOWN), LOWEST SPECTRAL TERMS, AND ELECTRON CONFIGURATION OF THE ELEMENTS.—
(Continued)

X-ray symbol a, l Electron wave function			→ K 1, 0 1s	→ L I 2, 0 2s	→ L II, III 2, 1 2p	→ M I 3, 0 3s	→ M II, III 3, 1 3p	→ M IV, V 3, 2 3d	→ N I 4, 0 4s	→ N II, III 4, 1 4p	→ N IV, V 4, 2 4d	→ N VI, VII 4, 3 4f	→ O I 5, 0 5s	→ O II, III 5, 1 5p	→ O IV, V 5, 2 5d	→ P I 6, 0 6s	→ P II, III 6, 1 6p	→ P IV, V 6, 2 6d	→ Q I 7, 0 7s					
Element	First ionisation potential	Lowest term																						
Cu 29	7 712	$^1S_{1/2}$	2	2	6	2	6	10	1															
Zn 30	9 378	$^1S_0$	28 electrons copper core							2														
Ga 31	5 988	$^2P_{1/2}$								2	1													
Ge 32	8.11	$^2P_0$								2	2													
As 33	10	$^4S_{3/2}$								2	3													
Se 34	9 5	$^3P_2$								2	4													
Br 35	11 8	$^3P_{3/2}$								2	5													
Kr 36	13 976	$^1S_0$	2	2	6	2	6	10	2	6			1											
Rb 37	4 170	$^2S_{1/2}$	36 electrons krypton core											2										
Sr 38	5 685	$^1S_0$																	2					
Y 39	6.6	$^2D_{3/2}$																1		2				
Zr 40	6 94	$^2F_7/2$																2		2				
Cb 41		$^4D_{1/2}$																4		1				
Mo 42	7 372	$^1S_0$																5		1				
43																		6		1				
Ru 44		$^3F_4$																7		1				
Rh 45	7 7	$^4F_{9/2}$																8		1				
Pd 46	8 3	$^1S_0$																10						
Ag 47	7 563	$^2S_{1/2}$	2	2	6	2	6	10	2	6	10		1											
Cd 48	8 979	$^1S_0$	46 electrons silver core											2										
In 49	5 776	$^2P_{1/2}$																2		1				
Sn 50	7 327	$^2P_0$																2		2				
Sb 51	8.5	$^4S_{3/2}$																2		3				
Te 52		$^3P_2$																2		4				
I 53	10.	$^2P_{3/2}$																2		5				
Xe 54	12 110	$^1S_0$	2	2	6	2	6	10	2	6	10		2	6										
Cs 55	3 887	$^2S_{1/2}$	54 electrons Xenon core													1								
Ba 56	5 202	$^1S_0$																				2		
La 57		$^2D_{3/2}$																						

Ce 58		$^2H_{1/2}$	2	2	6	2	6	10	2	6	10	1	2	6	1	2		
Pr 59		$^4K_{11/2}$	46 electrons silver core												2	6	1	2
Nd 60		$^4L_9$													3		1	2
P 61		$^4L_{11/2}$													4		1	2
Sm 62		$^4K_4$													5		1	2
Eu 63		$^2H_{3/2}$													6		1	2
Gd 64		$^2D_7$													7		1	2
Tb 65		$^4H_{17/2}$													8		1	2
Dy 66		$^4K_{10}$													9		1	2
Ho 67		$^4L_{13/2}$													10		1	2
Er 68		$^4L_{10}$													11		1	2
Tm 69		$^4K_{17/2}$													12		1	2
Yb 70		$^2H_6$													13		1	2
Lu 71		$^2D_{3/2}$													14		1	2
Hf 72		$^2F_{5/2}$	2	2	6	2	6	10	2	6	10	14	2	6	2	6	2	
Ta 73		$^2F_{3/2}$	68 electrons hafnium core												3		2	
W 74		$^4D_6$													4		2	
Re 75		$^4S_{3/2}$													5		2	
Os 76		$^4D_4$													6		2	
Ir 77		$^4F_4$													7		1	
Pt 78	8 9	$^2D_{3/2}$													9		0	
		$^2D_5$													9		1	
Au 79	9 22	$^2S_{1/2}$	2	2	6	2	6	10	2	6	10	14	2	6	10	1		
Hg 80	10 419	$^1S_0$	78 electrons gold core												2		2	
Tl 81	6.098	$^2P_{1/2}$													2		1	
Pb 82	7.404	$^3P_0$													2		2	
Bi 83	8.0	$^4S_{3/2}$													2		3	
Po 84		$^3P_1$													2		4	
— 85		$^3P_{3/2}$													2		5	
Rn 86	10 730	$^1S_0$													2		6	
— 87		$^2S_{1/2}$	2	2	6	2	6	10	2	6	10	14	2	6	10	2	6	1
Ra 88	10.2	$^1S_0$	86 electrons radon core															2
Ac 89		$^2D_{3/2}$																2
Th 90		$^3F_4$																2
Pa 91		$^3F_{3/2}$																2
U 92		$^1D_2$																4

Note.—The ionization potentials are mostly calculated, using modern values of the constants [Eq. (147a) in Sec. 108], from data in Bacher and Goudsmit, "Atomic Energy States—As Derived from the Analyses of Optical Spectra," 1932.

SOME USEFUL CONSTANTS AND RELATIONS

(Except in a_0, mc^2, M_1c^2, kT at 15° , the last figure given is uncertain)

c	velocity of light ²	2.9978×10^{10} cm. sec. ⁻¹
F	the faraday ¹	$9,648.8$ e.m.u. mole ⁻¹
e	electronic charge ¹	4.803×10^{-10} e.s.u.
m	electronic mass (Sec. 38)	0.9107×10^{-27} gram
e/m	ratio of electronic charge to mass ¹	1.7592×10^7 e.m.u. gram ⁻¹
M_{atom}	mass of atom of atomic weight 1 (Sec. 195)	1.6604×10^{-24} gram
M_1	atomic mass unit (Sec. 195)	1.6599×10^{-24} gram
h	Planck's constant ²	6.610×10^{-27} erg sec.
R	universal gas constant.	$\begin{cases} 8.315 \times 10^7 \text{ ergs deg.}^{-1} \text{ mole}^{-1} \\ 82.06 \text{ cm.}^3 \text{ atm. deg.}^{-1} \text{ mole}^{-1} \end{cases}$
N_0	Avogadro's number (Sec. 42)	6.023×10^{23} mole ⁻¹
k	Boltzmann's constant (Sec. 42)	1.381×10^{-16} erg deg. ⁻¹
σ	Stefan's constant ¹	5.735×10^{-5} erg cm. ⁻² sec. ⁻¹ deg. ⁻⁴
R_∞	Rydberg's wave number ¹	1.0973730×10^5 cm. ⁻¹
M_H/m	ratio of masses of hydrogen atom and electron (Sec. 97)	1.837
$a_0 = \frac{h^2}{4\pi^2me^2}$	radius of smallest Bohr orbit in hydrogen (Sec. 94)	5.27×10^{-9} cm.
$\alpha = \frac{2\pi e^2}{ch}$	"fine-structure constant" (Sec. 135)	$0.007314 = 1/136.7$
I	electron-volt (ev, Sec. 96) = 1.601×10^{-12} erg	
Mc^2	rest energy of 1 atomic mass unit (Sec. 206)	932×10^6 electron-volts
mc^2	rest energy of the electron (Sec. 206)	0.511×10^6 electron-volts
kT at 15°C.	(T = absolute temperature = 288.1°K.)	0.0248 electron-volts
If $h\nu = eV_s/300$	(V_s = potential difference in volts or energy in electron-volts), $\lambda = c/\nu$, $\lambda_A = \lambda \times 10^8$ = wave length in angstroms, $\tilde{\nu} = 1/\lambda$ or waves per cm. (Sec. 99),	

$$\lambda_A = \frac{12,378}{\tilde{\nu}}, \quad \tilde{\nu} = 8,079 V_s.$$

¹BIRGE, *Rev. Modern Phys.*, vol. 13. p. 233 (1941).
²DTNNINGTON, *Rev. Modern Phys.*, vol. 11, p. 65 (1939).

INDEX

Boldface numbers indicate references of special importance.

A

A coefficient, Einstein's, 281-283
 Aberration of light, 29
 Abraham, 55n., 61n.
 Absorption of light, 224
 coefficient of, 145n., 502
 by molecules, 402, 403, 406, 409, **411**
 Absorption limits (*see* X-rays)
 Absorptivity, 142
 Activity, definition of, 625
 Adams, 53n.
 Ader, 450n.
 Adiabatic principle, **372**
 Ageno, 653n.
 Agostino, 618n.
 Air pump, invention of, 20
 Aiya, 725
 Alberti, 85n.
 Alhazen, 10
 Alkali metals, doublet separations of,
 338
 spectra of, 321-329, 335-338
 term energies of, 326
 Allen, 438, 646
 Allison, 477n., 499n., 505n., (fig.) 513,
 517
 Alpha rays or particles, 554-561
 range of, 557-560
 scattering of, 197-199
 anomalous, 589
 theory of, 594-596
 (*See also* Nuclear reactions)
 Alvarez, 644
 Amaldi, 615n., 618
 Ampere, 38
 Anaxagoras, 5
 Anderson, 598, 599, 600n., 693, (fig.)
 694, 714, (fig.) 715
 Andrews, 505n.
 Angstrom, 185, 188

Angstrom (= 10^{-8} cm.), 185
 Angular momentum, of atom, 318-321,
 348
 with one electron, **273-276, 280**
 in Bohr's theory, 205-209
 of electron in central field, 290
 in magnetic field, **378**
 measurement of, **390**
 of molecule, 401, **405, 413**
 of nucleus, 392, **393**
 table of, 394
 orbital and spin, **286, 332**
 principle of, 28
 as vector, **273-276**
 Arago, ~~36~~^{*}, 41
 Archimedes, 9
 Argon, central field for, 295
 Aristarchus, 9
 Aristotle, 6M), 14
 Aston, 545, 546, 550w., 552n., 584, 733
 Atom, Bohr theory of, 204-210, 233
 early views of, 195
 nuclear theory of, 198-203
 static, 299
 wave-mechanical theory of, for com-
 plex atoms, 288, 293-298
 for one-electron atoms, 272-280
 Atomic heat, 424
 Atomic mass unit (a.m.u.), 550, 585, 742
 Atomic numbers, **202**, 298, 480, 733
 (table)
 Atomic weights, 551
 table of, **733**
 Auger, 534, 659n., **699, 713, 727**
 Auger effect, 533-536, 570 (in nucleus)
 Avogadro's number, 85, 163, **742**

B

B coefficient, Einstein's, **281-283**
Bhabha, 702n., 705n., 725
Bacher, 230n., 336w., 623n., 624n.,

- Back, 377, 387, 395n., 396
 Bacon, Roger, 10
 Backlin, 84, 527n.
 von Baeyer, 569n.
 Bainbridge, 548, (fig.) 549
 Baker, 623n., 624n.
 Balard, 36
 Baldwin, 626n., 638, 654n.
 Balmer, 188, 189
 Balmer series, 188, (photographs) 189, 215, 343, 344
 Bands, 399
 branches and systems of, 415-416
 electronic, 401, 415-419, (photo-graphs) 418
 head of, 399, 417, 418
 from isotopes, 419
 rotation, 400, 409, 419
 vibration-rotation, 400, 404
 effect of temperature on, 410
 in relation to heat of dissociation, 409
 in relation to rotation bands, 409
 Barkas, 733
 Barkla, 196, 459, 462, 463, 465-468
 Barnes, 536n.
 Barometer, invention of, 20
 Barschall, 642n.
 Bartholinus, 36
 Battat, 624n.
 Bearden, 476, 525?i., 527n., 540n.
 Beattie, 437n.
 Becker, J. A., 517, 576
 Becker, R., 55n.
 Becker, R. A., 646n., 648
 Becquerel, 554
 Bennett, 664n., 678
 Benoist, 185n., 455
 Berlin, 615n.
 Bernouilli, 28
 Bertin-Sans, 457
 Beta rays or particles ($/3\pm$), 554, 555, 567, 628-629, 645, (pairs) 647-648
 origin of, 570, 595
 primary and secondary, 567, 570
 (See also Nuclear reactions)
 spectra of, 569
Betatron, 613
 Bethe, 255, 560n., 580, 588n., 602n., 609n., Q26n., 703, 709n., 730
 Bierens de Hahn, 181ra.
 Biot, 36, 38
 Birge, 74, 84n., 183n., 192n., 21 In., 212n., 217, 499n., 733, 742
 Black, 29
 Black body or surface, 142
 Black-body radiation, 145-146, 154-160
 magnitude of, 151, (fig.) 158, 179, 183
 Planck's law of, 178-183, 283
 Rayleigh-Jeans formula for, 160, 169
 Wien's formula for, 159
 Blackett, 579, 600rc., 686, 718n.
 Bless, 518
 Blewett, 613n.
 Bloch, 518n., 533n., 702n.
 Bohm, 615n.
 Bohr, 184, 203, 204-213, 233, 234, 528, 652
 Bollman, 499n.
 Boltzmann, 149
 Boltzmann distribution law, 231
 Boltzmann's constant, 97, 163, 742
 Bonner, 630, 641, 644, 649
 Booth, 653n.
 Born, 112n., 236, 438
 Bothe, 518, 521n., 569n., 576, 682, 683
 Bowen, 680, (fig.) 681
 Bradley, 29
 Bragg, 470, 473, 474, 478, 479, 505n.
 Bragg planes, 470
 Bragg's law, 472
 Brahe, Tycho, 17, 29
 Breadth of spectral lines, 396-399, 536
 Breit, 593n.
 Bretscher, 580
 Brewster, 14n.
 Briekwedde 393n.
 Brode, 690
 Broglie, de, 236-239, 242, 245, 259
 Broussard, 697n.
 Brbxon, 729n.
 Brubaker, 641
 Brtiche, 675n.
 Buguet, 464
 Buisson, 397
 Bunsen, 36
 Burkhardt, 519
 Burmeister, 407
 Bursts of ionization, 605

C

- Cabeo, 20
 Cady, 54In.
 Calcite crystals, 476, 477
 Cameron, 660-662
 Canal rays, 388
 Cario, 227
 Carlisle, 37
 Carlson, 705n.
 Carnot, 33
 Cathode rays, 80
 Cavendish, 30, 31
 Cavendish Laboratory, 48
 Celsius scale, 29
 Chadwick, 555n., 556n., 563n., 572n.,
 574, 577-580, 600n., 615, 639-641,
 689n., 71In.
 Chain reactions, 656
 Chakrabarty, 705n.
 Characteristic frequencies, 429-431, 439
 Characteristic functions, 270
 Characteristic temperatures, 434
 Characteristic values, 270
 Characteristic X-rays, 477-486
 Charge, field of accelerated, 63-64
 field of moving, 60, 134
 Charge cloud, 277, 289, (fig.) 295, 297
 Charged particles, energy loss by, 701,
 (plot) 702
 ionization by, 689-691, 701
 measurement of charge, energy, mass,
 686-687, (plot) 688
 motion in magnetic field, 508, 667,
 686, (plot) 688
 of dipole, 667
 of earth, 673
 observations on individual, 571-573,
 681-686
 radiation by, 702-704
 Charlton, 614n.
 Chemical compounds of certain ele-
 ments, 303-307
 Childs, 419n.
 Chlorine, specific heat for, 441, 442, 449
 Clark, 470n.
 Classical theory, difficulties of, 94-96,
 107
 Clausius, 34, 48
 Clay, 664n., 667, 726n., 728
 Cleaves, 409n.
 Cloud chamber, 572
 Coblenz, 179n., 183n.
 Cockroft 605-608
 Cohesion, 308
 Coincidences, method of, 682
 Colby, 559n.
 Collins, 638
 Collisions of the second kind, 225
 Color vision, 48
 Compton, A. H., 463, 477n., (fig.) 513,
 514, 515, 517, 518, 525, 527, 659n.,
 664, 666, 678, 679, 730
 Compton, K. T., 88, 89
 Compton effect, 514-522, 566, 602
 Condon, 396
 Configuration, 347
 Constable, 577n., 639-641
 Contraction, in space and time, 116,
 122-124
 Lorentz-Fitzgerald, 116-117
 Cooksey, 61In.
 Coolidge, 454n.
 Cooper, J. N., 536n.
 Cooper, L., 8n.
 Copernican system, 11, 16, 18, 29
 Copernicus, 11
 Cork, 627n.
 Coruscles, 83
 Correspondence principle, 213
 Corson, 652n., 690
 Cosmic rays, 659-731
 absorption of, in air, 660, 662
 of particles, 683, 685, 698, 715, 718
 in water, 661, 664
 energy of, 680
 ionization due to, 659-667, 682
 bursts of, 695
 variation of, with altitude, 664
 with barometric pressure, 665
 with latitude, 667, 678
 with temperature, 679, 718
 with time, 729
 mesons (see Mesons in atmosphere)
 nature and origin of, 666, 680, 712,
 729
 neutrons, 709
 particles, 681-686
 penetrating and soft component[^] of,
 698, 716

Cosmic rays, protons, 710
 showers of, 692-700, 707-709
 (Auger) 727-728, 730
 stars, 711
 statistics on, 722-726, 730
 transmission curves for, 697, 709
 Coster, 532n., 534n.
 Coulomb, 30, 31
 Coulomb field, 289-295
 Counter telescope, 685
 Counters, Geiger or valve, 571
 Coupling, *jj*, 361
 LS or Russell-Saunders, 345-357
 in molecules, (AS) 413, («) 414
 transitional, 365
 Crane, 603n.
 Crussard, 713, 716
 Crystal diffraction grating, 468, (table)
 477
 Curie, Marie, 554
 Curie, Pierre, 554
 Curie, definition of, 564
 Curie-Joliot, 576, 600n., 603-605
 Cuykendall, 505n.
 Cyclotron, 610
 Czerny, 403

D

D, 349
d, 291
 D'Agostino, 618n.
 Da Vinci, Leonardo, 10
 Darwin, 286n.
 Das Gupta, 572n.
 Davis, 228
 Davisson, 248n., 250, 253, 254, 256
 Davy, 33, 40
 De Broglie, 236-239, 242, 245, 259
 Debye, 432
 Dee, 606n.
 Degeneracy, 273, 276, 291, 298, 348, 352
 nuclear, 393
 Degrees of freedom, 160
 energy of, 161, 163
 of waves in an enclosure, 164
 Democritus, 6
 Dempster, 546, 547n., 548, 733
 Dennison, 615n.
 Descartes, 22

Dessauer, 654n.
 Deuterium (heavy hydrogen), 393
 Deuteron (nucleus of heavy hydrogen),
 580
 disintegration of, 580, 637
 range of, 561
 Deutsch, 645
 Dewar, 187
 Dieke, 422
 Dirac, 236, 286
 Dispersion, 22
 Displacement currents, 48, 50, 53
 Displacement law, 157, 182
 Dissociation, heat of, 409-410
 Distribution law, Boltzmann, 231
 Doan, 525, 527
 Doppler, 36
 Doppler effect (moving mirror), 154
 on spectral lines, 124, 397
 Dow, 559n.
 Drinkwater, 344n.
 Drude, 100
 Druyvesteyn, 532n.
 Du Fay, 30
 Du Mond, 499n., 519
 Duane, 496n., 517
 DuBridge, 92n, 98, 107n.
 Dulong and Petit, law of, 424
 Dunlap, 618n., 691n.
 Dunning, 619n., 653w.
 Dunnington, 84, 181n., 499n., 742
 Dushman, 260n., 271n.

E

Eckart, 255
 Edison, 96
 Edlen, 218
 Edwards, 409n.
 Ehrenfest, 713
 Eigenvalues (eigenfunctions), 270
 Eighteenth century, physics in, 28-32
 Einstein, 91, 281, 283, 427, 430
 on relativity, 109, 117-120, 135-138
 Einstein frequency condition, 205
 Electricity, discoveries in, 30-32, 36-38
 Electrolysis, 37, 43
 Electromagnetic field, 48, 53-55
 energy in, 53
 momentum in, 53-54, 59

- Electromagnetic field, theory of, 33, 48-50
- Electromagnetic waves, intensity of, 58
- Electron, 74
 charge, 83-85, 742
 discovery of, 80-83
 e/m , 81-82, 85, 742
 emission due to photons, 60, 94-96
 (See also Photoelectric effect)
 energy loss by, (plot) 702
 energy, velocity and lip , (plot) 688
 mass, 84, 85, (m/M_H) 217, 551, 742
 motion in magnetic field, 78
 range in air or Pb, 705
 rest energy, 585, 742
- Electron microscope, 242
- Electron-volt (ev), 214, 742
 relation to wave number, 222
- Electrons in complex atoms, 288-298, 348-349, 360, 362
 transitions of, 316-318, 363
 creation of, 582, 600-601
 emission by photons, 601
 equivalent, 360
 free, 100-108
 K , L , M , 481
 in molecules, 412, 413
 recoil, 518, 521, 566-567
- Elliott, 645n.
- Ellis, 555rc., 556n., 563n., 565-567, 569, 572n., 689n.
- Elster, 72, 73, 87, 96, 659
- Emissive power, 140
- Emissivity, 145
- Empedocles, 5
- Ende, van den, 437n.
- Endoergic and exoergic reactions, 630-632
- Energy, conservation of, 33, 34, 44, 131
 per degree of freedom, 161
 electromagnetic, 53, (waves) 57
 inertia of, 130, 581
 kinetic, relativistic, 129, 509, 687
 loss by charged particles, 65
 of oscillators in equilibrium, 171-175, 177
 relativistic, 129-132
 released by fission, 654, 656-658
- Energy, released in nuclear reactions, 575, 608
 calculation of, 608, 630
 rest, 130, 585
 of thermal agitation, 101
- Energy levels, 175-177, 215, 270
 in cm.⁻¹, 214, 222
 of harmonic oscillator, 271
 in magnetic field, 374, 377, 379, 387
 molecular, 401, 405, 408, 412-416
 nuclear (see Nucleus)
 of one-electron atom, 208, 273, 339
 of uranium, 492
 width of, 597, 640
 X-ray, 487-488, 493, 532, (resonance) 537
- Epstein, 389, 675
- Equipartition of energy, 161
- Estermann, 258
- Ether, 117
- Euler, 28
- ev (= electron-volt), 214, 742
- Everblud, 593n.
- Exchange effect, 308
- Excitation of atoms, 220-230
- Excitation potentials, 223
- Exclusion principle, 296
- Expansion chamber, 572
- F
- F , 349
- /, 291
- Fabry, 185n., 397
- Fahie, 14n.
- Fahrenheit, 29
- Falling bodies, laws of, 8, 14
- Fankuchen, (fig.) 258
- Faraday, 38-45, 74, 75
- Faraday (unit of charge), 74, 85, 742
- Faraday effect, 44
- Farkas, 449n.
- Feather, 579n., (fig.) 617, 691n.
- Fermat's principle, 242
- Fermi, 615n., 617, 618, 650, 654, 655n., 657
- Fermi-Dirac gas, 101
- Field, Coulomb, 295
 modified central, 289, 294-295
- Field currents, 99

- Fine structure, in alkali-type sp<
 335
 in molecular spectra, 414, 415
 for one-electron atoms, 342
 Fine structure constants, 340, 742
 Fink, 619n.
 Fitzgerald, 116, 117
 Fizeau, 112
 Flerov, 655n.
 Flugge, 552n., 733
 Fluorescence, 226
 Forsterling, 439n.
 Foldy, 615n.
 Foote, 221, 222
 Force, 28, 129
 Fortrat diagram, 417
 Foucault, 36
 Fowler, A., 329n., 344n.
 Fowler, R. H., 438n.
 Fowler, W. A., 646n., 647-649
 Frames of reference, 110
 inertial (or Galilean), 136
 Franck, 224, 227n.
 Franklin, 30
 Fraunhofer, 36
 Frenkel, 332n.
 Fresnel, 35, 36, 112, 113
 Friedrich, 257, 469
 Frisch, 258, 651, 654
 Froman, (fig.) 685, 696n.
 Fryer, 621n.
- G
- Galileo, 12-17, 29
 Galvani, 37
 Gamertsfelder, 619n.
 Gamma-ray microscope, 268
 Gamma-ray spectra, 565, 567
 Gamma rays, 555
 definition of, 615
 emission of, 568
 excitation of, 633-637
 internal conversion of, 567
 origin of, 567-569, 633
 scattering of, 511-522
 spectra, 565, 567
 Gant, 654n.
 Gas constant, 163, **742**
 Geiger, 199, **201-203, 521n., 556n.**
 Geitel, 72, 73, 87, 96, **659**
 Gellibrand, 20
 Gerlach, 389
 Geomagnetic latitude, definition of, 675n.
 Germer, 248n., 250, 253, 254, 256, (fig.) 258
 Ghosh, 572w.
 Giauque, 419n.
 Gilbert, 20
 Gill, 695
 Glancing angle, 471
 Glazebrook, 46n.
 Glockler, 423n.
 Gockel, 660
 Goldhaber, 580, 61 On., 646
 Goucher, 228, 229
 Goudsmit, 230n., 286, 336w.
 Gouy, 475
 Graaf, van de, 609, 610
 Gram-atom, 74
 Gram-ion, 74
 Gram-molecule, 74
 Graves, 697n.
 Gravitation, 20, 21, 26, 135-137
 Gray, L. H., 603
 Gray, S., 30
 Green, G., 37
 Green, G. K., 633
 Greenhill, 8
 Greinacher, 605n.
 Greisen, 723, 729
 Grosse, 653n.
 Grossman, 137
 Ground state = normal state, 208
 Guericke, 20
 Guggenheimer, 644n.
 Guth, 638
- H
- h , 175
 Hacman, 560n.
 Hafstad, 636n.
 Hafner, 654n.
 Haga, 457
 Hagenow, 463
 Hahn, 552n., **650, 652, 733**
 Half-life, 561
 Hall, 726n.
 Hallwachs, 72

- Hamermesh, **710**
 Hamilton, 32, 238
 Hanstein, 655n.
 Harmonic oscillators, 67, 171-175, **177**,
 205-206, 271
 Harper, 684n.
 Hart, 21n.
 Hartley, 187
 Hartley's law, 187
 Hartree, 296n.
 Hatfield, 559n.
 Haworth, 605, 616n., 619/i.
 Haxby, 631n., 653n., 654n.
 Haxel, 560n.
 Haynes, 665
 Hazen, 689n., 691n., 695, 708, 711, 712,
 723n., 726, 729
 Heat, caloric theory of, 29, 33, 34
 discoveries in, 29, 33-34
 of dissociation, 404, 409
 kinetic theory of, 29, 32
 mechanical equivalent of, 34
 Heath, 5n.
 Heisenberg, 236, 267, 268, 419
 Heitler, 514n., 518n., 600n., 700-703,
 704, 705n.
 Helium, mass and binding energy of, 586
 production of, 557
 spectrum of ionized, 216, 345
 Helmholtz, 34, 50
 Henderson, 608n., 654
 Henry, 45
 Herb, 635n., 636n.
 Hertz, G., 224, 553
 Hertz, H., 50, 71, 72
 Herzberg, 356n., 399n., 409n, 412
 Hess, 660
 Hevesy, 555n.
 High potentials, production of, 605, 610
 High-speed ions, production of, 609
 Hilberry, **718**
 Hjalmar, 522
 Hoag, 718
 Hoffmann, 695
 Holes in subshells, 310, 491
 Hollnagel, 429n.
 Holloway, 559, 572
 Homonuclear molecules, 411, **423**
 Hooke, **28**
 Hoteko, 725
 Houston, 245n.
 Howe, 527n.
 Huber, 631n., 639?i.
 Hudson, C. M., 635n., 636n.
 Hudson, J. C., 485w.
 Hudspeth, 618n.
 Huggins, 188
 Hughes, 92-98n., 580n., (fig.) 692, 717n.,
 724n., 726n.
 Huizinga, 532n.
 Hull, A. W., 508n.
 Hull, G. F., 60
 Hunt, 496n.
 Hurmuzescu, 455
 Huygens, 28, 114, 115
 Hydrochloric acid, specific heat of, 450
 molecular spectrum of, 403, 406, 410
 Hydrogen,-Bohr's theory of, 204-210
 energy levels of, 208, (table) 214, 215,
 (diagram) 219, 325
 heavy, 393, 553
 orthohydrogen, 448
 parahydrogen, 448
 spectrum of, (figs.) 189, 215
 fine structure of, 342-343
 isotope structure of, 393
 Hyperfine structure, 391-396

 |

 /, 393
 Ibsen, 655
 Idei, 495n.
 Ikawa, 651 n.
 Imbert, 457
 Imes, 407
 Impenetrability of molecules, 309
 Indeterminacy principle, 267-269
 Induced emission, 282
 Induction accelerator, 613
 Intensity of electromagnetic waves, 58
 Interference of light, 35
 Invariant (relativistic), 135
 Ionic weight, 74
 Ionization, bursts of, 695
 Ionization of atoms, 220-231, 481
 by charged particles, 688-692, 705
 multiple, **223**
 Ionization chamber, 455

Ionization energy of potential of atoms,
 210, 220, 230, (table) 231, (plot)
 329, (table) 739
 Isobars, 550, 592
 Isomers, 550
 Isoelectronic atoms, 317
 Isothermal enclosure, 145
 Isotope structure, 391, 419
 Isotopes, 546, 548-553, (table) 733
 separation of, 552, 626, 658
 Isotopic weight, 549
 Ives, 124n.

J

Jeans, 160, 164, 372n.
 Jentschke, 653
 Jesse, 722, 725, 728, 730
 Johnson, 108n., 675n., 678n., «681n., 730
 Johnston, 419n.
 Joliot, 576, 600n., 603-605, 652n.
 Jones, 505n.
 Jongen, 664
 Joule, 34

K

K lines or series, 468, 479
k, 163
 Kaufman, 495n.
 Kayser, 186, 189, 336n.
 Keesom, 437n.
 Kellogg, 394n.
 Kelvin, 34
 Kennard, 102n., 164n., 171n., 449n.
 497n.
 Kepler, 17, 59
 Kepler's laws, 18, 19
 Kerr, 45n.
 Kerst, 613, 614n.
 Kikuchi, 256, (fig.) 257
 Kimura, 65In.
 Kinetic theory of gases, 48
 Kingshill, 727
 Kircher, 20, 29
 Kirchhoff's law, 143, 283
 Kirkpatrick, 501n., 519
 Klaiber, 626n.
 Klein, 520n.
 Klemperer, 603n.
 Knipping, 257, 469

Knock-on electrons, 728
 Knox, 36
 Koch, 654n.
 Kock, 638
 Kolhorster, 660, 682, 683
 Korff, 709n., 710
 Kossel, 539n.
 Kramers, 419
 Kretschmar, 510n.
 Krishnan, 420
 Kronig, 534
 Kruger, 600, (fig.) 601, (fig.) 632, 633,
 643
 Kuhn, 449n.
 Kuipers, 532 *n*.
 Kulenkampff, 501
 Kunz, 87n.
 Kurlbaum, 171, 429n.
 Kurie, 609n., 611n.
 Kusaka, 730n.
 Kussman, 183n.

L

L lines or series, 468, 479
 Ladenburg, 87, 183n.
 Lagrange, 28
 Landé, 380
 Landé *g* factor (splitting factor), 379-
 385, (table) 384
 Landers interval rule, 356
 Laplace, 37
 Lapp, 728n.
 Larmor, 372n.
 Larmor precession, 78, 372, 373
 Larsson, 523
 Laue, 257, 468
 Laue spots, 473
 Lauritsen, 603n., 646n., 648-649
 Lawrence, 608n., 611, 612
 Lea, 62In.
 Lead, spectrum of, 365
 X-rays from, 502
 Lebedev, 60
 Lees, 579
 Leibnitz, 28
 Lemaitre, 675, 676
 Lenard, 73, 85-87, 228
 Lenses, discoveries concerning, 20, 22,
 24

- Leprince-Ringuet, 713, 716
 Leu, 391
 Lewis, G. N., 553
 Lewis, L. G., 727
 Lewis, W. M. C., 438w.
 Lienhard, 639n.
 Light, corpuscular theory of, 94, 95
 Light, discoveries in, 29, 34-36
 electromagnetic theory of, 49
 velocity of, 111, 112, 118, 121, 136
 wave theory of, 24, 32-36, 95
 Lightning, 31
 Lrndemann, 430, 432
 Line breadth, 396-399
 Lipperhey, 14
 Little, 621
 Liveing, 187
 Livingood, 625n., 626n.
 Livingston, 559, 560n., 572, 608n, 609n.,
 611, 626n.
 Locher, 686
 Lodge, 14n.
 lx)mbardo, 726, 729
 liong, 621
 Lorentz, 75, 76, 100n., 112, 116, 117
 Lorentz transformation, 120-125
 Lorentz unit, 376
 Loria, 227w.
 Luebke, 605n., 616n., 619n.
 Lummer, 158, 171, 184
 Lyman series, 213, 215
- M
- McClelland, 96
 McDaniel, 623n., 624n.
 MacDonald, 553n.
 Mach, 109n.
 McMillan, 615n.
 Magnetic moment, of atom, 368
 measurement of, 390
 of nuclei, 391, 392, (table) 394
 of one-electron atom, 373
 due to orbital motion, 366
 due to spin, 286, 367
 of earth, 674
 Magnetism, 10, 20, 45, 314, 369
 Magnetron, 374
 (nuclear), 394
 Maltezos, 457
 Mandeville, 621
 Manley, 605n., 616n., 619n.
 Mann, 114n., 612n.
 Marsden, 199, 201, 203
 Mass, atomic unit of, 550
 conservation of, 131
 electromagnetic, 68
 of energy, 130-132, 585
 inertial, definition of, 127
 reduced, 209
 rest, 128
 variation of, 126, **128**
 Mass correction, 583, (fig.) 584
 Mass defect, 583w.
 Mass numbers, 550, (table) 733
 Mass spectrograph, 545
 Masses of atoms, 549, (table) 733
 spectrum of, (photograph) 549
 Mattauch,*396, 552rc., 733
 Matrix component, 284
 Maurer, 655
 Maxwell, 46-49, 50
 Maxwell's equations, 51
 May, 640, 71 In.
 Mayer, J. R., 171n.
 Mayer, M. G., 171n.
 Mayer, R. J., 34
 Mean life, 281
 Mechanics, discoveries in, 20, 27, 28,
 32
 as geometrical optics, 238
 relativistic, 126-132
 Mecke, 419n.
 Meggers, 221, 222
 Meitner, 600w., 651, 652, 654
 Mercury, spectrum of, 357
 Mersenne, 20
 Mesons (mesotrons), as cause of show-
 ers, 698, 730
 discovery of, 712-717
 energy loss by, (plot) 703
 in the atmosphere, 722-726, 728
 ionization by, 691
 mass of, 717
 mean life of, 717-719
 origin of, 720
 showers of, 726, 728
 Metals, electrons in, 100-108
 Metastable levels, 359, 569
 Michelson, 113, 185n., 391

- Michelson-Morley experiment, 113-117, 123, 343
 Micron, 185
 Microscope, electron, 242
 gamma-ray, 268
 Miller, 115n.
 Millikan, 83, 84, 90, 91, 164n., 659n., 660-666, 678, 680, (fig.) 681, 714, 730, 731
 Mitchell, 619n.
 Mohler, 221, 222
 Molecules, average energy of, 161-163
 ¹ number in gram-molecule, 163
 Möller, 583n.
 Moment of inertia of molecules, 402
 Momentum, electromagnetic, 53-54, (waves) 58
 indeterminacy of, 266-269
 relativistic, 128, 130
 Montgomery, 699n.
 Morgan, 698n.
 Morley, 113
 Moseley, 479, 480
 Moseley's law, 479, 482
 Mott-Smith, 686
 Multiple quantum states, 232
 Multiplets, *jj*, 364
 of alkalis, 335
 LS or Russell-Saunders, 354-357
 of one-electron atoms, 339-345
 of spectral lines, 355
 Murphy, 393n.
- N
- n, 579
 Nahring, 527
 Nassar, 708n.
 Neddermeyer, 693, (fig.) 694, 714, (fig.) 715
 Nedzel, 655
 Neher, 665, 678, 680, (fig.) 681, 725n., 730, 731
 Neon spectrum, 224
 Neptunium, 633, 658
 Nereson, 697n., 709n., 719n.
 Nernst, 432, 437n., 439
 Neueft, 629
 Neutrino, 571, 646
 Neutrons, 615
 capture of, 619, 620-624
 counting of, 616
 discovery of, 576
 forces on, 590-594
 magnetic moment of, 620
 mass of, 551, 580
 mean free path of, 619
 radioactivity of, 644
 scattering of, 620
 slow and thermal, 618, 632
 sources of, 615
 velocity of, 617
 (See also Nuclear reactions)
 Newton, 21-28, 30, 36, 109
 Nichols, 60, 429n., 430
 Nicholson, 37
 Nielsen, 698n.
 Nier, 550n., 653n.
 Nishina, 520n., 651n.
 Normal state, 208
 Normalization to unity, 26G
 Northrup, 61 On.
 Nuclear reactions, 575-649
 alternative bombardments, 627
 products, 626, 628, 646
 determination of mass by, 629
 due to, a-rays, 575, 579, 604, 616, 626, 635, 637
 deuterons, 616, 618, 626, 627, 629, 631, 635, 649
 electron capture, 644
 electron impact, 637-638
 neutrons, 616, 618, 619, 626, 627, 631, 632
 photons (photodisintegration), 627, 637
 protons, 606, 608, 616, 624, 626, 629, 631, 635, 636, 637, 646, (resonances) 634, 647
 emission of a-rays, 606, 608, 626, 629, 631, 632, 639, (spectra) 641
 of r-rays, 619, 626, 632, 633-637, 644-647
 of neutrons, 579, 604, 616, 618, 626, 627, 629, 631, 632, (spectra) 641, 650
 (See also Neutrons)
 of protons, 575, 626, 627, 631, 635, 639
 endothermic or exothermic, 630

Nuclear reactions, energy relations,
 575, 606-609, 630, 636, 639-640
 excitation of, 629, 647
 identification of initial isotope, 626,
 627
 identification of product, 604
 isotopic product, 632
 nature of, 624-625
 notation for, 575, 625, 626
 resonance levels for, 639-641, 646-
 649
 with radioactive products, 625-628,
 645, 647
 Ag^{106i} , 627
 Al^{28} , 626
 As^{74} , 628
 Au^{196} , 632
 Co^{68} , 645
 Mg^{27} , 626
 Mn^{64} , 645
 N^{13} , 604, 636
 Na^{22} , 628
 Na^{24} , 626, 628, 632, 633
 Np^{239} , 632
 O^{16} , 648
 P^{30} , 604, 626
 Pu^{239} , 632, 633
 S^{31} , 632
 U^{239} , 632
 $Z = (43)$; $Z = (93)$; $Z - (94)$, 633
 types of, 625
 $\text{A}^{27} + a \rightarrow \text{p}^{30} + n$, 604
 $\text{B}'' + \text{H}^1 \rightsquigarrow \text{Be}^8 + ct \text{ or } 3a$, 608
 $\text{Bio} + {}_n \rightarrow \text{Li}^7 + {}_a i$, 616
 $\text{B}^{10} + a \rightsquigarrow \text{N}^{13} + n$, 604
 $\text{Be}^9 + \text{H}^1 \rightarrow \text{Be}^8 - f \text{H}^2$, 624
 $\text{Be}^9 + \text{H}^2 \rightarrow \text{B}^{10} + n$, 641
 $\text{Be}^9 + a \rightarrow \text{C}^{12} + n$, 579
 $\text{Be}^9 - f a \rightarrow \text{N}^{14} + n$, 579
 $\text{F}^{19} + \text{H}^1 \rightarrow \text{O}^{16} + a$, 608, 648
 $\text{F}^{19} + \text{H}^2 \rightarrow \text{Ne}^{20} + n$, 634
 $\text{F}^{19} + a \rightarrow \text{Na}^{22} - f n$, 639
 $\text{F}^{19} - f ct \rightarrow \text{Ne}^{22} - 4\text{H}^1$, 639
 $\text{Hi} + n \rightarrow \text{H}^2$, 619
 $\text{H}^2 - 4\text{H}^2 \rightarrow \text{He}^3 - 4n$, 616
 $\text{Li}^7 - 4\text{H}^1 \rightarrow 2\text{He}^4$, 606, 635
 $\text{Li}^6 - 4\text{H}^2 \rightarrow \text{Li}^7 - 4\text{HS}$, 635
 $\text{Li}^6 - 4n \rightarrow \text{H}^* - 4$, 616
 $\text{N}^{14} + a \rightarrow \text{O}^{17} - 4\text{H}^1$, 575
 Nucleon, 583

Nucleus, 199-200
 angular momentum (or spin), 392,
 393-394
 binding energy, 586-588
 constituents of, 581
 density of, 589
 energy and mass of, 586-588
 energy from, 575, 656
 energy levels, 567, 597, 640, 644
 metastable, 569
 of B^{10} , 642
 of Mg^{24} , 633
 of Ne^{22} , 641
 fission of, 650-658
 isomeric, 627, 633
 magnetic moment of, 391, 392,
 (table) 394
 mass and rest energy of, 583-588
 motion of, 208, 280
 quantum states of, 590
 radius of, 589
 structure of, 581-594
 theory of, 588-598, 651
 transformation of, 561-564, 573-576,
 603-606
 (See also Nuclear reactions)

O

Occhialini, 600n., 686
 Oersted, 38
 Ogle, 600, (fig) 601, (fig) 632, 633
 Oliphant, 633n.
 Oppenheimer, 705n.
 Orbits in one-electron atom, 208-210,
 233, 280
 Oscillators, harmonic, energy levels of,
 271
 quantum theory of, 205-206
 radiation by, 67
 in thermal equilibrium, 171-175, 177
 wave-mechanical theory of, 271
 Otis, 662n.

P

P, 349
 p, 291
 Packing fractions, 584, (table) 733
 Page, 53n.
 Pair production, 600, 602, 633, 700, 705

- Paneth, 555n.
 Parratt, (fig.) 482, 494n., 531n., B32n., 536n., (fig.) 538, 539n.
 Particles, 264-267, 268-269
 in nucleus, 581, 588, 590
 Pascal, 20
 Paschen, 345, 377, 387
 Paschen-Back effect, 386-388
 Paschen series, 213, 215
 Pauli, 290, 392
 Pauling, 271n., 31 In.
 Pegram, 619n.
 Peierls, 588n.
 Pendulum, 14, 28
 Peregrinus, 10
 Period (of decay), 561
 Periodic system, 298-316
 Bohr's table for, 300
 standard table for, 738
 Perlow, 630
 Perot, 185n.
 Perrin, 80, 601
 Perturbation theory, 294
 zero-order, approximation, 296-298, 347, 348
 Petrzhak, 655n.
 Pfozter, 725n.
 Philipp, 600n.
 Photoelectric effect, 71-108
 and Compton effect, 520, 602
 discovery of, 71
 due to γ -rays, 565-567, 570
 due to X-rays, 506, 520
 nature of, 72-73, 85
 variation with frequency, 90
 with intensity, 87
 with potential, 86
 various properties of, 91, 106
 Photoelectrons, energy of, 87, 90
 source of, 92-94, 105-107
 Photon, 95, 237, 285
 Picard, 27
 Pickavance, 711n.
 Pickering, 710n., 725n., 730, 731
 Pierce, 505n.
 Placzek, 709rc.
 Plain, 635n., 636n.
 Planck, 13ft, 170, 175-180, 183, 184, 204
 Planck's constant h , 91, 102, 175, 212
 742
 Planck's radiation law, 178-183
 Plutonium, 629, 633, 658
 Pohl, 457
 Poisson, 37
 Polarization of light, 35, 36
 Pollard, 577n., 733
 Pollock, 615n.
 Pontecorvo, 618n.
 Pool, 627n.
 Pose, 640, 641, 655
 Positive rays, 542
 Positrons, 555, 598-603
 Potential barrier (nuclear), 595
 Powell, 71 In., 723n.
 Poynting's vector, 54, 58
 Preston, 160
 Preston's rule, 382-383
 Pringsheim, 158, 171, 184
 Probability amplitude, 266
 Probability density, 266, 277
 Probability of transition, 284
 Protons, energy, velocity, and H_p for, (plot) 688
 energy loss by, (plot) 702
 forces on, 590-594
 ionization by, 691
 mass of, 551
 range of, 560
 scattering of, 593
 (See also Nuclear reactions)
 Prout, 195
 Ptolemy, 9
 Pyrkosch, 659n.
 Pythagoras, 5
- Q**
- Q (nuclear heat of reaction), 606, 609, 630
 Quadrupole X-ray lines, 495
 Quantum defect, 328, 329
 Quantum numbers, for atoms, 272-273, 288, 387
 F, M_f , 394
 $n/l/n$, 272, 290, 296
 n/ljm , 332
 /, M_i , (nucleus) 393
 J M, 319, 351, 387
 LKXJIM, 322, 349, 351
 s n, 287

- Quantum numbers, *8 p d . . .*, 280
S P D . . ., 322
 of electrons in atoms, 289, 347, 362
 effective principal, 328
 in *K* and *L* shells, 292
 for molecules: (*J*) 401, (*v*) 404, (*A*) 413, (*S, K*) 413, (*I, M, tt*) 414
 of X-ray levels, 490
- Quantum of energy, 91, 176, 177
- Quantum states, 175, 269, 288, 293
 in central field, 289, 332-335
 due to spin, 286
 in external field, 365
 of harmonic oscillator, 173
 of ions (X-ray), 481
 in magnetic field, 373, 377, 379, 387
 of many-electron atom, 293
 of molecules, 411-414
 diatomic, 401, 404, 405
 multiple, 232
 normal, 208
 nuclear, 596
 of one-electron atom, 207
- Quantum theory, 170, 175, 204-207, 233
 (See also Wave mechanics)
- It
- R, R* (see Rydberg constant)
- Rabi, 391, 394c.
- Radiation, absorption and emission of,
 176, 177, 280-283
 by charged particles, 60-68, 702-704
 dipole, 285
 by harmonic oscillator, 67
 quadrupole, 285, 495
 density of, 143, 146
 in electromagnetic field, 53
 intensity of, 58
 nature of, 528
 pressure due to, 59, 147-149, 183
 resonance, 225
 from a surface, 141, 143-144
 thermal, 139
 (See also Black-body radiation)
- Radioactive series, 562
- Radioactivity, 553-567
 definition of, 624-625
 discovery of, 553
- Radioactivity, explanation of, 591,
 594-596
 induced, 603-605
- Raman, 420
- Raman effect, 419-423
- Ramberg, 533w., 536w.
- Rasetti, 555n., 557n., 571 n., 588n., 615n., 618n., 619n., 659n.
- Rawlinson, 507n.
- Rayleigh, 160, 164, 397, 419
- Rayleigh (R. J. Strutt), 226n.
- Rayleigh-Jeans formula, 160, 169
- Reflection from moving mirror, 152
- Reflectivity, 143
- Refraction, 9, 10, 20, 22
 of X-rays, 457, 522-527
- Regener, 662, 664
- Reich, 559n.
- Reimann, 96n., 544n.
- Relativity, 109-138
 and electromagnetism, 132-135
 mechanics, 126-132
 propagation of light, 111
 wave mechanics, 286
 general theory of, 135
 Newtonian, 109-111
 special theory of, 117-135
- Residual rays, 429
- Resonance potentials, 230
- Resonance radiation, 225
- Rest energy, 130, 585
- Richardson, 88, 89, 97, 344*.
- Richardson's equation, 97
- Richtmyer, 87n., 504n., 505n., 531n., 533n., 536w.
- Right-handed axes, 51
- Ritz, 193
- Roberts, 636n.
- Robinson, 507, 509
- Romer, 30
- Roentgen, 451-453, 466
- Rogozinski, 72Sw.
- Rojansky, 271n.
- Rosenberger, 5n.
- Ross, 517
- Rossi, 684, 685, 696, 718, 719n., 729
- Royal Institution, 33, 40
- Ruark, 233n., 372n., 399n.
- Rubens, 171, 429, 430, 439
- Rumbaugh, 636n.

- Rumford, 33
 Runge, 189
 Rutherford, 184, 197, 199-201, 203, 555n., 556, 563n., 572n., 574, 575, 578, 689n.
 Rydberg, 189, 190-193
 Rydberg constant, **192**, 211, 393, 742
 for one-electron atom, 216
 Rydberg formula, 192, 194, 328, 356
 Rydberg-Schuster law, 103, 219
- S
- s, 340
 s, 291
 Sadler, 450n., 465
 Saha, 641
 Savart, 38
 Saxena, 725
 Scattering of light, 419
 multiple and single, 100
 (See also under object scattered)
 Schein, 695, 722, 725, 728, 730
 Scheiner, 20
 Scherrer, 639n.
 Schrodinger, 236, 237, 259, 269, 436, 438n.
 Schuster, 193
 Screening of nucleus, 483
 Seaborg, 625n., 626n., 733
 Segre, 618n., 619n., 654n.
 Seitz, 108n.
 Selection rules, in atoms, 285, 318
 j, m, 333, 363
 J M, 320
 Z, 322, 333, 363
 L A S 2, 322, 350, 351, 387
 for configurations, 347, 353
 for molecules, J, 402, 406
 K, 414, 415
 AS, 413
 for Raman lines, 421
 for X-ray lines, 493, 495
 Serber, 614n., 707
 Series (spectral), discovery of, 186
 fundamental, 324
 interrelations between, 189, 193-195
 principal, sharp, diffuse, 190, 218-219, **323-325, 336**
relativistic terms, **194**
X-ray
- Shapiro, 711, 712
 Shaw, 525n.
 Shdanow, 712
 Shells and subshells, 291-292, (tables) 315, 490, 739
 Shimizu, 572
 Shockley, 108n.
 Shortley, 396
 Shoup, 631n., 653n., 654n.
 Shutt, 726
 Siegbahn, 476, 504n., 535n.
 Silberstein, 122n.
 Simon, 518
 Simultaneity, 118-120
 Skinner, 540w.
 Skobelzyn, 682, 602
 Slater, 311n.
 Smekal, 420
 Smith, 637n.
 Smythe, H. D., 552n., 655n., 656n., 658
 Smythe, W. R., 53n., 61n.
 Snell, 20, 655
 Snyder, 705n.
 Soddy, 555n.
 Sodium, energy levels, (diagrams) 210, 325, 326-329, (table) 337
 quantum defects, 328
 spectrum, 190, 218, 325, (table) 337
 excitation of, 225-226
 Zeeman effect for, 385
 Sommerfeld, 101, 233, 234, 340n., 345, 498
 Sound, discoveries in, 20
 Space quantization, 276, 380
 Space-time, 110
 Specific heats, 424-450
 of gases, classical theory of, 439-442
 as observed, 447
 quantum theory of, 442-447
 of solids, 424-438
 classical theory of, 426
 Debye's theory of, 432-438
 Einstein's theory of, 427
 of mixed solids, 438
 variation with temperature, 424-426, 435-438
 Spectra, absorption, 224
 alkali type of, 321-329, 335-338
 arc and spark, 187, 223
 continuous, **212, 224**

Spectra, from ionized atoms, 317
 from many-electron atoms, 316-318
 molecular, 399
 (See also Bands)
 from one-electron atoms, 210-216,
 342-345
 two-electron type, 357-358
 (See also Series, X-ray)
 Spectral lines, breadth of, 396-399,
 536
 Spin, electron, 286
 special correction due to, 340
 neutron and proton, 594
 nuclear, 393, (table) 394
 Spinks, 409w.
 Spin-orbit effect, 294
 in central field, 329
 on *LS* terms, 354
 on magnetic levels, 388
 in molecules, 414
 in one-electron atom, 340
 Stanton, 729
 Starling, 53n.
 Stark, 388
 Stark effect, 388
 Stationary states, 205, 269, 288, 293
 Statistical weights, 232
 Stearns, 678, (fig.) 685, 696n.
 Stebbins, 87n.
 Stefan constant, 742
 Stefan-Boltzmann law, 149-152
 Stenstrom, 522
 Stephens, 631n., 653n., 654n.
 Stern, 258, 389
 Stern-Gerlach experiment, 389
 Stevenson, 717
 Stilwell, 124w.
 Stormer, 675
 Stokes' law, 83
 Stoletow, 72
 Stoney, 74, 456
 Strassmann, 650, 652
 Straubel, 465
 Street, 717
 Streib, 647, 649
 Strutt, 226n.
 Sundarachar, 644n.
 Swann, 730w.
 Synchrotron, 615
 Szilard, 655n., 657

T

Tabin, 721
 Tarrant, 603
 Taylor, 391
 Telescope, 14, 15, 24
 Terms (spectral), 193-194, 214
 of alkali metals, 323, 326-329
 doublet, 335
 3*h* 363
 LS, 352
 singlet and triplet, 357-358
 Thales, 5
 Thaxton, 593n.
 Thermionic emission, 96-99
 origin of thermions, 103
 Thermodynamics, 33, 34
 Thermometer, invention of, 29
 Thibaud, 527n.
 Thompson, S. P., 38n.
 Thomson, G. P., 104n., (fig.) 282
 Thomson, J. J., 68, 96, 195, 543
 on electrons, 73, 80, 83, 85
 on X-rays, 195, 196-200, 453, 458
 Thornton, 627n., 652n.
 Time (in relativity), 109, 118-124
 Tolman, 174n.
 Tombouliau, 54In.
 Tomlinson, 648
 Torricelli, 20
 Townsend, 83
 Tracks, (photograph) 573
 Transitions, radiation less, 533
 Transmissivity, 143
 Transmutation of atoms, 561-564, 573-
 576, 605, 624-658 (see Nuclear
 reactions)
 Trautz, 450n.
 Trischka, 539w.
 Turner, 650n., 664n., 666, 678, 679
 Tuve, 684n.

U

Uhlenbeck, 286
 Ulrey, 496n.
 Uncertainty principle (see Indeter-
 minacy principle)
 Uranium, fission of, 652-658
 X-ray diagram for, 492
 Urey, 233n., 372n., 399n., 553n.

V

Vaidyanathan, 640
 Valence bonds, 308
 Vallarta, 675, 676
 Van Allen, 637rc.
 Van Atta, 610n.
 Van de Graaff, 610
 Van de Graaff generator, 609
 Van den Ende, 437n.
 Vector diagrams, 275
 for $J M$, 319
 for $L S J$, 361
 for Zeeman effect, 375, 378, 381
 Veksler, 615w.
 Velocities, transformation of, 124
 Villari, 454
 Vinci, Leonardo da, 10
 Viscosity of gases, 48
 Vogt, 485n.
 Volta, 37
 Volz, 620n., 621n.
 von Baeyer, 569n.
 von Wartenberg, 429n.

W

Waffler, 639n., 729n.
 Waldman, 638
 Walter, 457
 Walton, 605-608
 Warburton, 505n.
 Warren, 730n.
 Wave equation, 259-263, 286, 293
 for harmonic oscillator, 271
 for a many-electron atom, 293-294
 for a one-electron atom, 272, 278
 for a particle, 263, 270
 Wave functions, 270, 286
 of harmonic oscillator, 271
 of one electron, in complex atom, 296, 334
 of one-electron atom, 272, 278
 Wave length, de Broglie (of matter waves), 246-247
 measurement of, 35, 185, 476, 527
 Wave mechanics, of many-electron systems, 293-298, 335-349, 353
 of one particle, 334
 relativistic, **286**

Wave number, ν , 185, 214
 in terms of electron-volts, 222, 742
 Wave packets, 265, 596
 Waves, electromagnetic, 55-60
 matter (electron), 235-259
 diffraction of, 248-259, (photographs) 257-258
 molecular, 257
 physical significance of, 264-267
 refraction of, 239-244, 253-256
 wave length of, 244-248
 phase, 237
 Weber, 425
 Weighted average energy level, 356
 Weizel, 399n., 41 In.
 Wells, 609n., 631n., 653n., 654n.
 Wentzel, 532n.
 Westendorp, 614n.
 White, A. D., 16n.
 White, C. F., 361n., 365w.
 White, H. E., (fig.) 329, (fig.) 366, 384, (fig.) 392, 397n.
 Weidenbeck, 621w., 634, 635, 638n.
 Wien, 157
 Wien displacement law, 157
 Wien's formula, 159
 Williams, R. C., 343
 Williams, W. E., 344n.
 Wilson, A. H., 108n.
 Wilson, C. T. R., 518, 572n., 659, 689
 Wilson, E. B., Jr., 271n.
 Wilson, R. S., 632n.
 Wilson, V. C., 724n.
 Wilson, W., 689
 Wind, 457
 Winkelmann, 465
 Wollan, 664n., 722, 725, 728, 730
 Wood, 225, 245n., 422, 429n., 523n.
 Wooster, 569
 Work function, 97, 98, 104
 Wulff, 396
 Wynn-Williams, 571 n.

X

X-rays, 451-541
 absorption of, 464-468, 501-512
 absorption lines, 484-486, 536
 absorption limits, (table) 484, 486, 488, (table) **510**, 537, **539**

- X-rays, Auger effect, 533
 characteristic, 465
 continuous spectrum, 495-501
 definition of, 615
 diffraction of, 457; 468-473
 discovery and properties of, 451
 efficiency of production of, 501
 emission of, 456, 457
 energy levels for, 487-494, 532-
 (resonance) 537
 spectroscopic notation for, 494
 excitation of, 486
 first-order or diagram lines, 483, 531
 fluorescence, 466-468, 487
 frequencies, 480, (tables) 484-485
 hardness of, 465
 homogeneous, 465
 intensity of, continuous spectrum,
 495, 499
 lines, 494
 measurement of, 454, 499
 involving periphery of atom, 536-539
 from multiple ionized atoms, 531-533
 and outer part of the atom, 536
 photoelectric effect of, 506, 520
 polarization of, 462
 pulse theory of, 455-456, 458-462
 quadrupole lines, 495
 quantum theory of, 489-495
 reflection of, total, 525
 refraction of, 457, 522
 scattering of, 457, 462-464, 510-522,
 (plot) 512
 classical (Thomson) theory of, 458
 Compton, 514-522
- X-rays, second-order (or satellite) lines,
 531
 from solids, 539
 spectra, 468, 477-486, 536
 origin of, 481-486, 497
 plots of, 482
 selection rules for, 493, 495
 spectrometer for, 473
 target thickness, effect of, 498
 tubes for, 453
 wave lengths, 457, (table) 484
 measurement of, 473-477, 527
- X-unit (X.U.), 185
- Y
- Yalow, 646n.
 Yasaki, 651 n.
 Young, 34, 48, 95
 Yukawa, 644, 721
- Z
- Zacharias, 394n.
 Zeeman, 75, 76, 79
 Zeeman effect, 75-79, 80, 83, 365-388
 anomalous, 365
 classical theory of, 76-79, 369-373
 in huge field, 373-378
 nfjrinS-triplet, 365, 376, 383
 photographs of, 366
 in strong field, 386
 in weak field, 378-386
 Zinc, spectrum of, 187
 Zinn, 655n.

