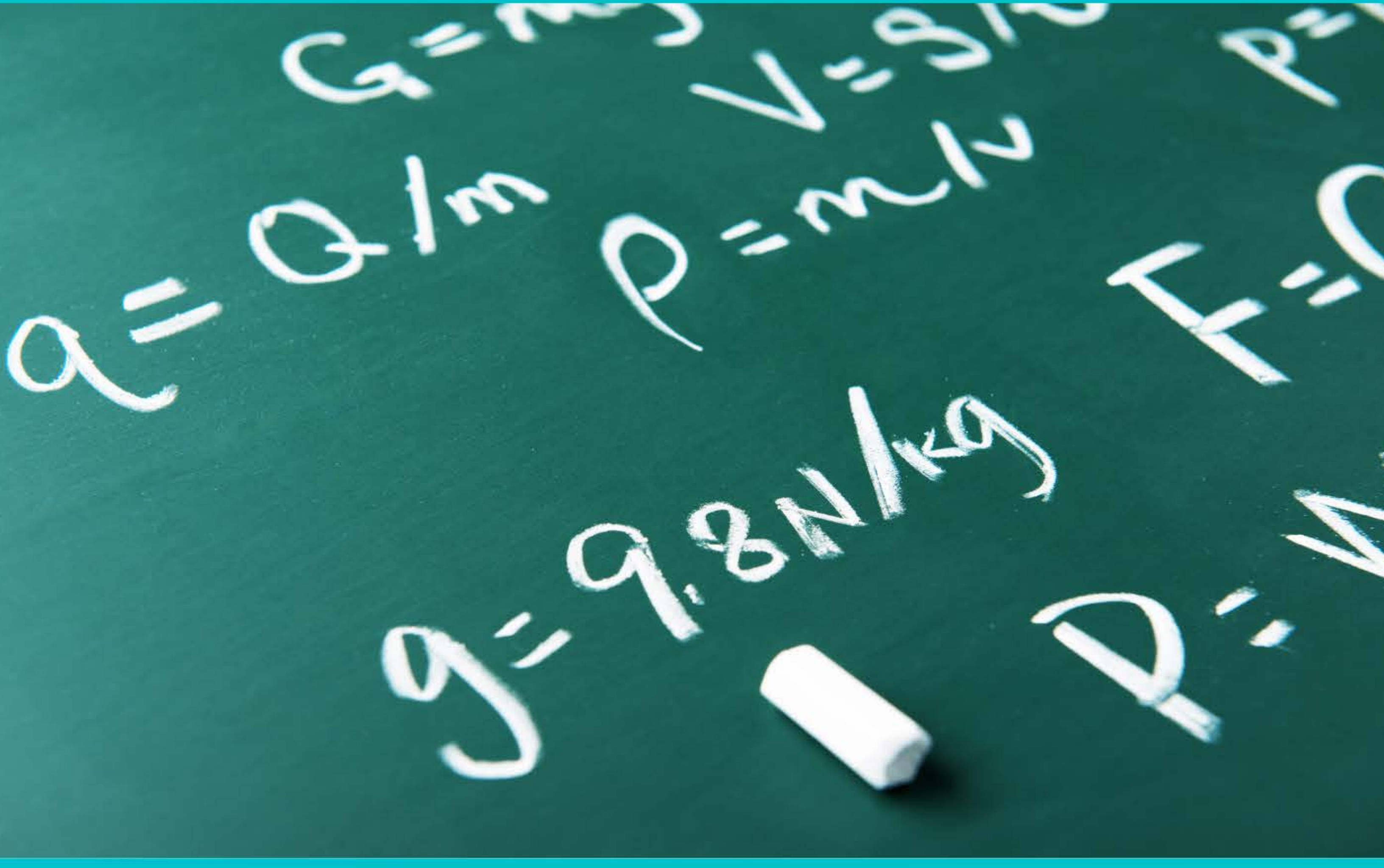
PHYSICS

For

Senior Secondary School

3



EDUBASE

SSS 3
FIRST TERM NOTES ON PHYSICS

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FIRST TERM

WEEK 1

Physics SS 3

Topic: **ELECTROMAGNETIC WAVES,**

Content

- **TYPES OF RADIATION IN ELECTROMAGNETIC SPECTRUM,**
- **DESCRIPTION AND USES**

We can distinguish between two general classes of waves: mechanical waves and electromagnetic waves.

Mechanical waves are waves that require material medium for their propagation.

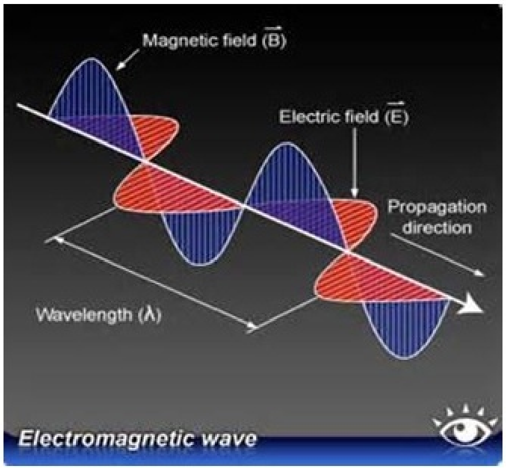
Examples of mechanical waves are water waves, sound waves, and waves on a rope. Water waves are due to vibrations of particles of water. Sound waves are due to vibrations of air particles. Water and air are the material media of propagation.

Electromagnetic waves are those that do not have a material medium for their propagation. They arise from the vibrations electric (E) and magnetic (M) fields. The combination of the electric and magnetic field waves is called an Electromagnetic (E-M) wave.

The magnetic and electric fields of an electromagnetic wave are perpendicular to each other and to the direction of the wave, examples of E-M waves are light wave, X-rays and gamma rays.

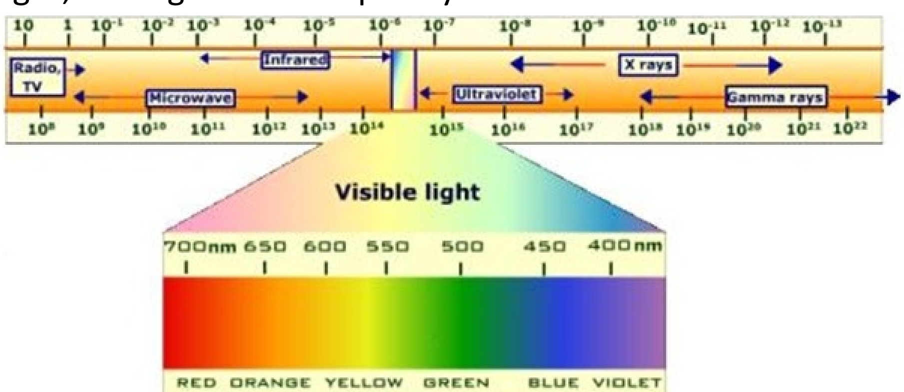
Another difference between mechanical and electromagnetic waves is in their velocities. Electromagnetic waves travel at the speed of light but mechanical waves travel at a speed less than that of light.

A mechanical wave may be transverse or longitudinal, but an electromagnetic wave is always transverse.



Wavelength and Frequency

Just like an ocean wave, an electromagnetic wave has peaks and troughs. The wavelength is the distance between two identical points of the wave from cycle to cycle, for instance, the distance between one peak, or crest, and the next. EMR can also be defined in terms of its frequency, which is the number of crests that pass by in a given time interval. All forms of EMR travel at the same speed: the speed of light. Therefore, the frequency depends entirely on the wavelength: the shorter the wavelength, the higher the frequency.



Types of Radiation

Electromagnetic waves were first predicted theoretically by the British Physicist, James Clerk Maxwell. He also proposed that light itself was electromagnetic radiation. The experimental evidence for the existence of E-M wave was provided by the German Physicist, Heinrich Hertz.

The term electromagnetic wave describes the way electromagnetic radiation (EMR) moves through space. Different forms of EMR are distinguished by their wavelengths, which vary from many yards (meters) to a distance smaller than the diameter of an atomic nucleus. The full range, in increasing order of wavelength, goes from **gamma rays, X-rays, ultraviolet rays, visible light, infrared rays, microwaves** and **radio waves** and is known as the electromagnetic spectrum. Electromagnetic waves have many applications, both in science and in everyday life. The different rays that constitute the spectrum all have the same basic property that they travel with the speed of light. In vacuum this speed is 3×10^8 ms^{-1} and is represented by the letter c. They also all exhibit the properties of reflection, refraction and interference diffraction.

They however differ in their wavelength (and frequency). Recall that wavelength (λ) and frequency (*f*) are connected by the relation

$\lambda = v/f$ or $v = fv$

Here $v = c$, the velocity of light. Hence we can write

$\lambda = c/v$ or $c = f$

The waves also differ in the way they interact with the matter. This way, which depends on their frequency, influences how we can detect each type of wave. Hence the waves in the various parts of the electromagnetic spectrum differ in the means of production and detection.

Uses of Electromagnetic Waves

Cosmic Rays – Hazard

Gamma Rays – Killing Cancer Cells

X-Rays – Medical Uses (X-Raying People)

Ultra Violet – Sterilizing and sun tanning

Visible Light – Violet light has a higher frequency and lower wavelength than red light. (Light we see, photography is a use).

Infra-Red Light – Night Vision, short distance communication and heating

Microwaves – Cooking, communication

Radio Waves – Include short wave radio, medium wave and long wave radio waves.

Radio: Yes, this is the same kind of energy that radio stations emit into the air for your boom box to capture and turn into your favorite Mozart, Madonna, or Justin Timberlake tunes. But radio waves are also emitted by other things … such as stars and gases in space. You may not be able to dance to what these objects emit, but you can use it to learn what they are made of.

Microwaves: They will cook your popcorn in just a few minutes! Microwaves in space are used by astronomers to learn about the structure of nearby galaxies, and our own Milky Way!

Infrared: Our skin emits infrared light, which is why we can be seen in the dark by someone using night vision goggles. In space, IR light maps the dust between stars.

Visible: Yes, this is the part that our eyes see. Visible radiation is emitted by everything from fireflies to light bulbs to stars … also by fast-moving particles hitting other particles.

Ultraviolet: We know that the Sun is a source of ultraviolet (or UV) radiation, because it is the UV rays that cause our skin to burn! Stars and other "hot" objects in space emit UV radiation.

X-rays: Your doctor uses them to look at your bones and your dentist to look at your teeth. Hot gases in the Universe also emit X-rays.

Gamma-rays: Radioactive materials (some natural and others made by man in things like nuclear power plants) can emit gamma-rays. Big particle accelerators that scientists use to help them understand what matter is made of can sometimes generate gamma-rays. But the biggest gamma-ray generator of all is the Universe! It makes gamma radiation in all kinds of ways.

The following puts waves in order of increasing wavelength and decreasing frequency.

Questions

1. The electromagnetic waves with the highest frequencies are called

A. Gamma rays B. X-rays C. Ultraviolet rays D. Infrared rays

2. In a vacuum all electromagnetic waves have the same what?

A. wavelengths B. Frequency C. Propagation Speed D. Characteristics

3. A certain radio station is assigned a frequency of 2000KHz. Estimate the wavelength of its radio wave.

A. 210m B. 150 m C. 201 m D. 170 m

4. Which of the following is not an electromagnetic radiation?

A. X-ray B. Radio waves C. Infrared radiations D. Sound waves.

5. In which of the following groups are the radiations arranged in the increasing order of their wavelength?

A. Radio waves, gamma rays, X-rays

B. X-rays, gamma rays, radio waves

C. Gamma rays, radio waves, X-rays

D. Gamma rays, X-rays, radio waves

Answer

1. A 2. C 3. B 4. D 5. D

WEEK 2

Topic: ELECTROMAGNETIC INDUCTION

INTRODUCTION

The term electromagnetic induction refers to the generation of an electric current by passing a metal wire through a magnetic field. The discovery of electromagnetic induction in 1831 was preceded a decade earlier by a related discovery by Danish physicist Hans Christian Oersted (1777–1851). Oersted showed that an electric current produces a magnetic field. That is, if you place a simple magnetic compass near any of the electrical wires in your home that are carrying a current, you can detect a magnetic field around the wires. If an electric current can produce a magnetic field, physicists reasoned, perhaps the reverse effect could be observed as well. So they set out to generate an electric current from a magnetic field.

That effect was first observed in 1831 by English physicist Michael Faraday (1791–1867) and shortly thereafter by American physicist Joseph Henry (1797–1878). The principle on which the Faraday-Henry discovery is based is shown in the figure on page 762. A long piece of metal wire is wound around a metal bar. The two ends of the wire are connected to a galvanometer, an instrument used to measure electric current. The bar is then placed between the poles of a magnet.

Electric current: A flow of electrons.

Electrical generator: A device for converting mechanical (kinetic) energy into electrical energy.

Galvanometer: An instrument used to measure the flow of electric current.

Potential difference: Also called voltage; the amount of electric energy stored in a mass of electric charges compared to the energy stored in some other mass of charges.

Transformer: A device that transfers electric energy from one circuit to another circuit with different characteristics.

As long as the bar remains at rest, nothing happens. No current is generated. But moving the bar in one direction or another produces a current that can be read on the galvanometer. When the bar is moved downward, current flows in one direction through the metal wire. When the bar is moved upward, current flows in the opposite direction through the wire. The amount of current that flows is proportional to the speed with which the wire moves through the magnetic field. When the wire moves faster, a larger current is produced. When it moves more slowly, a smaller current is produced.

Actually, it is not necessary to move the wire in order to produce the electric current. One could just as well hold the wire still and move the magnetic poles. All that is necessary is the creation of some relative motion of the wire and the magnetic field. When that happens, an electric current is generated.

Applications

Many electrical devices operate on the principle of electromagnetic induction. Perhaps the most important of these is an electrical generator. An electrical generator is a device for converting kinetic energy (the energy of an object due to its motion) into electrical energy. In a generator, a wire coil is placed between the poles of a magnet and caused to spin at a high rate of speed. One way to make the coil spin is to attach it to a turbine powered by water, as in a dam. Steam from a boiler can also be used to make the coil spin.

As the coil spins between the poles of the magnet, an electric current is generated. That current then can be sent out along transmission lines to homes, office buildings, factories, and other consumers of electric power.

Induced Current

This involves generating a voltage by changing the magnetic field that passes through a coil of wire.

First, connect a coil of wire to a galvanometer, which is just a very sensitive device we can use to measure current in the coil. There is no battery or power supply, so no current should flow. Now bring a magnet close to the coil. You should notice two things:

If the magnet is held stationary near, or even inside, the coil, no current will flow through the coil.

If the magnet is moved, the galvanometer needle will deflect, showing that current is flowing through the coil. When the magnet is moved one way (say, into the coil), the needle deflects one way; when the magnet is moved the other way (say, out of the coil), the needle deflects the other way. Not only can a moving magnet cause a current to flow in the coil, the direction of the current depends on how the magnet is moved.

How can this be explained? It seems like a constant magnetic field does nothing to the coil, while a changing field causes a current to flow.

To confirm this, the magnet can be replaced with a second coil, and a current can be set up in this coil by connecting it to a battery. The second coil acts just like a bar magnet. When this coil is placed next to the first one, which is still connected to the galvanometer, nothing happens when a steady current passes through the second coil. When the current in the second coil is switched on or off, or changed in any way, however, the galvanometer responds, indicating that a current is flowing in the first coil.

You also notice one more thing. If you squeeze the first coil, changing its area, while it's sitting near a stationary magnet, the galvanometer needle moves, indicating that current is flowing through the coil.

What you can conclude from all these observations is that a changing magnetic field will produce a voltage in a coil, causing a current to flow. To be completely accurate, if the magnetic flux through a coil is changed, a voltage will be produced. This voltage is known as the **induced e.m.f.**

The magnetic flux is a measure of the number of magnetic field lines passing through an area. If a loop of wire with an area A is in a magnetic field B, the magnetic flux is given by:

$\phi = BA \cos\phi$, where ϕ is the angle between the magnetic field B and vector A, which is perpendicular to the plane of the loop.

If the flux changes, an emf will be induced. There are therefore three ways an e.m.f. can be induced in a loop:

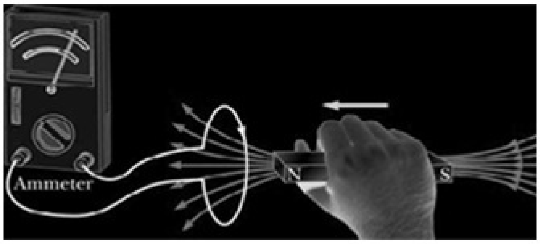
Change the magnetic field

Change the area of the loop

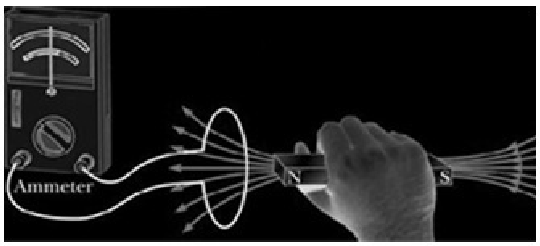
Change the angle between the field and the loop

Factors affecting the magnitude of the induced E.M.F.:

When a magnet is pushed into a coil as shown, the galvanometer deflects in one direction momentarily.



When the magnet is not moving, the galvanometer shows no reading.

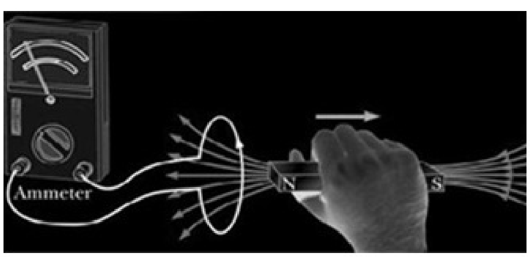


When the magnet is withdrawn from the coil, the galvanometer deflects in the opposite direction

When a magnet is pushed into a coil as shown, the galvanometer deflects in one direction momentarily.

When the magnet is not moving, the galvanometer shows no reading.

When the magnet is withdrawn from the coil, the galvanometer deflects in the opposite direction momentarily.



When the magnet is moved, its field lines are being "cut" by the coil. This generates an induced EMF in the coil that produces an induced current that flows in the coil, causing the deflection in the ammeter.

The magnitude of the deflection depends on the magnetic field density B, the speed of motion v of the magnet, and the number of turns N in the coil.

Faraday's law of induction

An e.m.f. can be induced in a coil if the magnetic flux through the coil is changed. It also makes a difference how fast the change is; a quick change induces more e.m.f. than a gradual change. This is summarized in Faraday's law of induction. The induced e.m.f. in a coil of N loops produced by a change in flux in a certain time interval is given by:

Faraday's law of Induction: $\varepsilon = -N\Delta\phi/\Delta t$

Recalling that the flux through a loop of area A is given by

$\phi = BA\cos\phi$,

Faraday's law can be written:

$\varepsilon = -N\Delta(BA\cos\phi)/\Delta t$

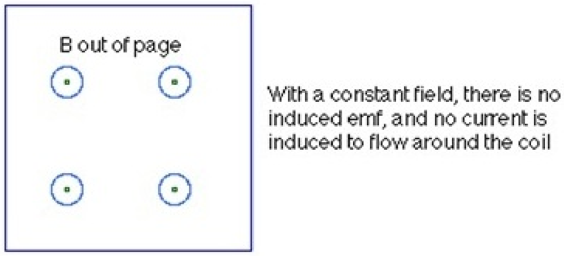
The negative sign in Faraday's law comes from the fact that the e.m.f. induced in the coil acts to oppose any change in the magnetic flux. This is summarized in Lenz's law.

Lenz's law: The induced emf generates a current that sets up a magnetic field which acts to oppose the change in magnetic flux.

Another way of stating Lenz's law is to say that coils and loops like to maintain the status quo (i.e., they don't like change). If a coil has zero magnetic flux, when a magnet is brought close then, while the flux is changing, the coil will set up its own magnetic field that points opposite to the field from the magnet. On the other hand, a coil with a particular flux from an external magnetic field will set up its own magnetic field in an attempt to maintain the flux at a constant level if the external field (and therefore flux) is changed.

An example

Consider a flat square coil with N = 5 loops. The coil is 20 cm on each side, and has a magnetic field of 0.3 T passing through it. The plane of the coil is perpendicular to the magnetic field: the field points out of the page.



(a) If nothing is changed, what is the induced e.m.f.?

There is only an induced e.m.f. when the magnetic flux changes, and while the change is taking place. If nothing changes, the induced e.m.f. is zero.

(b) The magnetic field is increased uniformly from 0.3 T to 0.8 T in 1.0 seconds. While the change is taking place, what is the induced e.m.f. in the coil?

Probably the most straight-forward way to approach this is to calculate the initial and final magnetic flux through the coil.

Initial magnetic flux: $\phi_0 = B_0A = 0.3\ (0.2)^2 = 0.012Tm^2$

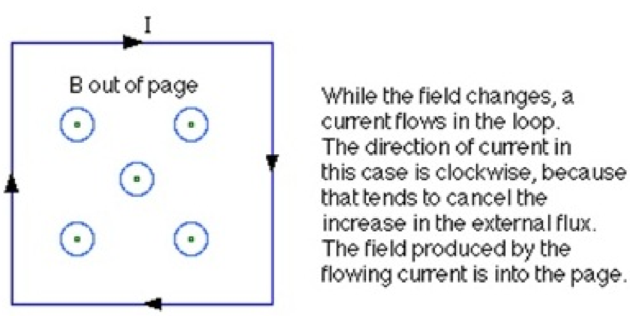
Final magnetic flux: $\phi_f = B_fA = 0.8\ (0.2)^2 = 0.032Tm^2$

The induced e.m.f. is then:

$\varepsilon = -N\Delta\phi/\Delta t = -N\ (\phi_f - \phi_0)/\Delta t = -5\ (0.032 - 0.012)/1.0 = -0.1v$

(c) While the magnetic field is changing, the e.m.f. induced in the coil causes a current to flow. Does the current flow clockwise or counter-clockwise around the coil?

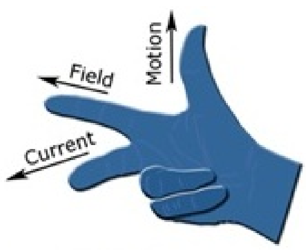
To answer this, apply Lenz's law, as well as the right-hand rule. While the magnetic field is being changed, the magnetic flux is being increased out of the page. According to Lenz's law, the emf induced in the loop by this changing flux produces a current that sets up a field opposing the change. The field set up by the current in the coil, then, points into the page, opposite to the direction of the increase in flux. To produce a field into the page, the current must flow clockwise around the loop. This can be found from the right hand rule.



One way to apply the rule is this. Point the thumb on your right hand in the direction of the required field, into the page in this case. If you curl your fingers, they curl in the direction the current flows around the loop – clockwise.

Fleming's Right Hand Rule

The rule states that, if our right hand is held so that the thumb, the forefinger and the middle finger are perpendicular to one another, the thumb will represent the direction of motion in the conductor, the fore finger will represent the direction of the magnetic field, while the middle or second finger will represent the direction of the induced current. The difference between Fleming's right hand and left hand rule is that the right hand rule is used for induced current or e.m.f. while the left hand refers to the force in the conductor.



Induction Coil

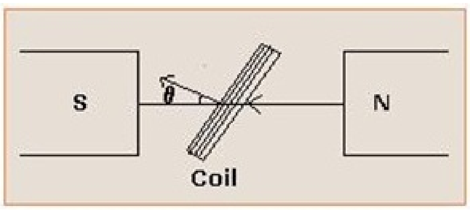
Induction Coil, a device for converting low-voltage direct current (DC) into high-voltage alternating current (AC). The coils are used chiefly in the electrical systems of automobiles and to operate X-ray tubes.

A typical induction coil has a core of soft iron, a primary coil, and a secondary coil. The primary coil consists of a few turns of fairly heavy wire around the core; the secondary consists of many turns of fine wire around the primary. The primary coil forms part of a circuit called the primary circuit that includes a direct current source and a circuit breaker, or interrupter.

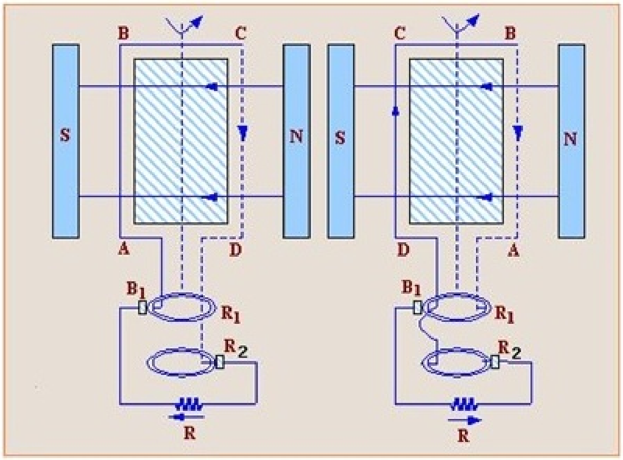
When the primary circuit is closed, direct current flows through the primary coil, producing a magnetic field. As the magnetic field builds up, it induces an electric current in the secondary coil. At the same time, the iron core becomes magnetized. The magnetized core draws the interrupter away from a metal contact, breaking the primary circuit. The direct current in the primary coil ceases and the coil's magnetic field collapses, again inducing an electric current in the secondary coil,

only in the opposite direction. Simultaneously, the core loses its magnetism and releases the interrupter, which is pulled back against the contact by a spring. The cycle continues to repeat rapidly, supplying an alternating current at the terminals of the secondary coil. The voltage in the secondary coil is higher than in the primary coil because of the greater number of turns in the secondary coil.
A capacitor, or condenser, is often used with an induction coil. The capacitor prevents sparking between the interrupter and contact by briefly storing the electric charge that would otherwise jump the gap between them.

Principle



AC Dynamo is based on the phenomenon of electromagnetic induction. That is, when the relative orientation between the coil and the magnetic field changes, the flux linked with the coil changes and this induces a current in the coil.
As the armature coil rotates, the angle Q changes continuously. Therefore, the flux linked with the coil changes.
Now, $\phi = N(B.A)$
$= NBA \cos q$
$= NBA \cos wt$
where q is the flux linked with the coil, N is the number of turns in the coil, A is the area enclosed by each three of the coil and B is the strength of the magnetic field.
$E = -d\phi/dt = – d(NBA \cos wt)/dt$ (from Faraday's law of EMF)
$= – NBA (-\sin wt)w$
$E = + NBA w \sin wt$
$e = e_o \sin wt$. This is the EMF Supplied by the A.C. generator
$i = \varepsilon/R = \varepsilon_0/R \sin \omega t = i_0 \sin \omega t$



Armature

ABCD is the armature coil consisting of a large number of turns of the insulated copper wire wound over a laminated soft iron core I. The coil can be rotated about the central axis.

Magnets

N and S are the pole pieces of a strong electromagnet in which the armature coil is rotated.

Slip rings

R_1 and R_2 are two hollow metallic rings to which both ends of the armature coil are connected. These rings rotate with the rotation of the coil.

Brushes

Brushes B_1 and B_2 are two flexible metal plates or carbon rods. These brushes are used to pass current from the coil to the external load resistance.

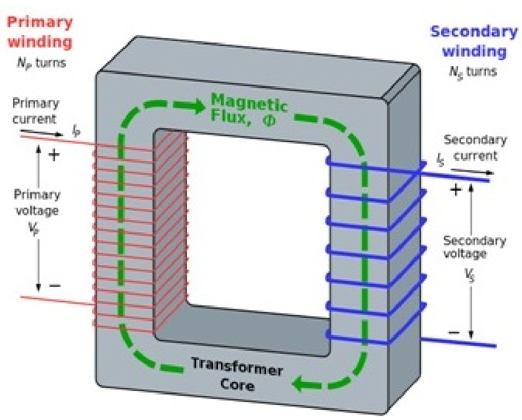
Working

To start with, suppose the plane of the coil is perpendicular to the plane of the paper in which the magnetic field is applied, with AB at the front and CD at the back, the flux linked with the coil is maximum in this position. As the coil rotates clockwise, AB moves inwards and CD moves outwards. According to Fleming's right hand rule, the current induced in AB is from A to B, and in CD, from C to D. In the

external circuit, current flows from B_2 to B_1. After half of the rotation of the coil, AB is at the back and CD is at the front. Therefore, AB starts moving outwards and CD inwards. The current induced in AB is from B to A, and in CD, from D to C. The current flows from B_1 to B_2 through the external circuit. We therefore see that the induced current in the external circuit changes direction after every half rotation of the coil, and hence is alternating in nature.
Alternator is an electric generator which changes the position or direction of flow of voltage and current. The current that changes is called the alternating current.

Transformers and Power transmission Distribution

Transformers also operate on the principle of electromagnetic induction. Transformers are devices that convert electric current from one potential difference (voltage) to another potential difference. For example, the current that comes from a power plant is typically high voltage current, much higher than is needed or than can be used in household appliances. A step-down transformer uses electromagnetic induction to convert the high voltage current in power lines to the lower voltage current needed for household appliances.



We have already seen that a change in flux induces an emf in a coil, given by Faraday's Law:

$\varepsilon = -N\, d\phi/dt$

We have also seen that a voltage in a coil induces a magnetic flux inside the coil. If we were to connect two coils with the same core, the flux, and the rate of change of flux, would be exactly the same inside both coils. We would have created a kind of flux circuit known as a transformer. The ratio between the voltage at the primary coil V_p and the voltage at the secondary coil V_s would have to be (since ϕ is constant):

$V_P/V_s = -N_p \, {}^{d\phi}/_{dt} \, / \, -N_s \, {}^{d\phi}/_{dt} = N_p/N_s$,

where N_p and N_s are the numbers of coils in the primary and secondary coils respectively.

In other words, we can change the voltage of some electricity by varying the number of coils in each coil. In order for this to work, the current used must be an alternating current (AC). This means that the current and voltage are constantly changing sinusoidally, and so there is a sinusoidal change in flux. This means that an e.m.f. is induced in the secondary coil. If the flux did not change (i.e. we were using direct current), then no e.m.f. would be induced, and the transformer would be useless except as a magnet (since it would still have a flux circuit in it).

Ideal Transformers

An ideal transformer is one in which all the electrical energy put into one coil comes out of the other coil. An ideal transformer does not exist, but, since it makes the maths easy, we like to pretend that it does. In this case, the power in must equal the power out:

$P = P_p = P_s = I_p V_p = I_s V_s$,

where I_p and I_s are the currents in the primary and secondary coils, respectively. So:

$V_p = P/I_p$ and $V_s = P/I_s$

By substitution into the transformer equation for voltage:

$N_s/N_p = V_s/V_p = P/I_s \, /P/I_p = /1/I_s \, /1/I_p = I_p/I_s$

$N_s/N_p = V_s/V_p = I_p/I_s$

So, in an ideal transformer, the ratio between the voltages is equal to the ratio between the numbers of coils, but the ratio between the currents is equal to the *reciprocal* of the ratio between the numbers of coils.

Mutual Inductance is the flow of induced current or voltage in a coil due to an alternating or varying current in a neighbouring coil.

Eddy Currents

In reality, the electrical energy is not all conserved, a lot of it is converted into heat by eddy currents. In a transformer, the magnetic flux created by the primary coil induces a current in the core. This occurs in order to oppose the change that produced the magnetic flux (Lenz's Law). The currents flowing in the core are called eddy currents.

These currents produce heat, using up energy and so causing inefficiency. One way of minimizing the effects of eddy currents is to make the core out of iron laminate. This is layers of iron separated by thin layers of an insulator such as varnish. The amplitude of the eddy currents produced is reduced as currents cannot flow through the layers of insulator.

Hysteresis Loss is wasted energy due to reversing the magnetization of core. The core is made to go through a cycle of magnetization during each alternating cycle of the primary current. Hysteresis loss is reduced by the use of special alloys in the core of the primary coil, or by the use of soft iron cores.

I^2 R (or Heat) Loss because the primary and secondary coils have resistance, some energy is lost in the form of heat ($I^2 R$) in the coils. This heat loss can be reduced by using thick wires, or low resistance coils.

Leakage of Magnetic Flux

Some energy is lost due to leakage of magnetic flux. This arises because not all the lines of induction due to current in the primary coil pass entirely through the iron core. This loss is reduced by special forms of coil winding, or by efficient core design.

It is because of the above losses that the efficiency of practical transformer is less than 100%.

The efficiency of a transformer is defined by

Efficiency = Output Power/Input Power x 100% = Power in the secondary coils/Power in the primary coils x 100%

Example

1. Find the turns ration in a transformer which delivers a voltage of 120 volts in the secondary coil from a primary voltage of 60 volts.

Solutions

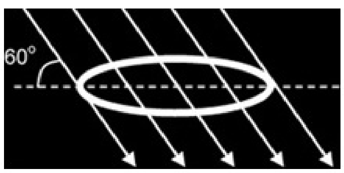
Turns ratio = N_s/N_p

$E_s/E_p = N_s/N_p$

$120/60 = N_s/N_p = 2$

Turns ratio = 2

2. A magnetic field of flux density 20 T passes down through a coil of wire, making an angle of 60° to the plane of the coil as shown. The coil has 500 turns and an area of 25 cm^2. Determine:



(i) the magnetic flux through the coil

$\phi = B\ A \sin \phi$

$= 20 (\sin 60°)\ 25 \times 10^{-4}$

$= 0.0433$ Wb

(ii) the flux linkage through the coil

$\Phi = N\ \phi$

$= 500 \times 0.0433 = 21.65$ Wb

Questions:

1. An ideal transformer transforms a 50A current into a 1A current. It has 40 coils on the primary coil. How many coils are in the secondary coil?

A. 5.1 KV B. 6.2 KV C. 7.3 KV 4. 8.4 KV

2. A step-down transformer has 300 coils on one coil, and 50 coils on the other. If 30 kV AC is put in, what voltage comes out?

A. 6 KV B. 5 KV C. 4 KV D. 3.5 KV

3. A magnetic field of flux density 20 T passes down through a coil of wire, making an angle of 60° to the plane of the coil as shown. The coil has 500 turns and an area of 25 cm^2.

Determine the magnetic flux through the coil

A. 0.786 Wb B. 0.0433 Wb C. 1.434 Wb D. 0.434 Wb

4. Which of these statements is not correct?

A. The magnetic flux is a measure of the number of magnetic field lines passing through an area.

1. When a magnet is pushed into a coil as shown, the galvanometer deflects in one direction momentarily.

C. In a transformer, the magnetic flux created by the primary coil induces a current in the core.

D. because the primary and secondary coils have resistance, some energy is lost in the form of heat ($I^2 R$) in the coils. This heat loss can be reduced by using flexible wires.

5. What year did the English physicist Michael Faraday observed the electromagnetic induction effect?

A 1830 B. 1831 C. 1832 D. 1833

Answers

1. A 2. B 3. B 4. D 5. B

WEEK 3

First Term SS 3, Physics,

Topic: ELECTRIC FIELD

Coulombs Law

The electric force between two point charges is directly proportional to the magnitude of each charge (q_1, q_2), inversely proportional to square of the separation between their centers (r), directed along the separation vector connecting their centers.

This relationship is known as Coulomb's Law. Charles Augustin de Coulomb (1736-1806) France. As an equation it is usually written in one of two forms …

$F = k_e q_1q_2/r^2$ or $F = 1/4\pi e_o. q_1q_2/r^2$

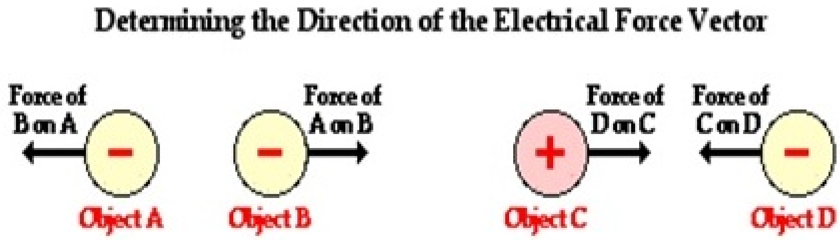
Electrostatic constant $k_e = 8.99 \times 10^9$ Nm/C^2,

Vacuum Permittivity $e_o = 8.85 \times 10^{-12}$C^2/Nm^2

When two charges have the same sign their product is positive, which means the force vector is directed with the separation vector and the action is repulsive.

When two charges have the opposite sign their product is negative, which means the force vector is directed against the separation vector and the action is attractive.

The electrical force, like all forces, is typically expressed using the unit Newton. Being a force, the strength of the electrical interaction is a vector quantity that has both magnitude and direction. The direction of the electrical force is dependent upon whether the charged objects are charged with like charge or opposite charge and upon their spatial orientation. By knowing the type of charge on the two objects, the direction of the force on either one of them can be determined with a little reasoning. In the diagram below, objects A and B have like charge causing them to repel each other. Thus, the force on object A is directed leftward (away from B) and the force on object B is directed rightward (away from A). On the other hand, objects C and D have opposite charge causing them to attract each other. Thus, the force on object C is directed rightward (toward object D) and the force on object D is directed leftward (toward object C). When it comes to the electrical force vector, perhaps the best way to determine the direction of it is to apply the fundamental rules of charge interaction (opposites attract and likes repel) using a little reasoning.



Example

Suppose that two point charges, each with a charge of +1.00 Coulomb are separated by a distance of 1.00 meter. Determine the magnitude of the electrical force of repulsion between them.

Q_1 = 1.00 C, Q_2 = 1.00 C, d = 1.00 m, F_{elect} =?

F = $1/4\pi e_o$. q_1q_2/r^2

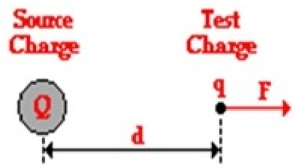
= (9.0 x 10^9 Nm^2/C^2) x (1.00 C) x (1.00 C) / $(1.00\ m)^2$

= 9.0 x 109 N

Electric Field Intensity

Electric field strength is a vector quantity; it has both magnitude and direction. The magnitude of the electric field strength is defined in terms of how it is measured. Let's suppose that an electric charge can be denoted by the symbol Q. This electric charge creates an electric field; since Q is the source of the electric field, we will refer to it as the **source charge**. The strength of the source charge's electric field could be measured by any other charge placed somewhere in its surroundings. The charge that is used to measure the electric field strength is referred to as a testcharge since it is used to *test* the field strength. The test charge has a quantity of charge denoted by the symbol q. When placed within the electric field, the test charge will experience an electric force – either attractive or repulsive. As is usually the case, this force will be denoted by the symbol F. The magnitude of the electric field is simply defined as the force per charge on the test charge.

Note that in the figure below d indicates the distance between the source charge and the test charge, which could also be denoted by 'r'.



Electric Field Strength = Force/Charge

If the electric field strength is denoted by the symbol E, then the equation can be rewritten in symbolic form as

$E = F/q$, $F = qE$, Electric Force = Electric Charge x Electric Field

We can obtain the equation for the Field Intensity E due to a point charge q at a distance r from that charge making use of these two equations $E = F/q$ and $F = qE$ we assume the point charge is situated vacuum

$F = 1/4\pi e_o. Qq/r^2$

$E = F/q = 1/4\pi e_o\ Qq/r^2q = Q/4\pi e_o\ r^2$

Hence we can write

$E= Q/4\pi e_o\ r^2$

This is the field intensity due to a charge Q at a distance r from the charge.

Electric Potential

In a gravitational field, forces of attraction acts between masses. Therefore in order to separate two masses, work must be done against the field.

Similarly, in electrostatic field where forces of attraction or repulsion act between charges, work must be done against this field in order to move charges against the field.

The gravitational potential at any point in a gravitational field is the work done per unit in bringing mass to that point from an arbitrarily chosen zero level of the earth. Similarly in an electrostatic or electric field the electric potential at a point is described as the work done per unit positive charge in bringing charge to that point from an arbitrarily chosen zero of potential. If work is done against the field, the potential is positive. If work is done by the field, the potential is negative. Points of positive potentials are said to be at a higher potential than those of negative potential. Thus in an electric cell, the positive terminal is at a higher potential than the negative terminal. The arbitrarily chosen zero of potential is taken at infinity, but in practical terms, ***the earth is taken to be at zero potential***. Thus whenever a conducting body is connected to the earth it is said to be ***'grounded' or 'earthed'*** or to have a zero potential.

Electric Potential (V) at a point is defined as the work done in bringing unit positive charge from infinity to that point against the electrical forces of the field.

Potential difference (V_{AB}) between two points A and B is the work done in taking unit positive charge from one point to the other in the electric field.

Both the potential and the potential difference are scalar quantities having the dimension of work/charge. The unit is the volt.

1 volt = 1 Joule/Coulomb

The potential difference between two points is one volt if the work done in taking one Coulomb of positive charge from one point to the other is one joule.

W = qV

(Joules) = (Coulomb x Volts) or ***V = W/q***

Electric intensity is related to the electric potential difference between two points through the equation.

E = v/d, where d is the distance between two points in an electric field

Hence we have, ***V =Ed***,volt = $NC^{-1}m$

The unit of ***E = v/d*** can also be in volt per meter but we have E = $q/4\pi e_o d^2$ or $q/4\pi e_0 d^2$

Putting this in ***V =Ed***, we have

$v = q/4\pi e_o d^2$ x d or $v = q/4\pi e_o d$

Where v is the potential at a point due to a charge q at a distance d from the charge.

Two parallel plates are charged to a voltage of 40V. If they are separated by a distance of 10.0 cm, calculate the electric intensity between them.

Solution

E = V/d

= 40 (volt)/0.1 m

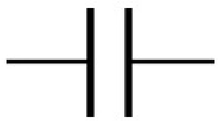
= 400 Vm^{-1}

Capacitors and Capacitance

Capacitors

A capacitor is an electronic device for storing charge. Capacitors can be found in almost all but the most simple electronic circuits. There are many different types

of capacitor but they all work in essentially the same way. A simplified view of a capacitor is a pair of metal plates separated by a gap in which there is an insulating material known as the dielectric. This simplified capacitor is also chosen as the electronic circuit symbol for a capacitor is a pair of parallel plates as shown below
The symbol above is for an unpolarised capacitor
On the picture are different capacitors look.

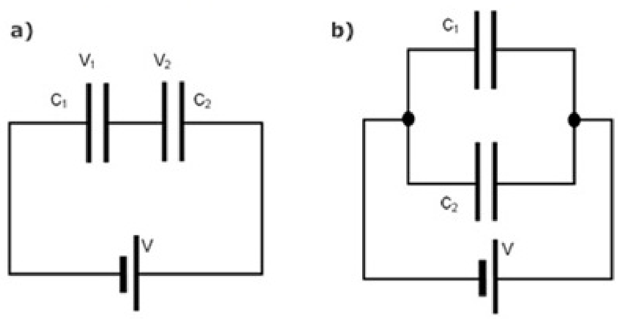


The symbol above is for an unpolarised capacitor

On the picture are different capacitors look.

Arrangement of Capacitors – Series and Parallel

Arrangement of Capacitors – Series and Parallel



Figures above are Capacitor arranged in Series and Parallel

Series

Consider the series network of capacitors shown in the figure above where the positive plate is connected to the negative plate of the next. What is the equivalent capacitance of the network? Look at the plates in the middle, these plates are physically disconnected from the circuit so the total charge on them must remain constant. It follows that when a voltage is applied across both of the capacitors, the charge $+Q$ on the positive plate of capacitor C_1 must be balanced by the charge –

Q on the negative plate of capacitor C_2. The net result is that both capacitors possess the same charge Q. The potential drops V_1 and V_2 across the two capacitors are in general, different. However, the sum of these drops equals the total potential drop V applied across the input and output wires. $V=V_1 + V_2$. The equivalent capacitance of the pair is again $C_T=Q/V$. Thus, $1/C_T = V/Q = (V_1 + V_2)/Q = V_1/Q + V_2/Q$ giving

By connecting capacitors in series you store less charge so does ever make sense to connect capacitors in series? It is sometimes done because capacitors have maximum working voltages, and with two 900 volt maximum capacitors in series, you can increase the working voltage to 1800 volts.

Parallel

For a parallel circuit such as in Figure above, the voltages are the same across each component. However the total charge is divided between the two capacitors since it must distribute itself such that the voltage across the two is the same. Also, since the capacitors may have different capacitances C_1 and C_2 the charges Q_1 and Q_2 must also be different. The equivalent capacitance C_T of the pair of capacitors is simply the ratio Q/V where $Q=Q_1+Q_2$ is the total stored charge. It follows that $C_T = Q/V = (Q_1+Q_2)/V = Q_1/V + Q_2/V$ giving

$1/C_t = 1/C_1 + 1/C_2$

$C_T = C_1 + C_2$

The overall capacitance increases by adding together capacitors in parallel so we create larger capacitances than is possible using a single capacitor.

Capacitor – Energy Stored

The energy stored in a capacitor can be expressed as

$W = 1/2\ CV^2$

Where W = energy stored (Joules), C = capacitance (Farad), V = potential difference (Voltage)

Questions

1. Two balloons are charged with an identical quantity and type of charge: -6.25 nC. They are held apart at a separation distance of 61.7 cm. Determine the magnitude of the electrical force of repulsion between them.

A. 8.0 x 10^{-6} N B. 9.23 x 10^{-7} N C. 7.54 x 10^{-8} N D. 54 x 10^{-5} N

2. Two balloons with charges of +3.37 µC and -8.21 µC attract each other with a force of 0.0626 Newton. Determine the separation distance between the two balloons.

A. +1.99 m B -1.99 m C. 7.34 m D. 2.65 m

3. The energy stored in a *10 µF* capacitor charged to *230 V* can be calculated as

A. 1.46 J B. 0.26 J C. 8 J D. 3.45 J

4. Electric field strength has

A. Only direction B Only Magnitude C. Magnitude and direction D. None of the options

5. By connecting capacitors in series you store

A. More charges B. No Charge C. Average charges D. Less charge

Questions

1. B 2. A 3. B 4. C 5. D

WEEK 4

First Term SSS3, Physics,

Topic: **CURRENT ELECTRICITY**

CURRENT ELECTRICITY

A continuous flow of charges is produced from primary and secondary cell.

a. **Primary Cells**

Primary cells are those from which current is produced as a result of non-reversible chemical changes taken place between the various components of the cell.

A simple cell is an example of a primary cell. When all the zinc of such a cell is used up through the chemical action with sulphuric acid, by which the current is produced, the cell cannot be restored or recharged to its original condition by passing a charging current through the cell in the reverse order. In order to cause primary cell to produce current once again, they must be resupplied with fresh active materials (zinc, copper, sulphuric acid).

The disadvantage of primary cell is that the chemicals in them are gradually used up when the cells are in use.

Secondary cells (or Accumulator)

The secondary class of cells are called secondary cells of accumulators.

These are cells whose chemical actions can be reversed by driving a current through them in a direction opposite to the current they supply. Such cells can therefore be recharged and used for a long time.

The main advantage of accumulators is that they have very low internal resistance (unlike primary cells). They can therefore provide large current in their terminal potential difference.

They are used in motor cars to provide energy for the spark igniting the petrol in the engine, and also in telephone exchanges to provide current in telephone cables when required.

Examples of Primary Cells

There are three types of primary cells; the simple cell, Daniel cell and Leclanche cell.

The Daniel Cell

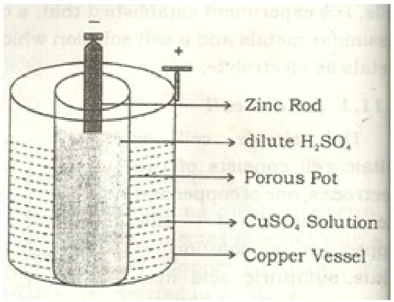
It consists of (i) a copper vessel filled with a saturated copper sulphate solution (ii) a porous pot containing dilute sulphuric acid (H_2SO_4) which is the electrolyte for this cell. (iii) an amalgamated zinc rod immersed in the acid. The zinc rod acts the negative terminal of the cells while the copper vessel acts as the positive terminal. The copper sulphate solution acts as the depolarizer for the Daniel cell.

The initial e.m.f produced by the Daniel cell is about 1.1 volts and cell provides only a small current for some time. The Daniel cell is no longer in use.

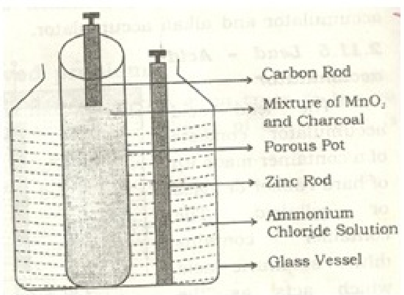
Leclanche Cell

a. **The wet Type**

The internal structures of this cell consist of (i) the carbon contained in a porous pot. This carbon rod acts as the positive terminal of the cell. (ii) A porous pot containing a mixture of manganese dioxide and powered carbon, parked around the carbon rod. (iii) A glass of porcelain jar filled with a solution of ammonium chloride. (iv) A zinc rod immersed in the in the ammonium chloride solution.



Daniel Cell



Leclanche Cell

In this cell the zinc rod is the negative terminal, the carbon rod is the positive terminal, the ammonium chloride is the electrolyte and the manganese dioxide is the depolarizing agent depolarization however takes place more slowly than the rate at which hydrogen is liberated by the action of zinc and ammonium chloride solution.

Therefore when the cell is in continuous use, some depolarization takes place and the e.m.f. of the cell drops. Another defect of the wet leclanche cell is that it is cumbersome to carry about without spilling the liquid.

The leclanche cell is used only where intermittent current is required e.g. in electric bells. The advantage of the cell is that its e.m.f. is relatively high, about 1.5 volts and its chemicals are cheap.

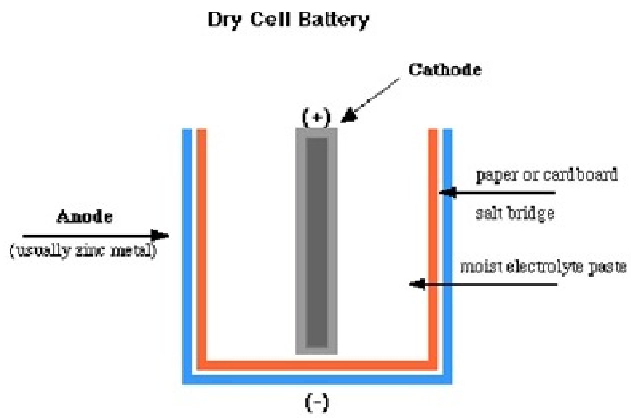
b. **Leclanche Cell (the dry type)**

1. The electrolyte is ammonium chloride in the form of a paste or jelly and it is mixed with starch and flour.
2. The positive terminal is a carbon rod.
3. The negative terminal is the zinc container.
4. The depolarizer of manganese dioxide mixed with powdered carbon is contained ina muslin bag round the carbon rod.

5. A cardboard disc is used at the bottom to prevent the carbon pole from touching the negative container to prevent the short circuiting of the cell.

The e.m.f. of a dry leclanche cell is about 1.5 volts and the chemical reaction taking place within this cell is the same as that of the wet type. The dry cell like the wet leclanche cell cannot maintain current for a long time because it polarizes if continuously owing to the hydrogen liberated at the carbon rod. Hydrogen is produced at a faster rate than can be depolarized by the manganese dioxide.

Dry cells are used in torches, bicycle lamps, electric bells and transistor radios.



Secondary Cells

There are two types

a. The lead acid accumulator and

b. The alkaline or Nickel – Iron (NIFE) accumulator

The simple form of the cells consists of (i) a positive pole of lead peroxide (chocolate or dark brown in cooler) (ii) a negative pole of lead plate (grey in colour) and an electrolyte of dilute sulphuric acid.

In modern commercial accumulators, both the positive and negative plates are made of grids of lead antimony alloy, the holes and grooves of which are filled with either the lead or the lead peroxide. The positive plate containing the lead peroxide and the negative plates containing lead are assembled alternatively in groups and are separated by insulators. All the positives are connected together to form one positive terminal of the accumulator, and the negative plates are also connected together to form the negative terminal. The whole frame work is enclosed in a plastic container filled with sulphuric acid.

When fully charged the e.m.f. of the cell is about 2.2 volts and the relative density of the dilute sulphuric acid is about1.25 the cell is considered completely discharged when the relative density of the acid falls to 1.15.

Using or Discharging the Accumulator

The accumulator is used to supply current to an external resistance e.g. an electric bulb. After some time of such use, the current will cease to flow and will be indicated by both the ammeter and the bulb. The accumulator is said to be discharged. A little red sulphate is formed at both of the plate and the relative density of the acid and the e.m.f. of the cell decrease below the original levels.

2. **Nickel Iron (NIFE) Accumulator**

The poles of this accumulator are positive nickel plates and negative iron plate. The electrolyte is a solution potassium hydroxide. The e.m.f. when fully charged is about 1.5 volts and falls to 1.3 volts after some time. The NOFE accumulator has a longer life span than the lead accumulator.

Charging or Recharging an Accumulator

To charge the accumulator we pass direct current (d.c.) through it in the opposite direction to that which it supplies. The positive terminal of the d.c. is connected to the positive terminal of the accumulator and the negative terminal of the d.c. is connected to the negative terminal of the accumulator.

The rheostat R in the charging circuit is essential as the sulphuric acid has a very low resistance such a 0.02 ohm. The ammeter A is also included in the circuit to ensure that only the charging current recommended by the manufacturer is used.

Charging the accumulator is considered complete when the relative density of the acid attains the recommended value of 1.25 and the e.m.f. of the accumulator is about 2.2 volts.

Questions

1. Primary cells are those from which current is produced as a result of

………………………………………………

taken place between the various components of the cell.

A. reversible chemical changes B. non-reversible chemical changes C. reversible chemical reactions D. non-reversible chemical reaction.

2. Which of these does not belong to the class of primary cells?

A. the simple cell B. Daniel cell C. Leclanche cell D. The lead acid accumulator

3. In Nickel Iron (NIFE) Accumulator the e.m.f. when fully charged is about ………………. and falls to ……………. volts after some time.

A. 4.3 to 3.6 volts B. 1.5 to 1.3 volts C. 2.5 to 2.2 volts D. 1.1 to 0.9 volts

4. The initial e.m.f produced by the Daniel cell is about

A. 1.1 volts B. 2.1 volt C. 3.1 volt D. 4.1 volts

5. Charging the accumulator is considered complete when the relative density of the acid attains the recommended value of ……………….

A. 2.5 B. 1.25 C. 3.0 D. 1.76

Answers

1. B 2. D 3. B 4. A 5. B

WEEK 5

First Term SSS3, Physics,

Topic: ELECTROLYSIS

Conduction Through liquids

Electrolysis is the process by which the movement of an electric current through a solution librates electrons. Chemicals in a container are decomposed in order to generate current.

Michael Faraday studied this process extensively and laid the foundation for the theory of electrolysis. Some liquids are good conductors while others are poor conductors of electricity. Good conductors are known as electrolytes, poor conductors are non-electrolytes.

Liquids such as solutions of acids, bases and salts are generally good conductors. Liquids such as benzene and paraffin or kerosene are poor conductors. Organic compounds are generally poor conductors. While pure water is also a poor conductor, water containing some dissolved salts conducts moderately.

A Voltameter is device for measuring the quantity of electricity passing through a conductor by the amount of electrolytic decomposition it produces, or for measuring the strength of a current by the amount of such decomposition in a given time.

1. **The Electrolytes** is the liquid or molten substance which conducts a current and is decomposed by it. i.e. it contains mobile ions and undergoes decomposition and is called electrolyte, e.g. acids, bases, common salts, etc.
2. **Non-electrolyte** this is a substance which, either in molten state or in solution, does not allow the flow of an electric current.
Thus, it is a substance that does not conduct current or undergo decomposition, e.g. organic solvents such as benzene, paraffin, sugar, salt, etc.
3. **Electrodes** are materials in the form of a rod or plate through which current enters or leaves the electrolyte. There are two electrodes: (a) the positive electrode through which current enters the electrolyte is called the anode. (b) The negative electrode through which liquid leaves the electrolyte is called the Cathode.

4. Anode is a positive (+ve) electrode at which the electron enters and current leaves the electrolyte.

5. **Cathode** is a negative (-ve) electrode at which the electron enters and current leaves the electrolyte

6. **Voltameter**: The whole apparatus consisting of the vessel, electrolyte and electrodes is called the voltameter.

7. **Ions**: They are charged particles which exists in electrolytes and take part in electrolysis

These are the immediate products of decomposition of an electrolyte. The ions which go to the anode are called anions, those ions which go to the cathode are called cations

Current Electricity

Current electricity consists of fast moving negatively charged electrons. Currents travels in material which allow the flow of electrons called conductors. Current is produced in a simple circuit consisting of a battery (which is source), a bulb (which is the lighting) and a tap key (which is used in controlling the simple system).

The current in the battery due to the force applied on it. The force is not visible but it is performed by some chemicals in the battery. The electrical pressure is called the voltage.

Dynamics of charged particles (ions) in electrolytes

The Ionic theory

In an electrolyte there are positively and negatively charged particles called ions. The molecules that constitute the electrolyte are split in solution into these ions through the process known as electrolytic dissolution. The dissolution of an electrolyte occurs irrespective of whether or not an electric field is applied to the electrolyte. Ions in an electrolyte execute random movements until a battery is connected to the electrodes of a voltmeter. As soon as a p.d. is set up across the electrodes the positive ions drift to the cathode, which is at a negative potential while the negative ions drift to the anode which is at a positive potential.

This directional movement of ions is the electric current flowing through the electrolyte. Such movement ceases as soon as the battery is disconnected, and the ions move randomly once again.
Thus electrolytic solutions are able to conduct electricity because the electrolytes in solution can dissociate into ions Non electrolytes in solution do not easily dissociate into ions.

Examples of Electrolysis
Pure water is a poor electrolyte. Because of that, a few drops of sulfuric acid will help to increase its conductivity by adding more mobile electrons
As such the ions present in the electrolyte now are
From Sulfuric Acid: H^+ and SO_4^{2-}
From Water: H^+ and $OH^–$
Inert electrodes like Carbon or Platinum electrodes are used so that the acid solutions do not corrode the electrodes easily.

At the Cathode:
H+ migrate to the cathode and accepts an electron to become a Hydrogen atom. After that, the Hydrogen atoms form a covalent bond with other hydrogen atoms to become Hydrogen gas molecules
$2H^+ + e^– –> H_2$
Observation:
Effervescence of colourless, odourless gas at the cathode
At the Anode:
SO_4^{2-} and $OH^–$ migrate to the anode, where $OH^–$ is preferentially discharged due to it having a lower position in the Electrochemical series as compared to SO_4^{2-}
$4OH^– –> 2H_2O + O_2 + 4e^–$
Therefore Overall Equation:
$2H_2O(l) –> 2H_2 (g)+ O_2(g)$

Michael Faraday's Laws of Electrolysis

First Law: It shows that the mass (m) of substance deposited at the cathode during electrolysis is directly proportional to the quantity of electricity (total charge q) passed by the electrolyte.

i.e. $M = ZQ$, where $Q = It$, $M = ZIt$, $Z = M/It$, where Z is the constant of proportionality and is known as electrochemical equivalent (E.C.E.) of the substance.

Thus, **electrochemical equivalent (ECE)** may be defined as "the mass of the ion deposited by passing a current of one Ampere for one second (i.e., by passing Coulomb of electricity)". It's unit is gram per coulomb.

Coulomb is the unit of electrical charge.

96500Coulombs electrons = 1 mole electrons.

1 Coulomb = $6.023 \times 1^{023}/96500 = 6.85 \times 10^{18}$ electrons, or 1 electronic charge 1.6×10^{-19} Coulomb.

Second law: It states that, "When the same quantity of electricity is passed through different electrolytes, the masses of different ions liberated at the electrodes are directly proportional to their chemical equivalents (Equivalent weights)." i.e.,

$M_1/W_2 = E_1/E_2$ or Z_1I_t/Z_2I_t or $Z_1/Z_2 = E_1/E_2$ ($\therefore W = ZI_t$)

Thus the electrochemical equivalent (Z) of an element is directly proportional to its equivalent weight (E), i.e., $E \propto Z$ or $E = FZ$ or $E = 96500 \times Z$

where, F = Faraday constant = 96500 C mol^{-1}

So, 1 Faraday = 1F =Electrical charge carried out by one mole of electrons.

1F = Charge on an electron × Avogadro's number.

$1F = e^- \times N = (1.602 \times 10^{-19}c) \times (6.023 \times 10^{23}\ mol^{-1})$

Number of Faraday = number of electrons passed/6.023×10^{23}

(3) Faraday's law for gaseous electrolytic product For the gases, we use $V = It\ Ve/96500$

where, V = Volume of gas evolved at S.T.P. at an electrode

Ve = Equivalent volume = Volume of gas evolved at an electrode at S.T.P. by 1 Faraday charge

(4) Quantitative aspects of electrolysis: We know that, one Faraday (1F) of electricity is equal to the charge carried by one mole (6.023×10^{23}) of electrons. So, in any reaction, if one mole of electrons is involved, then that reaction would consume or produce 1F of electricity. Since 1F is equal to 96,500 Coulombs, hence 96,500 Coulombs of electricity would cause a reaction involving one mole of electrons.

If in any reaction, n moles of electrons are involved, then the total electricity (Q) involved in the reaction is given by, $Q = nF = n \times 96500$ C

Thus, the amount of electricity involved in any reaction is related to,

(i) The number of moles of electrons involved in the reaction,

(ii) The amount of any substance involved in the reaction.

Therefore, 1 Faraday or 96,500 C or 1 mole of electrons will reduce,

(a) 1 mole of monovalent cation,(b) 1/2mole of divalent cation,

(c) 1/3 mole of trivalent cation, (d) 1/n mole of n valent cations.

Uses of Electrolysis

1. **Extraction or isolation of metals:** Electrolysis is a process used in extracting metals. Such elements are usually very reactive, e.g. sodium, potassium, chlorine, oxygen, aluminium, etc. Such elements are found at the top reactivity series. They react too readily and cannot be prepared by the electrolysis in an aqueous solution of one of the salts. They are prepared by the electrolysis of their fused salts.

2. **Purification of Metals:** many do not in free or uncombined state. They exist in a state of combination with other elements. To extract such metals from their ores, they have to be purified. The overall result of the experiment is that the metal is transformed from its impure to its pure state.

3. **Electroplating of Metals:** Electroplating is the process of coating a substance with the layer of another substance. It is done through electrolysis. Object to be plated is fairly cleaned to make sure that the anode deposits sticks firmly. If it is not a conductor, it is first coated with graphite, so that what is to be coated on it sticks firmly. The anode is the pure metal to be deposited, while the cathode is made the electrolytic cell. The electrolyte is a solution of soluble salt and the pure metal used as the anode.

Examples

1. Calculate the time in minute, required to plate a substance a $300cm^2$, a layer of copper 0.6mm thick, if a constant current of 2A is maintained. Assuming the density of copper is $8.8g/cm^3$ and one coulomb liberates 0.00033g copper.

Solution

Given that Area = $300cm^2$, thickness = 0.6mm = 0.06cm

Mass = 0.00033g, density = $8.8g/cm^3$

Density = mass/volume, mass = density x volume

Mass = 8.8 x 300 x 0.06 = 158.4g

Using M = Zit

T = m/ZI = 158.4/2 x 0.00033 = 24000secs = 4000mins

Questions

1. Find the mass of copper deposited on the cathode of a copper voltameter if a current of 0.53A is placed through it for 30 minutes (e.c.e. of copper = 3.3×10^{-4} gC^{-1})

A. 0.465 g B. 0.425 g C. 0.325 g D. 0.315 g

2. Calculate the time in minutes required to electroplate an article of area 300 cm^2 with a layer of copper 0.06 cm thick if a constant current of 2 A is maintained. Assumed that the density of copper is 8.8 gcm^{-3} and that one coulomb liberates 0.00033 g of copper.

A. 400 minutes B. 4000 minutes C. 4500 minutes D. 5000 minutes

3. A copper and a silver voltameter are connected in series, and at the end of a period of time, 5.0 g of copper was deposited, calculate the mass of silver deposited at the same time. Chemical equivalent of copper = 31.5. Chemical equivalent of silver = 108.

A. 47 g B. 17.14 g C. 20.17 g D. 23 g

4. The electrochemical equivalent of a metal is 0.126×10^{-6} kgC^{-1}. The mass of the metal that a current of 5 A will deposit from a suitable bath in 1 hour is

A. 0.0378×10^{-3} kg B. 0.227×10^{-3} kg C. 0.378×10^{-3} kg D. 0.595×10^{-3} kg

5. Ions are .... particles which exists in electrolytes and take part in electrolysis

A. Non charged B. Charged C. Neutral D. Positive

Answers

1. D 2. B 3. B 4. C 5. B

WEEK 6

First Term SSS3, Physics,

Topic: **ELECTRIC MEASUREMENT**

RESISTIVITY

Resistivity: The resistance of a wire or material conductor maintained at a constant temperature is related to its length (l) and its cross-sectional area (A) by the expression

$R = \rho\, l/A$

Where ρ is a constant of proportionality known as the resistivity of the material

Then we could deduce $\rho = RA/l$

Thus ρ = R when l = 1 and A = 1 Hence we define resistivity as follows:

Resistivity is the resistance of unit length of material of unit cross-sectional area.

Where R is measured in ohms, A in m, l in m, the unit of ρ is in ohm-meter (Ωm).

We recall our definition of resistance as the ability of a material to oppose the flow of current through it. The greater the resistivity of a wire, the poorer it is as an electrical conductor. Because of this, the term conductivity is used to specify the current-carrying ability of a material. The greater the conductivity the more easily current flows through the material. Thus materials of high conductivity also have low resistivity. Conductivity, $\sigma = l/\rho$

Electrical conductivity is a measure of the extent to which a material will allow current to flow easily through it when a p.d. is applied at a specific temperature. It is the reciprocal of resistivity.

Example The resistance of a wire of length 100 cm and diameter 0.3 mm is found to be 3.0 ohms. Calculate (i) the resistivity, (ii) the conductivity of the material of the wire.

Solution

(i) Resistivity

$\rho = RA/l$

R= 3 Ω, l = 100 cm = 1.0 m, r = (0.3/2 x 10-3) m

Area = πr2 = π(0.3/2 x 10-3)2 m2

= π(1.5 x 10-4)2 m2

ρ = 3 x 22//7 x (1.5 x 10-4)2/1 Ωm

= 21.21 x 10-8 Ωm

= 2.12 x 10-7 Ωm

(ii) Conductivity, σ = 1/ρ

= 1/2.12 x10=7 (Ωm)-1

= 4.7 x 106 (Ωm)-1

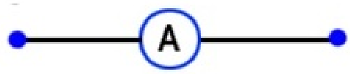
Conversion of Galvanometer to Ammeter

An ammeter is used for measuring electric currents. A galvanometer is used for detecting and measuring very small currents.

We can convert the galvanometer into an ammeter by connecting a suitable resistor in parallel with the galvanometer. A resistor used for this purpose is known as shunt. The shunt is a low resistance wire and is used to divert a large part of current being measured but to allow only a small current to pass through the galvanometer.

Since Galvanometer is a very sensitive instrument therefore it can't measure heavy currents. In order to convert a galvanometer into an Ammeter, a very low resistance known as "shunt" resistance is connected in parallel to Galvanometer. Value of shunt is so adjusted that most of the current passes through the shunt. In this way a Galvanometer is converted into Ammeter and can measure heavy currents without fully deflected

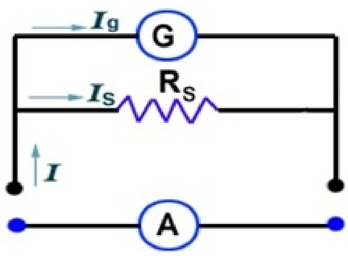
An ammeter is always connected in series to a circuit



Value of Shunt resistance

Let resistance of galvanometer = R_g and it gives full-scale deflection when current I_g is passed through it.
Then, $V_g = I_gR_g$ -------(i)

Let a shunt of resistance (R_s) is connected in parallel to galvanometer. If total current through the circuit is I.



Then current through shunt:

Is = (I-Ig)

potential difference across the shunt:

Vs= IsRs

Vs = (I – Ig)Rs ——-(ii)

But Vs =Vg

(I – Ig)Rs = IgRg

Rs = Ig Rg/I – Ig

Voltmeter is an electrical measuring device, which is used to measure potential difference between two points in a circuit.

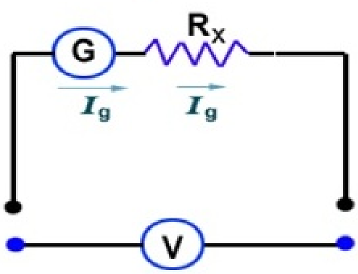
Connection of Voltmeter in Circuit

Conversion of Galvanometer to Voltmeter. The Multiplier

Since Galvanometer is a very sensitive instrument, therefore it cannot measure high potential difference. In order to convert a Galvanometer into voltmeter, a very high resistance known as "series resistance" is connected in series with the galvanometer.

Value of series resistance

Let resistance of galvanometer = R_g and resistance R_x (high) is connected in series to it. Then combined resistance = $(R_g + R_x)$.



Voltmeter is always connected in parallel to a circuit.

Conversion of Galvanometer to Voltmeter. The Multiplier

Since Galvanometer is a very sensitive instrument, therefore it cannot measure high potential difference. In order to convert a Galvanometer into voltmeter, a very high resistance known as "series resistance" is connected in series with the galvanometer.

Value of series resistance

Let resistance of galvanometer = Rg and resistance Rx (high) is connected in series to it. Then combined resistance = (Rg + Rx).

If potential between the points to be measured = V and if galvanometer gives full-scale deflection, when current "Ig" passes through it. Then,

$V = I_g (R_g + R_x)$

$V = I_gR_g + I_gR_x$

$V - I_gR_g = I_gR_x$

$R_x = (V - I_gR_g)/I_g$

$R_x = V/I_g - R_g$

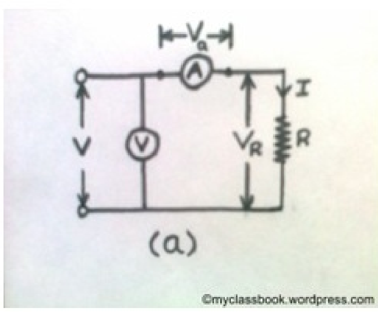
Measurement of Resistance by Ammeter-Voltmeter Method

This is very popular method for measurement of medium resistances since instruments required for this method are usually available in laboratory. The two types of connections employed for ammeter voltmeter method are shown in figures below. In both the methods if readings of ammeter and voltmeter are taken then we can measure value of resistance by using formula:

R_m = voltmeter reading/ammeter reading = V/I

The measured value of resistance R_m, would be equal to the true value R, if the ammeter resistance is zero and the voltmeter resistance is infinite, so that the conditions in the circuit are not disturbed. But in actual practice this is not possible and hence both methods give inaccurate results.

Consider circuit of figure (a):



(a)

Voltmeter-Ammeter method

In this method ammeter measures the true value of current flowing through resistance but voltmeter does not measures the true value of the voltage across the resistance. The voltmeter indicates the sum of the voltage across resistance and ammeter.

Let R_a be the resistance of the ammeter.

Therefore, voltage across the ammeter

$V_a = IR_a$

Measured value of the resistance

$R_{m1} = V/I = V_r + V_a/I = I_r + IR_a/I = R + R_a$

Therefore, true value of resistance, $R = R_{m1} – R_a$

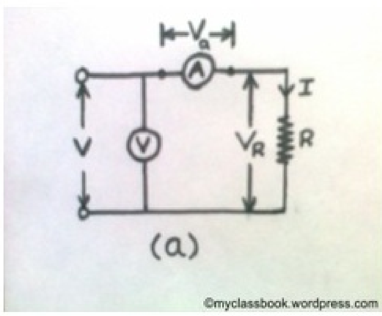
$= R_{m1}(1 – R_a/R_{m1})$

Hence the measured value of the resistance is higher than the true value. It is also clear from the above equation that the true power is equal to the measured only if the ammeter resistance is zero.

Relative error, $E_r = R_{m1} – R/R = R_a/R$

It is clear from the above equation that the error will be small if the value of the measuring resistance is large as compare to the internal resistance of the ammeter .therefore circuit should be used when measuring resistances are high.

Consider circuit of figure (a):



Voltmeter-Ammeter method

In this circuit the voltmeter measures the true value of the voltage across the measuring resistance but the ammeter does not measure the true value of the current flowing through the resistance. The current through the ammeter is the sum of the current through the voltmeter and resistance.

Let R_v be the resistance of the voltmeter.

Therefore current through the voltmeter, $I_v = V/R_v$

Measured value of the resistance $R_{m2} = V/I = V/I_r + I_v = V/V/R + V/R_v = R/1 + R/R_v$

True value of resistance;

$R = R_{m2}R_v/R_v – R_{m2} = R_{m2}\ (1/1 – R_{m2}/R_v)$

From the above equation it is clear the true value of the resistance will be equal to the measured value only when the voltmeter resistance is equal to the infinite. However, if the resistance of the voltmeter is very large as compared to the resistance under measurement:

$R_v >> R_{m2}$

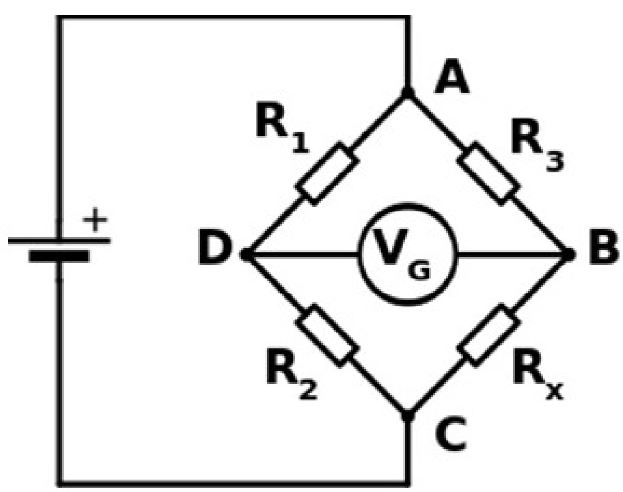
And therefore R_{m2}/R_v is very small

We have $R = R_{m2}\ (1 + R_{m2}/R_v)$

Thus, the measured value of the resistance is smaller than the true value.

Relative error, Er = Rm2 – R/R = R2m2/Rv R
The value of Rm2 is approximately equal to Er = -R/Rv
It is clear from the above equation that the relative error will be small if the resistance under measurement is very small as compared to the resistance of the voltmeter .hence the circuit should be used when the measuring values of resistances are low.

The Wheatstone Bridge
A Wheatstone bridge is an electrical circuit used to measure an unknown electrical resistance by balancing two legs of a bridge circuit, one leg of which includes the unknown component. Its operation is similar to the original potentiometer. It was invented by Samuel Hunter Christie in 1833 and improved and popularized by Sir Charles Wheatstone in 1843. One of the Wheatstone bridge's initial uses was for the purpose of soils analysis and comparison



In the figure, is the unknown resistance to be measured; R1, and are resistors of known resistance and the resistance of is adjustable. If the ratio of the two resistances in the known leg (R2/R1) is equal to the ratio of the two in the unknown leg (Rx/R3), then the voltage between the two midpoints (B and D) will be zero and no current will flow through the galvanometer Vg. If the bridge is unbalanced, the direction of the current indicates whether R2 is too high or too low. R2 is varied until there is no current through the galvanometer, which then reads zero.

Detecting zero current with a galvanometer can be done to extremely high accuracy. Therefore, if R1, R2 and R3 are known to high precision, then Rx can be measured to high precision. Very small changes in Rx disrupt the balance and are readily detected.

At the point of balance, the ratio of

$R2/R1 = Rx/R3$

$= Rx = R2/R1 . R3$

Alternatively, if R1, R2 and R3 are known, but R2 is not adjustable, the voltage difference across or current flow through the meter can be used to calculate the value of Rx, using Kirchhoff's circuit laws (also known as Kirchhoff's rules). This setup is frequently used in strain gauge and resistance thermometer measurements, as it is usually faster to read a voltage level off a meter than to adjust a resistance to zero the voltage.

Derivation

First, Kirchhoff's first rule is used to find the currents in junctions B and D:

$I3 – Ix + IG = 0$

$I1 – I2 – IG = 0$

Then, Kirchhoff's second rule is used for finding the voltage in the loops ABD and BCD:

$(I3 . R3) – (IG .RG) – (I1 . R1) = 0$

$(Ix . Rx) – (I2 . R2) + (IG . RG) = 0$

When the bridge is balanced, then IG = 0, so the second set of equations can be rewritten as:

$I3 . R3 = I1 . R1$

$Ix . Rx = I2 . R2$

Then, the equations are divided and rearranged, giving:

$Rx = I2 . R2 . I3 . R3/ I1 . R1. Ix$

From the first rule, I3 = Ix and I1 = I2. The desired value of Rx is now known to be given as:

$Rx = R3 . R2/R1$

If all four resistor values and the supply voltage (VS) are known, and the resistance of the galvanometer is high enough that IG is negligible, the voltage across the bridge (VG) can be found by working out the voltage from each potential divider and subtracting one from the other. The equation for this is:

VG = (Rx/R3 + Rx – R2/R1 + R2) Vs

where VG is the voltage of node B relative to node D.

Significance

The Wheatstone bridge illustrates the concept of a difference measurement, which can be extremely accurate. Variations on the Wheatstone bridge can be used to measure capacitance, inductance, impedance and other quantities, such as the amount of combustible gases in a sample, with an explosimeter. The Kelvin bridge was specially adapted from the Wheatstone bridge for measuring very low resistances. In many cases, the significance of measuring the unknown resistance is related to measuring the impact of some physical phenomenon (such as force, temperature, pressure, etc.) which thereby allows the use of Wheatstone bridge in measuring those elements indirectly.

Questions

1. If a wire has a resistance of 1.32Ω, a length of 110 cm and a n area of cross-sectional of 0.00415 cm2, Find the resistivity of the material of which it is made.

A. 6.45 x 10-7Ωm B. 4.98 x 10-7Ωm C. 3.46 x 10-7Ωm D. 2.39 x 10-7Ωm

2. A galvanometer has as resistance of 5Ω. By using a shunt wire of resistance 0.05Ω, the galvanometer could be converted to an ammeter, capable of reading 2A. What is the current through the galvanometer?

A. 0.02A B. 0.06A C. 2.09A D. 4.15 A

3. A Wheatstone bridge is an electrical circuit used to measure an unknown ………………………….. by balancing two legs of a bridge circuit, one leg of which includes the unknown component.

A. Potential difference B. electrical resistance C. Voltage D. Current

4. In other to concert a galvanometer into ammeter, the shunt resistance must be connected in …………………… to the galvanometer.

A. Series B. Parallel C. direct D. Indirect

5. A resistance wire of length 2m and of uniform cross-sectional area 5.0 x 10-7 m2 has a resistance of 2Ω. Calculate its resistivity.

A. 5.0 x 10-6 m B. 5.6 x 10-7 m2 C. 5.0 x 10-7 m2 D. 4.7 x 10-5 m2

Answers

1. B 2. A 3. B 4. B 5. C

WEEK 7

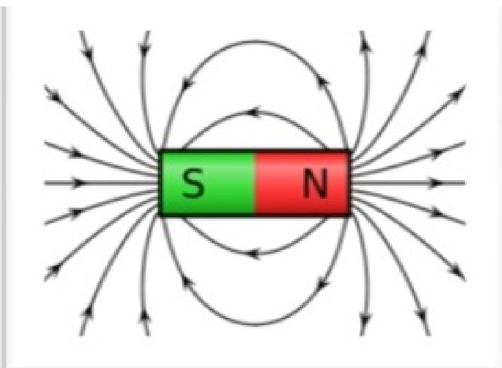
First Term SSS3, Physics,

Topic: MAGNETIC FIELD

INTRODUCTION

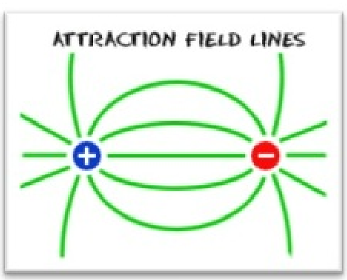
Magnetic field has a region around a magnet in which the influence of the magnet can be felt or detected.

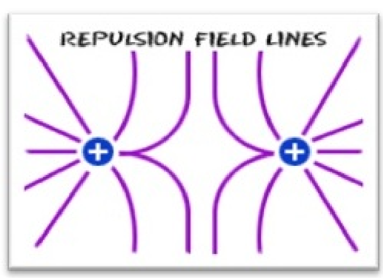
The area around a magnet in which it can attract or repel objects or in which a magnetic force can be detected is called the magnetic field of the magnet.



Patterns of Magnetic Field

Magnetic field patterns can conveniently be observed using iron fillings. The magnet is placed on paper and the iron fillings are sprinkled lightly on the paper around the magnet. The paper is now tapped gently and the iron fillings will be seen to turn and settle in definite directions. The pattern of the magnetic field and for other magnetic arrangements are as depicted below





Magnetic field is a mathematical description of the magnetic influence of electric currents and magnetic materials. The magnetic field at any given point is specified by both a direction and a magnitude (or strength); as such it is a vector field. The term is used for two distinct but closely related fields denoted by the symbols B and H. B refers to magnetic flux density, and H to magnetic field strength.

Magnetic flux density is most commonly defined in terms of the Lorentz force it exerts on moving electric charges

Alternative names for B

Magnetic flux density, Magnetic induction, Magnetic field

Alternative names for H

Magnetic field intensity, Magnetic field strength, Magnetic field, Magnetizing field

The magnetic field can be defined in several equivalent ways based on the effects it has on its environment.

Often the magnetic field is defined by the force it exerts on a moving charged particle. It is known from experiments in electrostatics that a particle of charge q in an electric field E experiences a force F = qE. However, in other situations, such as when a charged particle moves in the vicinity of a current-carrying wire, the force also depends on the velocity of that particle. Fortunately, the velocity dependent portion can be separated out such that the force on the particle satisfies the Lorentz force law;

$F = q(E + v \times B)$.

Here v is the particle's velocity and × denotes the cross product.

The vector B is termed the magnetic field, and it is defined as the vector field necessary to make the Lorentz force law correctly describe the motion of a charged particle. This definition allows the determination of B in the following way.

The command, "Measure the direction and magnitude of the vector B at such and such a place," calls for the following operations: Take a particle of known charge q. Measure the force on q at rest, to determine E. Then measure the force on the particle when its velocity is v; repeat with v in some other direction. Now find a B that makes [the Lorentz force law] fit all these results—that is the magnetic field at the place in question.

SI units, B is measured in teslas (symbol: T) and correspondingly Φ_B (magnetic flux) is measured in webers (symbol: Wb) so that a flux density of 1 Wb/m^2 is 1 tesla. The SI unit of tesla is equivalent to (newton•second)/(coulomb•metre). In Gaussian-cgs units, B is measured in gauss (symbol: G). (The conversion is 1 T = 10,000 G.) The H-field is measured in amperes per metre (A/m) in SI units.

Temporary and Permanent Magnets

Series of experiments carried out have shown that iron, when brought close to a magnetic material, the magnetic property is lost easily.

When iron nails are clung to a magnet, they form a single file in their arrangement, but when the magnet bearing heavy iron nail to form an arrangement is removed, every other nail looses their property, thereby falling off. The conclusion is that iron easily magnetizes and also demagnetizes, where strong magnetism is required for a short time. Examples of temporary magnet are soft iron and electromagnet.

Temporary magnets are employed in the following devices: electric bells, induction coil, telephone ear-piece, magnetic relay, etc.

Steel is not easily magnetized to a magnetic material, because it takes time for the magnetic molecules in steel to be arranged. Steel is not easily demagnetized because it retains its magnetic properties even after the removal of magnets. Steel keeps its magnetism much longer than iron because of these differences in their magnetic properties. Therefore steel is used and most preferred for making permanent magnets. Examples of permanent magnets are steel alloy, etc.

Permanent magnets are employed in the following devices: electric motors, D.C Dynamo, radio loud speaker, aerials of transistor, etc.

Differences between electromagnet and ordinary magnet

Electromagnets	Ordinary Magnets
1 They are very strong in terms of uses.	They are not strong in terms of uses.
2 They are made from cast iron.	They are made from steel.
3 They are temporary magnets.	They are permanent magnets.

Magnetization

Magnetization is the process by which magnetic material is attracted to a single magnet. An example is an experiment using a bar of horse-shoe magnet and many nails. The first nail clings to the magnet followed by second, third, fourth, etc until the magnet force could no longer retain.

Methods of making magnets

1. Single touch method:

This is done by continuous stroking of a steel bar by a permanent magnet. The magnet is raised each time it reaches the end of the steel bar. A stage is reached when the last touch of the stroking process produces or results in a pole opposite.

2. Method of divided touch (Double stroke):

Using the divided method, each half of the steel is stroked continuously in opposite directions by the opposite poles N ad S of the two bar magnets. Divided touch stroking starts in the middle. The same principle is followed in single touch method.

3. Hammering in the earth’s field:

A magnet is made through the influence of the earth’s field (magnetic). The bar is first placed in a north- south direction and inclined at angle 700, to the horizontal axis. The upper part of the magnet is hammered repeatedly. It is found that the lower part has a weak North pole.

4. Electrical method:

This involves using electrical method which is the best way of making magnet. A solenoid of a long coil of insulated copper wire is connected to a battery at both terminals or end points. A steel bar is placed inside the solenoid, current is switched on for few seconds and switched off. A test after the experiment will show that the steel bar is found to be a magnet with North and South poles.

NB: This position of the poles depends on the direction of the current. When the current is clockwise in the coil, the bar has a South Pole at its end and vice-versa.

Demagnetization

This simply means the removal or loss of magnetism from a magnetic material i.e. destroying magnetism. Demagnetism, on the other hand is a process by which the property or substance is removed, causing a breakdown in the magnetic circuit. This can be done through the following methods:

1. Heating method: When a magnet is heated until it is red hot and placed in .a

East-West direction, the magnetic property is lost and would no more behave like a magnet again.
2. Hammering method: The magnet is repeatedly hammered while pointing in an E-W direction, that is, about 900 to the earth's magnetic field of direction. Hammering randomly disorganizes the arrangement of its magnetic property.
3. Electric method: The best way to demagnetize a magnet is by electrical method. A magnet is connected to an AC source and current through a steel bar placed inside a solenoid coil pointing in the direction of E-W, after some seconds, the magnets are taken away from the solenoid and are placed a distance away from the solenoid.

Earth's Magnetic Field
The earth magnetism: The reason for the earth's magnetism is not understood, though it is generally agreed that the earth contains the electrical charges. Field also shows the same similarity to the production of North and South poles by a magnet inside the earth, slightly inclined to the geographical axis. The behaviour is like the south-seeking pole pointing towards the North pole.
If a magnet placed in a uniform sphere of non-magnetic material is mounted on an axis joining the point on the N and S poles of the earth, the line of force will run as follows. Below are very key information as regards the magnetic field of the earth.
As per the most established theory it is due to the rotation of the earth where by the various charged ions present in the molten state in the core of the earth rotate and constitute a current.
(1) The magnetic field of earth is similar to one which would be obtained if a huge magnet is assumed to be buried deep inside the earth at its centre.
(2) The axis of rotation of earth is called geographic axis and the points where it cuts the surface of earth are called geographical poles (Ng, Sg). The circle on the earth's surface perpendicular to the geographic axis is called equator.
(3) A vertical plane passing through the geographical axis is called geographical meridian.
(4) The axis of the huge magnet assumed to be lying inside the earth is called magnetic axis of the earth. The points where the magnetic axis cuts the surface of

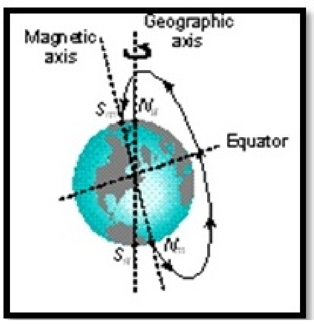
earth are called magnetic poles. The circle on the earth's surface perpendicular to the magnetic axis is called magnetic equator.

(5) Magnetic axis and Geographic axis don't coincide but they make an angle of 17.5° with each other.

(6) Magnetic equator divides the earth into two hemispheres. The hemisphere containing south polarity of earth's magnetism is called northern hemisphere while the other, the southern hemisphere.

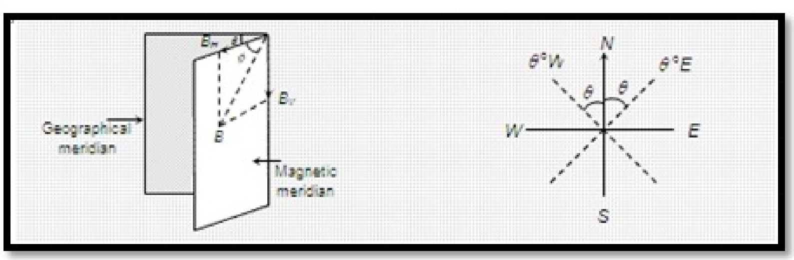
(7) The magnetic field of earth is not constant and changes irregularly from place to place on the surface of the earth and even at a given place in varies with time too.

(8) Direction of earth's magnetic field is from S (geographical south) to N (Geographical north).



Elements of Earth's Magnetic Field

The magnitude and direction of the magnetic field of the earth at a place are completely given by certain quantities known as magnetic elements.
(1) Magnetic Declination: It is the angle between geographic and the magnetic meridian planes.



Declination at a place is expressed at or depending upon whether the north pole of the compass needle lies to the east or to the west of the geographical axis.
(2) Angle of inclination or Dip (ø): It is the angle between the direction of intensity of total magnetic field of earth and a horizontal line in the magnetic meridian.

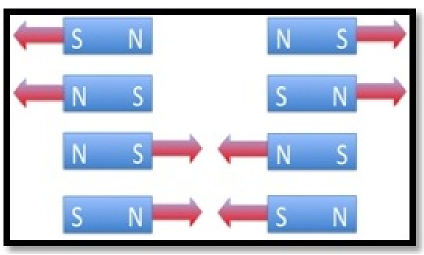
(3) Horizontal component of earth's magnetic field (BH): Earth's magnetic field is horizontal only at the magnetic equator. At any other place, the total intensity can be resolved into horizontal component (BH) and vertical component (BV).

Magnet

What we typically refer to as a magnet, ie. a material that spontaneously produces a magnetic field, is in fact a ferromagnet. The name comes from the region where ferromagnetic stones were found in ancient Greek times, but magnets were also known in the same time period in India and China.
A compass is a freely suspended ferromagnet that can be used for navigation, or, as we will use it in this lecture, to determine the direction of a magnetic field.

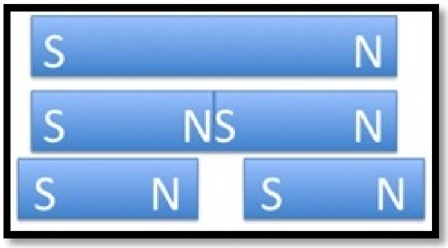
Poles

We are familiar with the idea that magnet has poles, and that like poles repel and unlike poles attract.



What is a Pole?

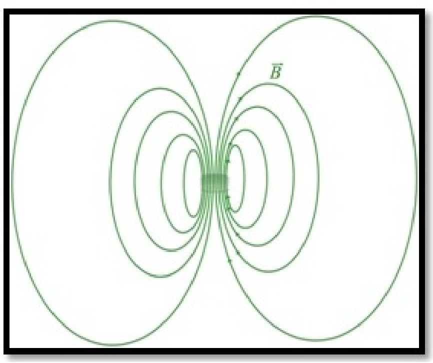
Poles always come in pairs, magnetic monopoles would be highly theoretically interesting, but have not been observed in experiment. A magnetic monopole would be the magnetic equivalent of charge and would act as a source or sink of magnetic field.
We can think of a magnet as having a particular magnetization direction, and we can then understand why if we break a magnet we end up with the creation of another pair of poles. In the context of the bar magnet we can consider the poles to describe the ends of a magnetic material.



Magnetic Field Lines

As with electric fields it can be useful to draw lines which reflect the magnetic field at a point. Field lines point from North to South.

The lack of magnetic monopoles means that magnetic field lines do not begin or end anywhere. So in the case of a bar magnet we can see that the field lines that we can measure outside the magnet continue within it to close the loop.



Earth's magnetic field

The fact that a compass works demonstrates that the Earth has a magnetic field. The magnetic field of the earth is also important in shielding the earth from cosmic radiation. We should note that the magnetic North Pole is actually a magnetic south pole, and the magnetic South Pole is actually a magnetic north pole. The Earth's magnetic poles move around, and in fact can flip their direction completely. This might suggest to us that the magnetic field, unlike in our bar magnet, is due to some kind of dynamic process. In fact the Earth's magnetic field is due to electric currents in the outer liquid core. So we should discuss how an electric current gives rise to a magnetic field.

Magnetic field due to a current carrying wire

If we put current through a wire we can measure that it produces a magnetic field. The direction of the field can be determined by a right hand rule. The experiment we will perform today was first performed by Hans Christian Oersted who first noticed the effect on a compass due to a current during a lecture in 1820.

Magnetic field due to a current carrying loop

If we make the wire in to a loop we can again apply the right hand rule to determine the direction of the magnetic field. We can compare the field distribution to the one we measured for the bar magnet.

Force on a current carrying wire

Having seen that a current carrying wire produces a magnetic field, we can now see whether a magnetic field exerts a force on a current carrying wire. In doing so we will be able to produce a definition of the magnetic field.

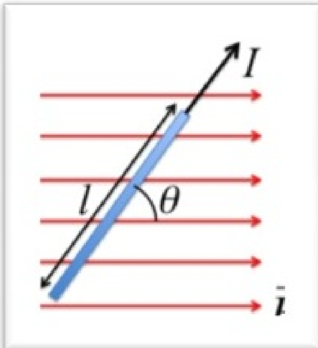
The direction of the force can be determined in a Jumping wire experiment.

Further experiment would reveal that the force on a wire is always perpendicular to both the current and the field, so we can see that we can use a right hand rule to determine the direction of the force.

I make my parallel fingers the field lines, my thumb the current and the force is the direction of the palm of my hand (the direction I would push).

Magnitude of magnetic field

If we consider wire of which length lies within a magnetic field we find that the force depends on as well as the current. We can write an equation that contains this information as well as the right hand rule for the direction that we identified earlier. We can give the length of the wire a direction and make it a vector. The current is then defined to be positive when it flows in the direction of the length vector. The force is then or in the diagram below



We can also chop the length up in to infinitesimal pieces which produce infinitesimal forces to accommodate a wire that changes its direction with respect to a magnetic field, or a non-uniform magnetic field.

Questions

1. Which is not true about electromagnets?
A. They are very strong in terms of uses. B. They are made from cast iron. C. They are temporary magnets. D. They are made from steel.
2. Demagnetization can be done through the following methods, except
A. Heating B. Hammering C. Wetting D. Electric
3. Often the magnetic field is defined by the force it exerts on a moving charged particle. It is known from experiments in electrostatics that a particle of charge q in an electric field E experiences a force. The force is mathematically denoted thus:
A. F = qE B. F = q/E C. F = q + E D. F = q(-E)
4. A compass is used to determine the
A. direction of a magnetic field B. Magnetic field strength C. Magnetic flux density D. Magnetizing field
5. Magnetization is the process by which magnetic material is attracted to
A. Quadruple magnets B. Triple magnets C. Double magnets D. a single magnet

Answers

1. A 2. C 3. A 4. A 5. D

WEEK 8

First Term SSS3, Physics,

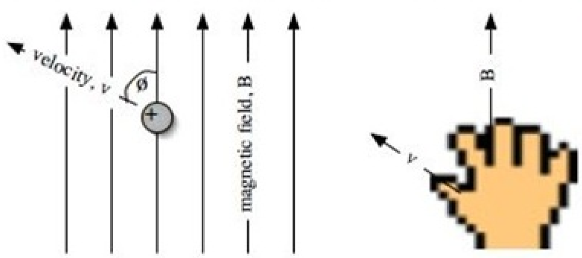
Topic: MAGNETIC FIELD AROUND CURRENT CARRYING CONDUCTOR

MAGNETIC FIELDS

A **magnetic field** is a region of space where a north magnetic monopole experiences a force. The direction of the field is by definition the direction of the force on the north end of a magnet. Since most texts contain diagrams of magnetic fields they will not be reproduced here.

CHARGES IN A MAGNETIC FIELD

Use your **right hand** to determine the direction of **force** on a moving **positively charged particle** in a magnetic field. With the fingers pointing from south to north (the same direction as the field), and the thumb pointing in the direction of the velocity of the particle, the palm points in the direction of the force on the particle. In the example below, the force is away from the observer.



For a negative particle, use your **left hand**.

The magnitude of the force on the particle, F, is

F = qvBsinø. Notice that F = 0 when any of the following are true:

q= 0 ; v = 0; ø = 0^0; ø = 180^0 . Notice that F is greatest when the particle is moving at right angles to the field.

EXAMPLES

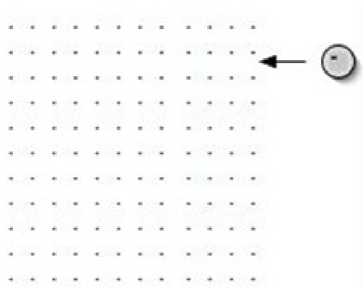
1. An alpha particle (two protons and two neutrons) traveling east at 2.0×10^5 m/s enters a magnetic field of 0.20 T pointing straight up. What is the force acting on the alpha particle?

answer:

With the fingers of the right hand pointing straight up, and the thumb pointing east, the palm points south.

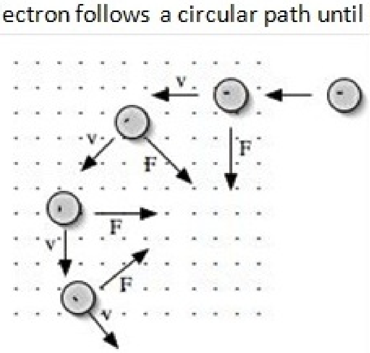
$F = qvB\sin\phi = (2 \times 1.6 \times 10^{-19}\ C)(2.0 \times 10^{5}\ m/s)(0.20\ T)\sin 90^{0} = 1.28 \times 10^{-14}\ N$

2. An electron traveling to the left moves into a magnetic field directed toward the observer. Trace the path of the particle, assuming it eventually leaves the field.



answer:

The moment the electron enters the field, it experiences a force perpendicular to its velocity. The electron follows a circular path until it leaves the field.



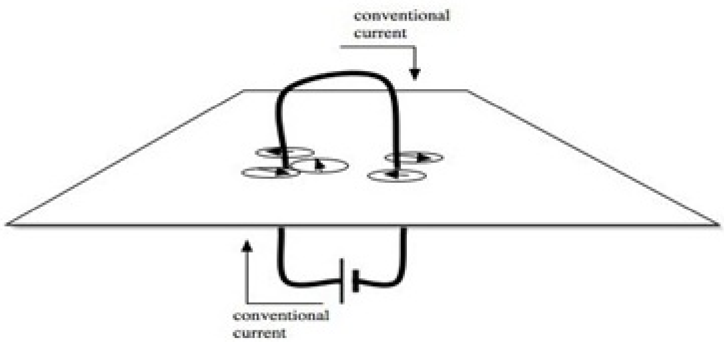
answer:

The moment the electron enters the field, it experiences a force perpendicular to its velocity. The electron follows a circular path until it leaves the field.

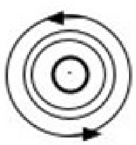
FIELD AROUND A CURRENT-CARRYING CONDUCTOR

When current flows through a conductor a magnetic field forms. The field lines form concentric circles around the conductor. Hold a straight conductor (wire), in your right hand with your thumb pointing in the direction of conventional current (positive flow). Your fingers circle the wire in the direction of the magnetic field.

The compasses in the following diagram indicate the direction of the magnetic field near the conductor. Use your left hand for electron flow.

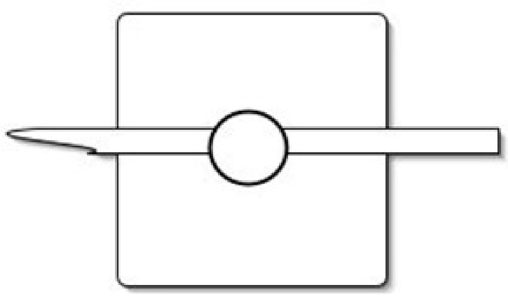


The following diagram shows a conductor carrying conventional current out of the page (toward the observer), and the direction of the field near the conductor. With the thumb of the right hand pointing toward the observer, the fingers of the right hand circle around the conductor in the direction of the magnetic field.



EXAMPLE

3. The following diagram is a view from above a table with a current-carrying conductor stretched across the table top. Conventional current is flowing from left to right. A compass is laying flat on top of the conductor. What is the direction of the compass needle?

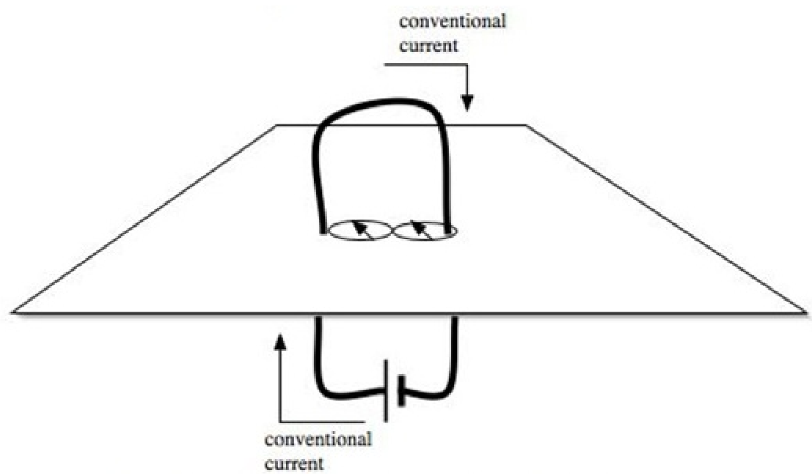


answer:

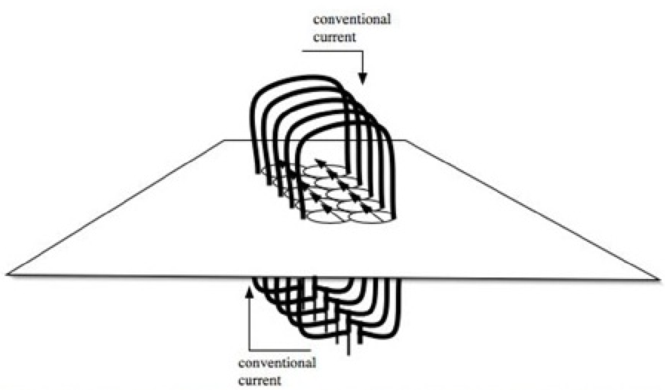
To the bottom of the page.

SOLENOIDS

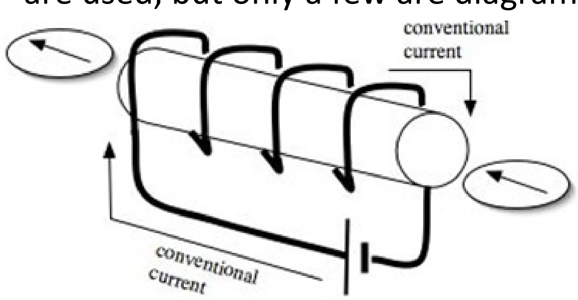
Notice the fields formed by either side of a looped conductor reinforce each other inside the loop.



One might imagine carrying a step further using many loops to produce a stronger magnetic field.



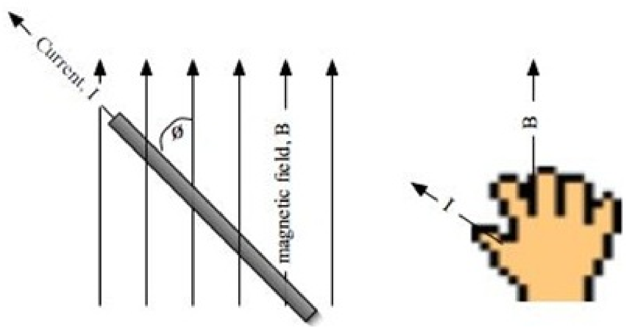
In practice, this is achieved by looping a wire many times and attaching it to one power source. This is called a "**solenoid**." An iron core is often used to help gather magnetic field lines into a more intense field. Hundreds or thousands of loops or "turns" are used, but only a few are diagrammed for clarity.



Notice that by wrapping the fingers of your right hand around the solenoid in the direction of conventional current flow, the thumb points north. For electron flow use your left hand.

MAGNETIC FORCE ON A CURRENT-CARRYING CONDUCTOR

When a conductor carries current through a magnetic field, a **magnetic force** is produced on the conductor. (**This is known as the motor principle**). Use the right to determine the direction of magnetic force on a current-carrying conductor in a magnetic field. The thumb points in the direction of conventional (positive) current. The fingers point in the direction of the field (from north to south). The palm points in the direction of the force on the conductor. In the diagram below, the force on the conductor is into the page (or away from the observer). **For current defined as electron flow**, use the left hand.



The magnitude of the force on the conductor, F, is

F = BILsinø where L is the length of the conductor in the field.

Notice that F = 0 when any of the following are true:
B= 0 ; I = 0; L = 0; ø = 0^0; ø = 180^0 .

Notice that F is greatest when the conductor is at right angles to the field.

Electric Current is the rate of flow of charges. An increase in current through a conductor means more number of electrons are crossing any cross section of the wire per second.

A moving charge experiences a force in magnetic field because a moving charge creates a magnetic field around it. The two interacting magnetic fields cause the force..

EXAMPLES

4. A horizontal conductor is carrying 5.0 A of current to the east. A magnetic field of 0.20 T pointing straight up cuts across 1.5 m of the conductor. What is the force acting on the conductor?

answer:

With the fingers of the right hand pointing straight up, and the thumb pointing east, the palm points south.

$F = BIL\sin\emptyset = (0.20\ T)(5.0\ A)(1.5\ m)\sin 90^0 = 1.5\ N$.

5. A 50.0 cm horizontal section of conductor with a mass of 8.00 g is in a 0.400 T magnetic field directed to the west. What is the magnitude and direction of current required to make this section of the conductor seem weightless?

answer:

The magnetic force must be opposite and equal to the weight of the section of the conductor.

With the fingers of the right hand pointing west, and the palm facing straight up, the thumb points north.

The weight of the conductor is $mg = (0.00800\ kg)(9.8\ N/kg) = 0.0784\ N$.

The magnetic force on the conductor is

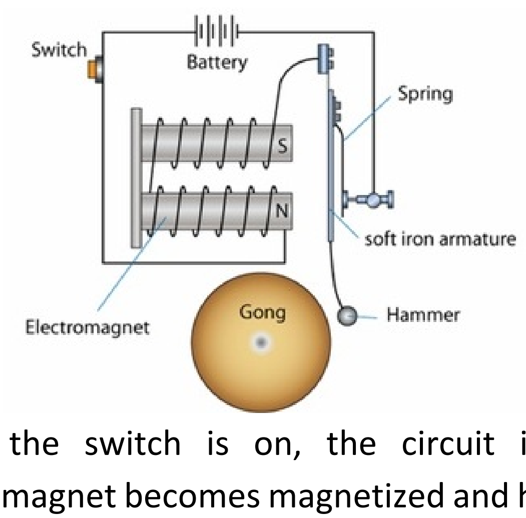
$F = BIL\sin\emptyset$, so

$0.0784\ N = (0.400\ T)I(0.50\ m)\sin 90^0$

$I = 0.392\ A$ [N]

Applications and Uses of Electromagnet

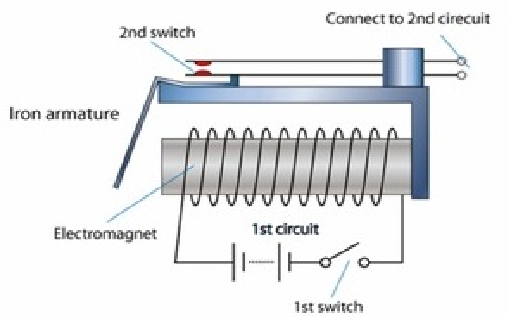
Electric Bell



When the switch is on, the circuit is completed and current flows. The electromagnet becomes magnetized and hence attracts the soft-iron armature and at the same time pull the hammer to strike the gong.

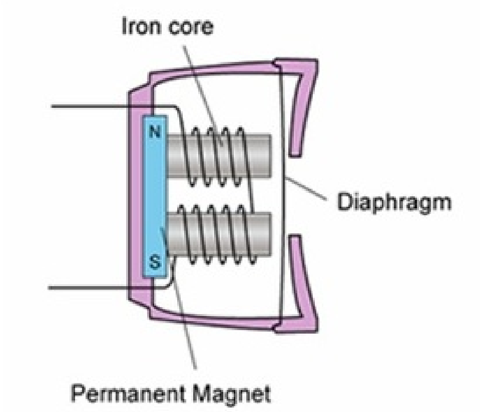
As soon as the hammer moves towards the gong, the circuit is broken. The current stops flowing and the electromagnet loses its magnetism. This causes the spring to pull back the armature and reconnect the circuit again.
When the circuit is connected, the electromagnet regain its magnetism and pull the armature and hence the hammer to strike the gong again. This cycle repeats and the bell rings continuously.

Electromagnetic Relay



A relay is an electrical switch that opens and closes under the control of another electrical circuit.
A relay has at least two circuits. One circuit can be used to control another circuit.
The 1st circuit (input circuit) supplies current to the electromagnet. The electromagnet is magnetized and attracts one end of the iron armature.
The armature is then closes the contacts (2nd switch) and allows current flows in the second circuit. When the 1st switch is open again, the current to the electromagnet is cut, the electromagnet loses its magnetism and the 2nd switch is opened. Thus, current stops to flow in the 2nd circuit.

Telephone Earpiece



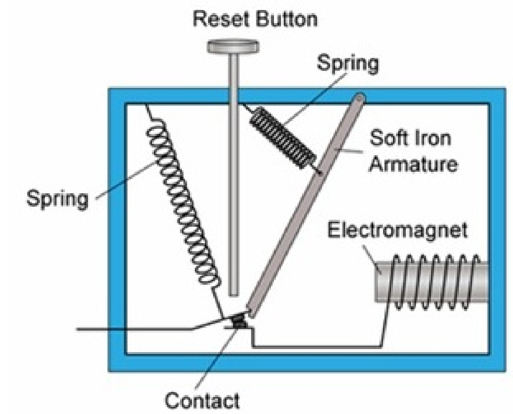
An electromagnet is used in the earpiece of a telephone. The figure shows the simple structure of a telephone earpiece.

When you speak to a friend through the telephone, your sound will be converted into electric current by the mouthpiece of the telephone. The current produced is a varying current and the frequency of the current will be the same as the frequency of your sound. The current will be sent to the earpiece of the telephone of your friend.

When the current passes through the solenoid, the iron core is magnetised. The strength of the magnetic field changes according to the varying current. When the current is high, the magnetic field will become stronger and when the current is low, the magnetic field become weaker.

The soft-iron diaphragm is pulled by the electromagnet and vibrates at the frequency of the varying current. The air around the diaphragm is stretched and compressed and produces sound wave. The frequency of the sound produced in the telephone earpiece will be the same as your sound.

Circuit Breaker



The figure shows the structure of a circuit breaker. A circuit breaker is an automatic switch that cut off current in a circuit when the current become too large.
When the current in a circuit increases, the strength of the electromagnet will increase in accordance; this will pull the soft iron armature towards the electromagnet.
As a result, the spring pulls apart the contact and disconnects the circuit immediately, and the current stop to flow.
We can reconnect the circuit by using the reset button. The reset button can be pushed to bring the contact back to its original position to reconnect the circuit.

Questions

1. Find the magnetic flux density in the center of a 4.0 cm long air-core solenoid made with 4900 turns of wire and carrying a 2.5A current.
A. 0.385 T B. 0.78 T C. 0.298 T D. 3.850 T
2. Calculate the force on the wire, when it makes an angle of 30^0in the direction of a straight wire of 40cm long and carrying a current of 4A is placed in a uniform magnetic field of strength 3×10^{-2} weber/m^2.
A. 3.4×10^{-2}N B. 2.4×10^{-2}N C. 4.5×10^{-2}N D. 5×10^{-2}N
3. When the current in a magnetic field increases, the strength of the magnetic field
A. Decreases B. Increases C. Stabilizes D. Extincts
4. Electromagnet is applicable in all, except
A. Telephone earpiece B. Electric bells C. Electromagnetic relays D. Ammeter
5. In the Flemings left hand rule, the fore-finger points in the

A. Direction of the magnetic field B. Direction of the conventional current C. direction of the force on the conductor D. None of the above is correct.

Answers

1. A 2. B 3. B 4. D 5. A

WEEK 9

First Term SSS3, Physics,

Topic**: ELECTROMAGNETIC FIELD**

Electromagnetic field

Electromagnetic field is a field representing the joint interaction of electric and magnetic field.

Electromagnetic field exerts force on charged particles. The force on a charged q moving with a velocity v (less than the velocity of light) is given by

$F = q(E + v \times B)$

The field can be viewed as the combination of an electric field and a magnetic field. The electric field is produced by stationary charges, and the magnetic field by moving charges (currents); these two are often described as the sources of the field. The way in which charges and currents interact with the electromagnetic field is described by Maxwell's equations and the Lorentz force law.

Electromagnetic field; a property of space caused by the motion of an electric charge. A stationary charge will produce only an electric field in the surrounding space. If the charge is moving, a magnetic field is also produced. An electric field can be produced also by a changing magnetic field. The mutual interaction of electric and magnetic fields produces an electromagnetic field, which is considered as having its own existence in space apart from the charges or currents (a stream of moving charges) with which it may be related. Under certain circumstances, this electromagnetic field can be described as a wave transporting electromagnetic energy.

Fleming's left hand rule

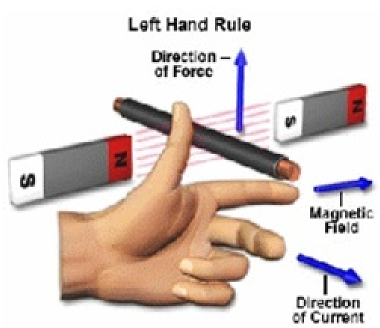
If an electric current flows through a magnetic field, the field exerts a force on the current.

This can happen when there is a magnet close to a wire carrying a current. It can also happen when there is a magnet close to a beam of electrons. The force causes the wire or the beam to move.

The direction of this force depends on the directions of the field and the current. The way to find this direction is called Fleming's left hand rule. Stretch out the thumb and first two fingers of your left hand. Hold them out so that they are all perpendicular to one another. Your thumb points in the direction of the force, your first finger points in the direction of the field, and your middle finger points in the direction of the current.

To find the direction of the force, point your first finger in the direction of the field and your middle finger in the direction of the current. Then your thumb will tell you the direction of the force. (When you do this, your thumb and two fingers MUST stay perpendicular to one another.)

It is well know that a current carrying conductor in a magnetic field experiences a force, and that if a conducting loop of wire moves relative to a magnetic field, a current is generated. Fleming's Left Hand Rule enables us to calculate the direction of the force in the first case and the direction of the field in the second case. This is illustrated below.



In this diagram the two fingers are horizontal at right angles and the thumb is vertical and at right angles to both. In fact, first finger, second finger and thumb are all at right angles to each other.

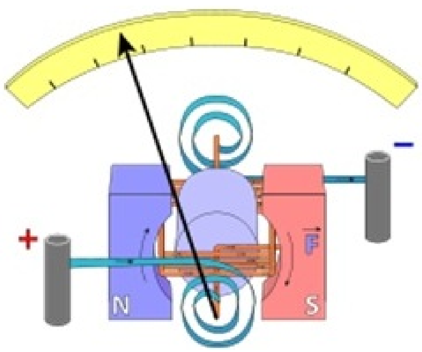
Application of Electromagnetic Field

Direct Current Motors (DC Motors)

The purpose of a direct current motor is to transform electrical energy into mechanical energy. When a current is supplied to a coil with parallel plane to the magnetic field, the coil turns only when the current passes through it. A permanent magnet with North and South poles place either side with a rectangular coil placed in between the two permanent magnets with different poles placed side by side to

other , i.e. N-S. Split rings are connected and are insulated from one another to the end of the coil, which are soldered to two brass segment communicators. Two carbon brushes are used to press lightly on the two segments. Battery is connected to the current in order to give current to the coil.
A downward force acts on the coil at the North pole and upward force acts on the coil at the South pole.

Moving Coil Galvanometer



This galvanometer is one of the most sensitive and accurate methods for detecting or measuring extremely small currents or potential difference
Galvanometer, is an electromechanical instrument used to indicate the presence, direction, or strength of a small electric current. The typical galvanometer is a sensitive laboratory instrument used to detect and compare currents, but the primary purpose of galvanometer is the detection of electric current not the measurement of current.
The galvanometer makes use of the fact that an electric current flowing through a wire sets up a magnetic field around the wire. In the galvanometer, the wire is wound into a coil. When current flows through the coil, one end of the coil becomes a north magnetic pole, the other a south magnetic pole. When a permanent magnet is placed near the coil, the two fields—the one from the coil and the one from the magnet—interact. The like poles will repulse each other and the unlike poles will attract. The amount of attraction and repulsion increases as the strength of the current increases.
In the moving-magnet galvanometer, the permanent magnet is a needle (much like a compass needle) mounted on a pivot and surrounded by the coil. In the moving-coil galvanometer—the most common type—the coil is mounted on pivots or

suspended by thin metal strips. The coil lies between the poles of a permanent magnet in such a way that it rotates when current flows through it. The direction of the rotation depends on the direction of the current through the coil, and the amount of rotation depends on the strength of the current. A galvanometer is often used to indicate when the current in a circuit has been reduced to zero, as in the operation of the Wheatstone bridge, a device for measuring electrical resistances precisely.

A moving-coil mechanism similar to that used in a galvanometer is used in some ammeters. Like the galvanometer, these instruments measure the strength of a current but they can handle a stronger current; unlike the galvanometer, they cannot indicate the current's direction. A moving-coil mechanism is also used in some voltmeters (which measure the voltage in a circuit) and ohmmeters (which measure the resistance in a circuit). In some instruments, a selector switch connects the moving-coil mechanism to different internal circuits so that a single mechanism can be used in making all three types of measurements.

The principle upon which the operation of the galvanometer is based was discovered in 1820 by Hans Christian Oersted when he observed that a magnetic needle could be deflected by an electric current. The first galvanometer was made by Johann Schweigger in 1820. In 1882, Jacques Arsene D'Arsonval introduced the moving-coil galvanometer. Edward Weston made important improvements to the device a few years later.

Advantages of moving coil galvanometer

1. It has a linear scale
2. Extreme magnetic field has no effect on it.
3. It can be used to measure current, voltage or resistance.

Sensitivity of a moving coil galvanometer

The deflecting couple on the coil is equal to the force on the vertical side multiplied by the width of the coil.

The force on a vertical side will be proportional to its length, hence the deflecting couple will be proportional to length x width which is equal to the area of the coil.

Thus, the sensitivity of a galvanometer is defined as the scale deflection by microampere. For high sensitivity the galvanometer must have:

(i) a coil of large area

(ii) a large number of turns in the coil

(iii) a special alloy permanent magnet which gives high magnetic flux

(iv) weak hair springs to give a small control couple.

1. A galvanometer as an electromechanical device is majorly used to detect ……………………………… amount of electric current.

A. Large B. Small C. Extra large D. infinite

2. In Fleming's left hand rule, the middle finger indicates the direction

A. Field B. Force B. Voltage D. Current

3. The purpose of a direct current motor is to transform ……………………… energy into …………………………… energy

A. Mechanical, Electrical B. Electrical, Heat C. Mechanical, Heat D. Electrical, mechanical

4. The first galvanometer was made by Johann Schweigger in

A. 1821 B. 1824 C. 1820 D. 1902

5. The positioning of the fore-finger in the Fleming's left hand rule is to determine the direction of the

A. Current B. Motion C. Force D. Field

Answers

1. B 2. D 3. D 4. C 5. D

<u>WEEK 10</u>

First Term SSS3, Physics

Topic: ELECTROMAGNETIC INDUCTION

INDUCED CURRENT

Definition: Electromagnetic induction (or sometimes just *induction*) is a process where a conductor placed in a changing magnetic field (or a conductor moving through a stationary magnetic field) causes the production of a voltage across the conductor. This process of electromagnetic induction, in turn, causes an electrical current – it is said to *induce* the current.

Michael Faraday is given credit for the discovery of electromagnetic induction in 1831, though some others had noted similar behavior in the years prior to this. The formal name for the physics equation that defines the behavior of an induced electromagnetic field from the magnetic flux (change in a magnetic field) is Faraday's law of electromagnetic induction.

If a coil of wire is placed in a changing magnetic field, a current will be induced in the wire. This current flows because something is producing an electric field that forces the charges around the wire. (It cannot be the magnetic force since the charges are not initially moving). This "something" is called an **electromotive force**, or **emf**, even though it is not a force. Instead, emf is like the voltage provided by a battery. A changing magnetic field through a coil of wire therefore must *induce* an emf in the coil which in turn causes current to flow.

The law describing induced emf is named after the British scientist Michael Faraday, but Faraday's Law should really be called Henry's Law. Joseph Henry, an American from the Albany area, discovered that changing magnetic fields induced current before Faraday did. Unfortunately, he lived in the age before instantaneous electronic communication between Europe and America. Faraday got published and got famous before Henry could report his findings. Interestingly enough, Henry had to explain the results to Faraday when the two met a few years later.

Briefly stated, Faraday's law says that a changing magnetic field produces an electric field. If charges are free to move, the electric field will cause an emf and a current. For example, if a loop of wire is placed in a magnetic field so that the field passes through the loop, a change in the magnetic field will induce a current in the loop of wire. A current is also induced if the area of the loop changes, or if the area enclosing magnetic field changes. So it is the change in **magnetic flux**, defined as

$\phi_B = \int B.dA = BA \cos q$

that determines the induced current. **A** is the area vector; its magnitude is the area of the loop, and its direction is perpendicular to the area of the loop, and θ is the angle between **A** and the magnetic field **B**. The last equality (removing the integral) is valid only if the field is uniform over the entire loop.

Faraday's Law says that the emf induced (and therefore the current induced) in the loop is proportional to the rate of change in magnetic flux:

$e = iR = -\Delta\phi_B/\Delta t$

is the emf, which is the work done moving charges around the loop, divided by the charge. It is similar in concept to voltage, except that no charge separation is necessary. The magnetic flux Φ_B equals the magnetic field *B* times the area *A* of the loop with magnetic field through it if (a) the magnetic field is perpendicular to the plane of the loop, and (b) the magnetic field is uniform throughout the loop. For our purposes, we will assume these two conditions are met; in practical applications, however, magnetic field will vary through a loop, and the field will not always be perpendicular to the loop.

Since all applications of Faraday's Law to magnetic storage involve a coil of wire of fixed area, we will also assume that (c) the area does not change in time. We then have a simpler expression for the current induced in the coil:

$iR = -A\, dB/dt$

The induced current depends on both the area of the coil and the change in magnetic field. In a coil of wires, each loop contributes an area *A* to the right-hand side of the equation, so the induced emf will be proportional to the number of loops in a coil. But doubling the number of loops doubles the length of wire used and so doubles the resistance, so the induced current will not increase when loops are added.

Laws of Electromagnetic Induction

Faraday's Law

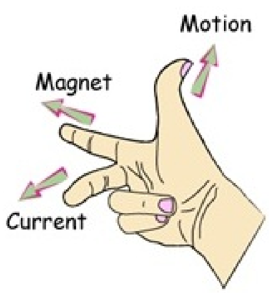
The magnitude of the induced e.m.f is determined from Faraday's Law.Faraday's Law states that the magnitude of the induced e.m.f is directly proportional to the rate of change of magnetic flux through a coil or alternatively the rate of the magnetic flux being cut.

When a magnet is moved into and out of a coil, the induced current that flows through the coil can be determined from Lenz's Law. Lenz's Law states that the induced current always flows in the direction that opposes the change in magnetic flux.

Lenz's Law

Lenz's Law obeys the principle of conservation of energy. Work is done to move the magnet against the repulsive force. This work done is converted to electric energy which manifests as an induced current.

For a conductor in a closed circuit moving perpendicular to a magnetic field and hence cutting its magnetic flux, the direction of the induced current is determined from Fleming's Right-Hand Rule.



Fleming's Right-Hand Rule is used to determine the direction of the induced current that flows from the wire when there is relative motion with respect to the magnetic field.

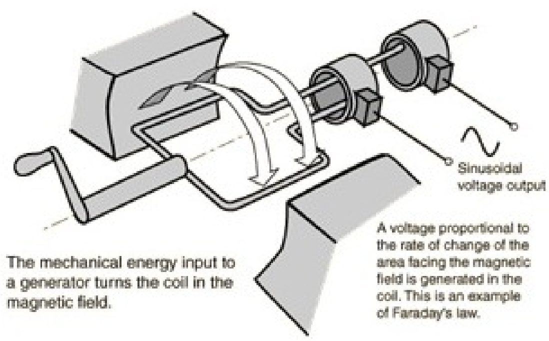
Induction Coil

An induction coil consists of two coils of insulated copper wire wound around a common iron core. One coil, called the primary winding, is made from relatively few (tens or hundreds) turns of coarse wire. The other coil, the secondary winding, typically consists of many (thousands) turns of fine wire.

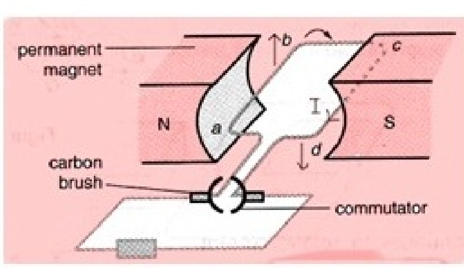
An electric current is passed through the primary, creating a magnetic field. Because of the common core, most of the primary's magnetic field couples with the secondary winding. The primary behaves as an inductor, storing energy in the associated magnetic field. When the primary current is suddenly interrupted, the magnetic field rapidly collapses. This causes a high voltage pulse to be developed across the secondary terminals through electromagnetic induction. Because of the large number of turns in the secondary coil, the secondary voltage pulse is typically many thousands of volts. This voltage is often sufficient to cause an electric spark, to jump across an air gap separating the secondary's output terminals. For this reason, induction coils were called spark coils.

AC Generator

The turning of a coil in a magnetic field produces motional emfs in both sides of the coil which add. Since the component of the velocity perpendicular to the magnetic field changes sinusoidally with the rotation, the generated voltage is sinusoidal or AC. This process can be described in terms of Faraday's law when you see that the rotation of the coil continually changes the magnetic flux through the coil and therefore generates a voltage.



DC generator

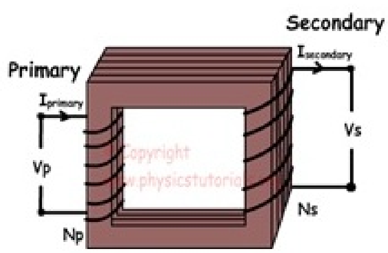


DC generator consists of 4 basic parts:
Magnet, coil, carbon brush & commutator
Basic working principle:
When the coil is rotated, side ab move upward, side cd moves downward.
Both sides cut the magnetic field
Induced current is produced, which then flows in the coil to external circuit
Commutator ensure the current flows in one direction.

Transformers

A transformer is an apparatus for changing a given electrical current into another current of different voltage. There are two kinds of transformers: step down and step up. Step up transformers increase the voltage where step down transformers decrease the voltage.

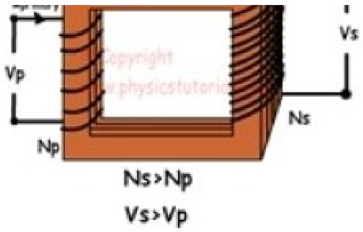
Structure of simple transformer is given below;



Voltage is applied to primary coil and, we take transformed voltage from secondary coil. There are two types of transformer, step up and step down. We use step down transformers in electrical devices like radio, and step up transformers in welding machine.

Step Up Transformer

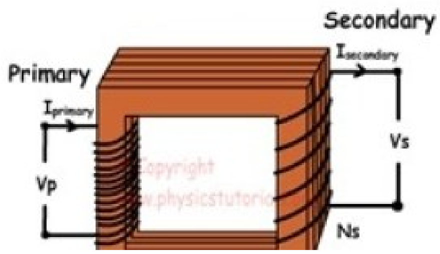
This type of transformer used for increase the incident voltage. Number of turns in secondary coil is larger than the number of turns in primary coil.



The N_s/N_p is called the turns ratio of the transformer. It is more than 1 for a step-up transformer and less than 1 for a step-down transformer.

Step Down Transformer

This type of transformer used for decrease incident voltage. Number of turns in primary coil is larger than the number of turns in secondary coil.



Transformer Equations

Vp is the potential, Ip is the current, Np is the turn on the primary coil and Vs is the potential, Is is the current, Ns is the turn on the secondary coil. We use following equations to find potential, current or number of turns of any coil;

$$N_p/N_s=V_p/V_s=I_s/I_p$$

Mutual Inductance is the flow of induced current or voltage in a coil due to an alternating or varying current in a neighbouring.

Eddy Currents

Such currents produced by the varying flux cutting the iron core, reduce efficiency because they consume power form the primary. Energy is lost in the form of heat in the iron core due to eddy current. Such core can be reduced by laminating the core. Laminations of core reduce eddy currents by breaking up their path of flow.

Power Transmission

Power generated at power stations are distributed over large distances to consumers through metal cables. Power is given by P = IV, and can thus be transmitted either at low current and at high voltage or at high current and low voltage.

Because the metal cables through which the power is transmitted have a certain amount of electrical resistance, transmitting power at high current will lead to loss of energy in the form of heat in the cables ($H \alpha I^2R$). To avoid this, power is therefore transmitted at high voltage and low current. This is known as high tension transmission.

Low currents lead to low energy loss in form of I^2R. Also since lower currents require thinner cables, cost of cable materials is considerably reduced if power is transmitted with low currents and high voltage.

Step down transformers are used to reduce the high transmitted voltages to lower voltages required in homes and factories.

Question

1. Which of the following statements about a generator is not correct?

A. It can produce direct current

B. It can produce alternating current

C. It requires an external supply of current to the coil.

D. It may require the use of a commutator

2. A transformer has 500 turns in the primary coil and 300 turns in the secondary coil. If the primary coil is connected to a 220 V mains, what voltage will be obtained from the secondary coil?

A. 123 volts B. 132 volts C. 150 volts D. 230 volts

3. DC generator consists of 4 basic parts namely, except

A. Magnet B. Coil C. Carbon brush D. Alternator E. Commutator

4. A transformer with 5500 turns in its primary is used between a 240 V a.c. supply and a 120 V kettle. Calculate the number of turns in the secondary.

A. 11,000 B. 275 C. 460 D. 232

5. Which of these is correct about the step-up transformer?

A. Primary turns is greater than the secondary B. Secondary turns is greater than the primary turns C. Secondary turns is equal to the primary turns D. All of the above.

Answer

1. C 2. B 3. D 4. B 5. B

WEEK 11

First Term SSS3, Physics,

Topic: **GRAVITATIONAL FIELD AND LAW, GRAVITATIONAL POTENTIAL, ESCAPE VELOCITY , POTENTIAL ENERGY IN GRAVITATIONAL FIELD**

Concept of Gravitational Field

There are two types of forces – contact forces and force fields. Contact forces are most common in everyday life. For example, you push or pull on wheelbarrow, a tennis racket exerts a force on a tennis ball when they make contact, your foot exerts a force on a football when you kick it. Force fields, e.g. gravitational forces, act even when the two bodies are not in contact. The earth, for example exerts a force on a falling mango fruit. It also exerts a force on a on the moon which is about 385,000km away. The sun itself exerts a force on the earth even though the earth is about 1.5×10^8 km distant from the sun.

In order to explain the observation of forces acting at a distance, it has been postulated that a gravitational field surrounds everybody that has mass, and this field fills up all of space. A second body at a particular location near the first body experiences a force because of the gravitational field that exists there. The gravitational field of the first body acts directly on the second body.

Gravitational field is a region or space around a mass in which the gravitational force of the mass can be felt.

Gravitational Forces between two masses

Gravitation is the force of attraction is the force of attraction exerted by a body on all other bodies in the universe. Hence a gravitational force exists between a body and all other bodies around it. Gravitational forces act between all masses and hold together planets. stars and galaxies, each mass has a gravitational field around it.

It was Sir Isaac Newton who first proposed the relationship between the gravitational force F, between two masses, m_1 m_2 and the distance, r, between these masses. He proposed his famous *Law of Universal gravitation* which we can state as follows:

Every particle in the universe attracts every other particle with a force that is proportional to the product of their masses and inversely proportional to the square of the distance between them. This force acts along the line joining the centre of the two particles.

F µ m_1 m_2/r^2

The magnitude of this force can be written as

$F = Gm_1m_2/r^2$

Where m_1 and m_2 are the masses of the two particles, r is the distance between them and G is a universal constant of gravitation or simply gravitational constant.

Gravitational constant, G, has the same numerical value for all objects ($G = 6.67 \times 10^{-11}\ Nm^2kg^{-2}$).

Newton's law of gravitation refers to the force between two particles or bodies. Actually such a gravitational force is a pair of forces, an action-reaction pair. Although the masses of the particles may be different, forces of equal magnitude act on each other, and the action line of both forces lies along the line joining the bodies. Mass m_1 attracts m_2 with a force given by the equation $F = Gm_1m_2/r^2$. By Newton's third law of motion, action and reaction are equal and opposite. Therefore, m_2 also attracts m_1 by an equal but opposite force.

Gravitational attraction keeps the moon in its orbit around the earth and the earth in its orbit around the sun. Gravitational forces are always those of attraction.

Example

Two 5.0 kg spherical balls are placed so that their centres are 50.0 cm apart. What is the magnitude of the gravitational force between the two balls? ($G = 6.67 \times 10^{-11}\ Nm^2kg^{-2}$)

Solution

$F = Gm_1m_2/r^2$

$= 6.67 \times 10^{-11} \times 5.0 \times 5.0/0.5^2$

$= 6.67 \times 25 \times 10^{-11}/0.5^2$

$= 6.67 \times 10^{-9}$ N

Because of the universal law of gravitation, there is a gravitational force of attraction between the sun and the planets, between earth and the moon, and also between other planets.

Relation between the Gravitational Constant 'G' and the acceleration of gravity at the earth surface 'g'

The earth is supposed to be a sphere of radius, r_e, with its mass m_e concentrated at the earth's centre. The distance of any object on the earth's surface to the centre of the earth is r_e the earth's radius. The gravitational force of attraction of the earth on any mass, m, on the earth's surface is given by

$F = Gm_e m/r_e^2$

This is the force of gravity on the mass due to the earth, that is, the weight of the object, mg, where 'g' is the acceleration due to gravity.

Thus, $F = Gm_e m/r_e^2 = mg$

The force per unit mass, F/m, is given by

$F/m = Ge_{ms}/r_e^2 = g$

Hence $g = Gm_e/r_e^2$

This means that the acceleration due to gravity 'g' can be considered as the force per unit mass on the earth's surface. According to this equation $g = Gm_e/r_e^2$, the acceleration of gravity, 'g' at the surface of the earth, is determined by m_e (the earth's mass), and r_e (the earth's radius), hence from equation $g = Gm_e/r_e^2$ we should 'g' to be slightly greater at the top of the mountain than at the sea level. This is what actually obtains in practice.

If 'g' and 'G' are actually known we can use equation $g = Gm_e/r_e^2$ to calculate the earth's mass, m_e.

$M_e = gr_e^2/G$.

The radius of the moon is one-fourth, and its mass is one eighty-first that of the earth. If the acceleration due to gravity on the surface of the earth is 9.8 ms^{-2}, what is its value on the moon's surface.

The relation between g and G is given by

$g = Gm_e/r_e^2$

For the earth, $g_e = Gm_e/r_e^2$ (i)

For the moon, $g_m = Gm_m/r_m^2$ (ii)

Since G is the universal constant of gravitation it is constant for both equations.

$r_m = r_e/4$ (iii)

$m_m = m_e/81$ (iv)

From equation (i)

$G = g_e r_e^2/m_e$ (v)

Putting (v) in (ii) we have

$g_m = g_e r_e^2/m_e \,.\, m_m/r_m^2$

$= 9.8 \times (r_e/ r_m)^2 \times m_m/m_e$ (vi)

From (iii) $r_e/r_m = 4$

From (iv) $m_m/m_e = 1/81$

Putting these in equation (vi) we have: $g_m = 9.8 \times (4)^2 \times 1/81$

$= 9.8 \times 16 \times 1/81 = 1.9 \text{ ms}^{-2}$

Gravitational Potential

The work done in raising a mass, m from the ground surface to a height, h above the ground is given by

$W = mgh$

The work has been done against the gravitational pull of the earth. This work appears as the gravitational potential energy (PE) of the body.

This potential energy is dependent of the height, h, or the relative position of the body from the ground or zero level where the PE is considered to be zero.

In general points in any gravitational field possess *gravitational potential*. If free to move, a body will tend to move from a point ay higher gravitational potential to points of lower gravitational potential

Gravitational Potential (G) at a point is defined as the work done in taking unit mass from infinity to that point. Unit is jkg^{-1}.

This gravitational potential is given by

$V = -Gm/r$

Where m is the mass producing the gravitational field and r is the distance of the point to the mass. The gravitational potential decreases as r increases and become zero where r is infinitely large, the negative sign in the equation above indicates that the potential at infinity (zero) is higher that the potential close to the mass.

Escape Velocity

There are many man-made satellites that circle around the earth at the present time. One common feature of these bodies is that they are held in an approximately circular path by the earth's gravitational pull. It is this force that provides the needed earth's centripetal force required to keep the satellites in their orbits. The velocity (*v_s*) of the satellite as it orbits round the earth is given by

$Mv_s^2/r_e = F = Gmm_e/r_e^2$ (centripetal force) = (Gravitational force)

Hence $v_s = \text{Ö}Gm_e/r_e$

This is the velocity with which the satellite moves round the earth. Notice that the mass of the satellite does not enter into this $v_s = \text{Ö}Gm_e/r_{e.}$ All satellites in orbit with radius r_e must have the same speed. For a satellite to escape from the earth and never return, it must be launched with a velocity greater than that required to make it orbit.

We define the escape velocity (*v_e*) as the minimum velocity required for an object (e.g. satellite or rocket) to just escape or leave the gravitational influence or field of an astronomical body (e.g. the earth) permanently.

We can obtain the formula for the Escape velocity using the Newton's law of universal gravitation which is an inverse square law:

$F = Gm_em/r^2$

Let m be the mass of the satellite, and m_e, the mass of the earth.

The work done in carrying a mass m from a point at a distance r from the centre of the earth, to a distance so great that the gravitational field is negligibly weak is given by

$W = F \times r$

But from equation 2.2, $F = Gm_em/r^2$

Hence, $W = Gm_em/r^2.r = Gm_em/r$

This work must equal the kinetic energy of the body of mass m at this point, having a velocity v_e. This kinetic energy is given by:

$KE = ½\, mv_e^2$

$½\, mv_e^2 = Gm_em/r$

$v_e^2 = 2Gm_e/r$

$v_e = \text{Ö}2Gm_e/r$

If we launch the mass m from the earth's surface, where r = R, we then have that

$v_e = \text{Ö}2Gm_e/R$

But from equation $m_e = gr_e^2/G$, $m_e = gr^2/G$

Hence $v_e = \text{Ö}2G/R.gR^2/G$

v_e = Ö2gR

Where R is the earth's radius

Questions

1. Determine the mass of the earth if the radius of the earth is approximately 6.38 x 10^6 m, G = 6.67 x 10^{-11} Nm^2kg^{-2} and g = 9.80 ms^{-2}.

A. 5 x 10^{24} kg B. 5.98 x 10^{25} C. 5.98 x 10^{24} D. 5.6 x 10^{25}

2. Determine the force of attraction between the sun (m_s = 1.99 x 10^{30} kg) and the earth (m_e = 5.98 x 10^{24} kg). Assume the sun is 1.50 x 10^8 km from the earth.

A. 3.53 x 10^{18} B. 3.53 x 10^{22} N C. 3.05 x 10^{32} D. 3.3 x 10^{22}

3. Which of the following is correct for Escape Velocity?

A. v_e = Ö2Gr B. v_e = Ö2Gr C. v_e = Ö2gr D. v_e = Ö2gR

4. The numerical value for gravitational constant G is

A. 6.42 x 10^{-11} Nm^2 kg^2 B. 6.67 x 10^{-11} Nm^2 kg^2 C. 7.8 x 10^{-11} Nm^2 kg^2 D. 6.67 x 10^{-15} Nm^2 kg^2

5. The expression F= Gm_1m_2/r^2 indicates

A. The gravitational force of attraction B. The gravitational force of repulsion C. Gravitational force of constant acceleration D. Force of repulsion

Answers

1. C 2. B 3. D 4. B 5. A

SS 3
SECOND TERM NOTES ON PHYSICS

TABLE OF CONTENT

SECOND TERM

WEEK 1

Physics SS 3

Topic: ALTERNATING CURRENT CIRCUIT

ALTERNATING CURRENT

A.C. circuits are circuits through which an alternating current flows. Such circuits are used extensively in power transmission, radio, telecommunication and medicine.

Alternating currents are produced by time dependent alternating voltages given by the relation $E = E0 \sin \omega t$. Much of what we learned about d.c. circuits also apply to a.c. circuits. The effects on such voltages on resistors, capacitors and inductors will be discussed.

Nomenclature in A.C Circuits

An Alternating Current (A.C.) is one that varies sinusoidally or periodically, in such a way as to reverse its direction periodically. The commonest form of such an a.c. can be represented by

$I = I0 \sin 2\pi ft$

$I0 \sin \omega t$

Where I is the instantaneous current at a time t, I0 is the maximum (or peak) value of current or its amplitude; f is the frequency and ω (= $2\pi ft$) is the angular velocity, (ωt) is the phase angle of the current. Alternating is also represented by

$V = V0 \sin 2\pi ft$

$= V0 = \omega t$

Here, v, v0 are the instantaneous and peak (or maximum) values of the voltages or its amplitude

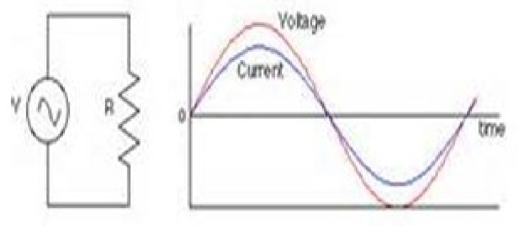
Example

If an a.c. voltage is represented by the relation $V = 4 \sin 900v\pi t$, the peak voltage V0 = 4

V and $2\pi ft = 900\pi t$ or $f = 900/2 = 450$ Hz. Then $\omega = 2\pi f = 900\pi$.

Peak, and r.m.s Values of A.C.

An alternating current (or voltage) varies sinusoidally as shown below which is a sine waveform. The amplitude or peak value of the current I0, is the maximum numerical value of the current.



The root mean square (r.m.s.) value of the current is the effective value of the current.

Root-mean-square current is that steady current which will develop the same quantity of heat in the same time in the same resistance.

The r.m.s. value for the current is given by

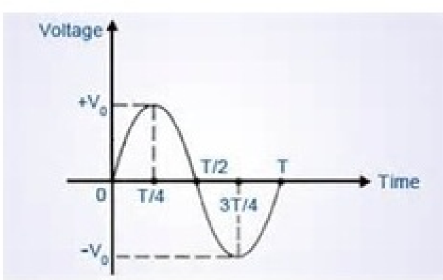
Ir.m.s =I0/√2

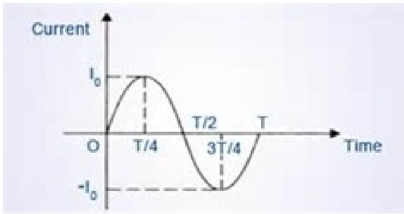
= 0.070I0

The moving iron and the hot-wire meters measure the average value of the square of the current called the mean square current. They are however calibrated in such a way as to indicate the r.m.s. current directly. Thus most a.c. meters read the effective or r.m.s. values. The average value of an a.c. voltage or current is zero.

Resistance in A.C Circuit

At any instance the current through the resistor (R) is I and the voltage across it is V



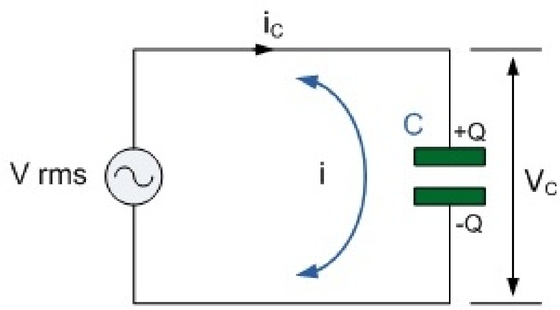


From Ohm's law we have that V = IR
Thus the current is given by I = V/R
If we put V = V0 sin ωt, then the current is also given by
I =V/R = V0sin ωt/R
= I0 sin ωt
The voltmeter and ammeter connected in the circuit will read the r.m.s. value of voltage and current.
Hence we can also write that Ir.m.s. = Vr.m.s. /R
The voltage and the current are said to be in phase or in step with each other. This means that both of them attain their maximum, zero and minimum values at the same instant in time.
Example
Find the root mean square value of the sinusoidal voltage with peak value at 260V.
Solution
Using V0 = √2 x Vr.m.s.
Given that V0 = 260V, Vr.m.s. =?
V = √2Vr.m.s. ; 260 = √2Vr.m.s.
260 = 1.414Vr.m.s.
Vr.m.s = 260/1.414 = 183.867 = 184V

Capacitance in AC Circuits
When capacitors are connected across a direct current DC supply voltage they become charged to the value of the applied voltage, acting like temporary storage devices and maintain or hold this charge indefinitely as long as the supply voltage is present. During this charging process, a charging current, (i) will flow into the capacitor opposing any changes to the voltage at a rate that is equal to the rate of change of the electrical charge on the plates.

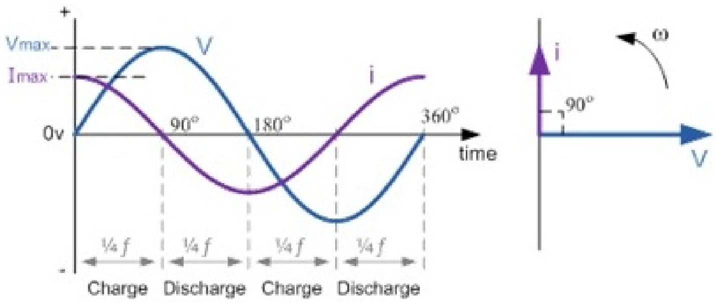
This charging current can be defined as: i = CdV/dt. Once the capacitor is "fully-charged" the capacitor blocks the flow of any more electrons onto its plates as they have become saturated. However, if we apply an alternating current or AC supply, the capacitor will alternately charge and discharge at a rate determined by the frequency of the supply. Then the Capacitance in AC circuits varies with frequency as the capacitor is being constantly charged and discharged.
We know that the flow of electrons through the capacitor is directly proportional to the rate of change of the voltage across the plates. Then, we can see that capacitors in AC circuits like to pass current when the voltage across its plates is constantly changing with respect to time such as in AC signals, but it does not like to pass current when the applied voltage is of a constant value such as in DC signals. Consider the circuit below.

AC Capacitor Circuit



In the purely capacitive circuit above, the capacitor is connected directly across the AC supply voltage. As the supply voltage increases and decreases, the capacitor charges and discharges with respect to this change. We know that the charging current is directly proportional to the rate of change of the voltage across the plates with this rate of change at its greatest as the supply voltage crosses over from its positive half cycle to its negative half cycle or vice versa at points, 0o and 180o along the sine wave. Consequently, the least voltage change occurs when the AC sine wave crosses over at its maximum or minimum peak voltage level, (Vm). At these positions in the cycle the maximum or minimum currents are flowing through the capacitor circuit and this is shown below.

AC Capacitor Phasor Diagram



Capacitive Reactance

Capacitive Reactance in a purely capacitive circuit is the opposition to current flow in AC circuits only. Like resistance, reactance is also measured in Ohm's but is given the symbol X to distinguish it from a purely resistive value. As reactance can also be applied to Inductors as well as Capacitors it is more commonly known as Capacitive Reactance for capacitors in AC circuits and is given the symbol Xc so we can actually say that Capacitive Reactance is Resistance that varies with frequency. Also, capacitive reactance depends on the value of the capacitor in Farads as well as the frequency of the AC waveform and the formula used to define capacitive reactance is given as:

Capacitive Reactance

$X_c = 1/2\pi fc = 1/\omega C$

Where:

F is in Hertz and C is in Farads.

$2\pi F$ can also be expressed collectively as the Greek letter Omega, ω to denote an angular frequency.

From the capacitive reactance formula above, it can be seen that if either of the Frequency or Capacitance where to be increased the overall capacitive reactance would decrease. As the frequency approaches infinity the capacitors reactance would reduce to zero acting like a perfect conductor. However, as the frequency approaches zero or DC, the capacitors reactance would increase up to infinity, acting like a very large resistance. This means then that capacitive reactance is

“Inversely proportional” to frequency for any given value of Capacitance and this shown below:

Question

1. Find the r.m.s. value of an alternating current whose peak value is 5 amps.

A. 3.53 Amps B. 4.5 Amps C. 2.19 Amps D. 6.50 Amps

2. Which of these is not correct, A.C. circuits are circuits through which an alternating current flows. Such circuits are used extensively in

A. Power transmission B. Radio C. Telecommunication D. Automobile

3. In an A.C. circuit the peak value of the potential difference is 180 V. What is the instantaneous p.d, when it has reached 1/8th of a cycle?

A. 80 Volts B. 90√2 Volts C. 80√2 Volts D. 90 Volts

4. A 240V supply is connected with a resistor of 20 in an A.C. circuit. Find the current in the circuit

A. 15 A B. 23 A C. 12 A D. 34 A

5. At frequency of 50Hz, a capacitor of 5µF is connected to a circuit of 230V supply. Find the capacitive reactance.

A. 630Ω B. 636.62Ω C. 790Ω D. 430.77Ω

Answer

1. A 2. D 3. B 4. C 5. B

WEEK 2

Physics SS 3

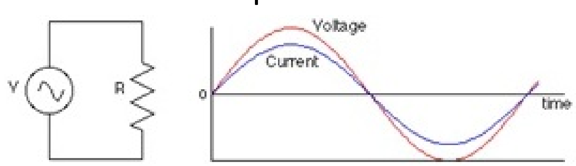
Topic: **AC IN RESISTOR, INDUCTOR AND CAPACITOR**

Alternating current

Direct current (DC) circuits involve current flowing in one direction. In alternating current (AC) circuits, instead of a constant voltage supplied by a battery, the voltage oscillates in a sine wave pattern, varying with time as $V = V_o \sin \omega t$.

In a household circuit, the frequency is 60 Hz. The angular frequency is related to the frequency, f, by $\omega = 2\pi f V_o$ represents the maximum voltage, which in a household circuit in North America is about 170 volts. We talk of a household voltage of 120 volts, though; this number is a kind of average value of the voltage. The particular averaging method used is something called root mean square (square the voltage to make everything positive, find the average, take the square root), or rms. Voltages and currents for AC circuits are generally expressed as r.m.s. values. For a sine wave, the relationship between the peak and the r.m.s. average is:

r.m.s. value = 0.707 peak value



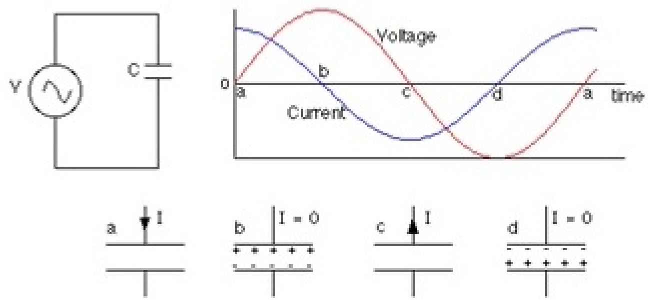
The relationship $V = IR$ applies for resistors in an AC circuit, so $I = V/R = (V_o/R)\sin(\omega t) = I_o \sin(\omega t)$

Resistance in an AC circuit

The relationship $V = IR$ applies for resistors in an AC circuit, so $I = V/R = (V_o/R) \sin(\omega t) = I_o \sin(\omega t)$

In AC circuits we'll talk a lot about the phase of the current relative to the voltage. In a circuit which only involves resistors, the current and voltage are in phase with each other, which means that the peak voltage is reached at the same instant as peak current. In circuits which have capacitors and inductors (coils) the phase relationships will be quite different.

Capacitance in an AC circuit



Consider now a circuit which has only a capacitor and an AC power source (such as a wall outlet). A capacitor is a device for storing charging. It turns out that there is a 90° phase difference between the current and voltage, with the current reaching its peak 90° (1/4 cycle) before the voltage reaches its peak. Put another way, the current leads the voltage by 90° in a purely capacitive circuit.

To understand why this is, we should review some of the relevant equations, including: relationship between voltage and charge for a capacitor:

CV = Q relationship between current and the flow of change: I = ΔQ/Δt

The AC power supply produces an oscillating voltage. We should follow the circuit through one cycle of the voltage to figure out what happens to the current.

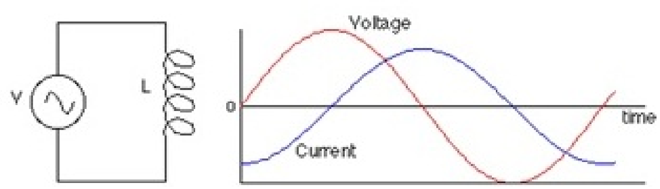
Step 1 – At point a (see diagram) the voltage is zero and the capacitor is uncharged. Initially, the voltage increases quickly. The voltage across the capacitor matches the power supply voltage, so the current is large to build up charge on the capacitor plates. The closer the voltage gets to its peak, the slower it changes, meaning less current has to flow. When the voltage reaches a peak at point b, the capacitor is fully charged and the current is momentarily zero.

Step 2 – After reaching a peak, the voltage starts dropping. The capacitor must discharge now, so the current reverses direction. When the voltage passes through zero at point c, it’s changing quite rapidly; to match this voltage the current must be large and negative.

Step 3 – Between points c and d, the voltage is negative. Charge builds up again on the capacitor plates, but the polarity is opposite to what it was in step one. Again the current is negative, and as the voltage reaches its negative peak at point d the current drops to zero.

Step 4 – After point d, the voltage heads toward zero and the capacitor must discharge. When the voltage reaches zero it's gone through a full cycle so it's back to point a again to repeat the cycle.
The larger the capacitance of the capacitor, the more charge has to flow to build up a particular voltage on the plates, and the higher the current will be. The higher the frequency of the voltage, the shorter the time available to change the voltage, so the larger the current has to be. The current, then, increases as the capacitance increases and as the frequency increases.
Usually this is thought of in terms of the effective resistance of the capacitor, which is known as the capacitive reactance, measured in ohms. There is an inverse relationship between current and resistance, so the capacitive reactance is inversely proportional to the capacitance and the frequency:
A capacitor in an AC circuit exhibits a kind of resistance called capacitive reactance, measured in ohms. This depends on the frequency of the AC voltage, and is given by Capacitive reactance $X_c = 1/\omega C = 1/2\pi fC$
We can use this like a resistance (because, really, it is a resistance) in an equation of the form V = IR to get the voltage across the capacitor:
$V = IX_c$
Note that V and I are generally the r.m.s. values of the voltage and current.

Inductance in an AC circuit



An inductor is simply a coil of wire (often wrapped around a piece of ferromagnet). If we now look at a circuit composed only of an inductor and an AC power source, we will again find that there is a 90° phase difference between the voltage and the current in the inductor. This time, however, the current lags the voltage by 90°, so it reaches its peak 1/4 cycle after the voltage peaks.
The reason for this has to do with the law of induction:
$e = -N\Delta\phi/\Delta t$ or $e = – L\Delta I/\Delta t$
Applying Kirchoff's loop rule to the circuit above gives:

$V – L\Delta I/\Delta t = 0$ so $V = L\Delta I/\Delta t$

As the voltage from the power source increases from zero, the voltage on the inductor matches it. With the capacitor, the voltage came from the charge stored on the capacitor plates (or, equivalently, from the electric field between the plates). With the inductor, the voltage comes from changing the flux through the coil, or, equivalently, changing the current through the coil, which changes the magnetic field in the coil.

To produce a large positive voltage, a large increase in current is required. When the voltage passes through zero, the current should stop changing just for an instant. When the voltage is large and negative, the current should be decreasing quickly. These conditions can all be satisfied by having the current vary like a negative cosine wave, when the voltage follows a sine wave.

How does the current through the inductor depend on the frequency and the inductance? If the frequency is raised, there is less time to change the voltage. If the time interval is reduced, the change in current is also reduced, so the current is lower. The current is also reduced if the inductance is increased.

As with the capacitor, this is usually put in terms of the effective resistance of the inductor. This effective resistance is known as the inductive reactance. This is given by $X_L = \omega L = 2\pi f L$, where L is the inductance of the coil (this depends on the geometry of the coil and whether it's got a ferromagnetic core). The unit of inductance is the henry.

As with capacitive reactance, the voltage across the inductor is given by: $V = IX_L$

Where does the energy go?

One of the main differences between resistors, capacitors, and inductors in AC circuits is in what happens with the electrical energy. With resistors, power is simply dissipated as heat. In a capacitor, no energy is lost because the capacitor alternately stores charge and then gives it back again. In this case, energy is stored in the electric field between the capacitor plates. The amount of energy stored in a capacitor is given by energy in a capacitor: Energy = $\frac{1}{2} CV^2$

In other words, there is energy associated with an electric field. In general, the energy density (energy per unit volume) in an electric field with no dielectric is: Energy density in an electric field = $\frac{1}{2} e_0 E^2$

With a dielectric, the energy density is multiplied by the dielectric constant.
There is also no energy lost in an inductor, because energy is alternately stored in the magnetic field and then given back to the circuit. The energy stored in an inductor is:
Energy in an inductor: Energy = ½ LI^2
Again, there is energy associated with the magnetic field. The energy density in a magnetic field is Energy density in a magnetic field = $B^2/(2\mu_0)$.

Power in an A.C. Circuit
The average power in an a.c. circuit is given by
$P = IV\cos\phi$
where I, V, are the effective (r.m.s.) values of the current and voltage respectively and ϕ is the angle of lag or lead between them. The quantity $\cos\phi$ is known as the power factor of the device. The power factor can have any value between zero and unity for ϕ varying from 90^o (or $\cos\phi = 0$) average power P, is zero. A power of zero means that the device is a pure reactance, inductance or capacitance. Thus no power is dissipated in an inductance or capacitance. However if I is the r.m.s. value of the current in a circuit containing a resistance, R, then the power absorbed in the resistance is given by
$P = I^2R$
For an a.c. circuit, the instantaneous power is given by P =IV (instantaneous values.)
$\text{Cos}\,\phi = R/Z$ = Resistance/Impedance

Resonance in RLC Series Circuit
The current, I, in an RLC series circuit is given by
$I = V/Z = V/\surd R^2 + (X_L - X_C)^2$
Where $X_L = 2\pi fL$ and $X_C = 1/2\pi fC$. The maximum current is obtained in the circuit when the impedance Z, in the above equation is minimum. This happens when
$X_L = X_C$ or $2\pi fL = 1/2\pi fC$

Resonance is said to occur in a.c. series circuit when the maximum current is obtained from such a circuit.

The frequency at which this resonance occurs is called the resonance frequency (f_0). This is the frequency at which $X_L = X_C$ or $2\pi f_0 L = 1/2\pi f_0 C$.
Hence solving the above equation we obtain that f_0 is given by
$f_0 = 1/2\pi\sqrt{LC}$
or since $\omega = 2\pi f$, we can write the condition of resonance as:
$\omega_0 = 1/\sqrt{LC}$
The variation of current I, and frequency f, in an RLC series circuit

Application of Resonance

The resonant circuit finds applications in electronics. It is used to tune radios and TVs. Its great advantage is that it responds strongly to one particular frequency. The other frequencies are very little effect.
Hence such a resonant circuit can select one signal of a definite frequency from a jumble of other signals available to it. That resonance frequency corresponds to that of a particular incoming radio signal. When this happens, maximum current is obtained and the distant radio station is loudly and clearly heard.

Example

An a.c. voltage of amplitude 2.0 volts is connected to an RLC series circuit. If the resistance in the circuit is 5 ohms, and the inductance and capacitance are 3 mH and 0.05µF respectively, calculate
i. the resonance frequency f_0
ii. the maximum a.c. current at resonance
Solution
$f_0 = 1/2\pi\sqrt{LC}$
$= 1/2\pi\sqrt{3 \times 10^{-3} \times 5 \times 10^{-8}} = 1/2\pi\sqrt{15 \times 10^{-11}}$
$= 1299.545$ Hz
$= 13$ KHz
At resonance $Z = R$ since $X_L = X_C$
$I_0 = V_0/R = 2/5 = 0.4$ Amps

CAPACITIVE REACTANCE

So far you have been dealing with the capacitor as a device which passes ac and in which the only opposition to the alternating current has been the normal circuit resistance present in any conductor. However, capacitors themselves offer a very real opposition to current flow. This opposition arises from the fact that, at a given voltage and frequency, the number of electrons which go back and forth from plate to plate is limited by the storage ability-that is, the capacitance-of the capacitor. As the capacitance is increased, a greater number of electrons change plates every cycle, and (since current is a measure of the number of electrons passing a given point in a given time) the current is increased.

Increasing the frequency will also decrease the opposition offered by a capacitor. This occurs because the number of electrons which the capacitor is capable of handling at a given voltage will change plates more often. As a result, more electrons will pass a given point in a given time (greater current flow). The opposition which a capacitor offers to ac is therefore inversely proportional to frequency and to capacitance. This opposition is called CAPACITIVE REACTANCE.

You may say that capacitive reactance decreases with increasing frequency or, for a given frequency, the capacitive reactance decreases with increasing capacitance. The symbol for capacitive reactance is X_C.

Now you can understand why it is said that the X_C varies inversely with the product of the frequency and capacitance. The formula is:

$X_c = 1/2\pi fc$

Where: X_C is capacitive reactance in ohms f is frequency in Hertz C is capacitance in farads 2π is 6.28 (2 X 3.1416).

The following example problem illustrates the computation of X_C.

Given : f = 100 Hz, C = 50µF

Solution: $X_c = 1/2\pi fc$

$= 1/6.28 \times 100\text{Hz} \times 50\mu\text{F}$

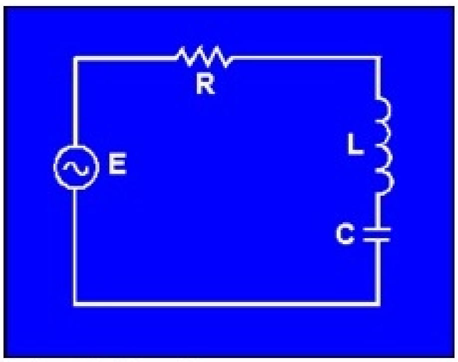
$= 1/0.0314\Omega$

$= 31.8\Omega$ or 32Ω

REACTANCE, IMPEDANCE, AND POWER RELATIONSHIPS IN AC CIRCUITS

Up to this point inductance and capacitance have been explained individually in ac circuits. The rest of this part will concern the combination of inductance, capacitance, and resistance in ac circuits.

To explain the various properties that exist within ac circuits, the series RLC circuit will be used. Figure (4) is the schematic diagram of the series RLC circuit. The symbol shown in figure 4-4 that is marked E is the general symbol used to indicate an ac voltage source.



Series RLC circuit.

REACTANCE

The effect of inductive reactance is to cause the current to lag the voltage, while that of capacitive reactance is to cause the current to lead the voltage. Therefore, since inductive reactance and capacitive reactance are exactly opposite in their effects, what will be the result when the two are combined? It is not hard to see that the net effect is a tendency to cancel each other, with the combined effect then equal to the difference between their values. This resultant is called REACTANCE; it is represented by the symbol X; and expressed by the equation $X = X_L - X_C$ or $X = X_C - X_L$. Thus, if a circuit contains 50 ohms of inductive reactance and 25 ohms of capacitive reactance in series, the net reactance, or X, is 50 ohms – 25 ohms, or 25 ohms of inductive reactance.

For a practical example, suppose you have a circuit containing an inductor of 100 µH in series with a capacitor of .001 µF, and operating at a frequency of 4 MHz What is the value of net reactance, or X?

Given f = 4MHz, L = 100µH, C =0.001µF

Solution: $X_L = 2\pi fL$

$= 6.28 \times 4MHz \times 100\mu H = 2512\ \Omega$

$X_c = 1/2\pi fC$

$= 1/6.28 \times 4MHz \times 0.001\mu F$

$= 1/0.2515\Omega$

$= 39.8\Omega$

$X = X_L - X_C$

$X = 24722\Omega$ (inductive)

IMPEDANCE

From your study of inductance and capacitance you know how inductive reactance and capacitive reactance act to oppose the flow of current in an ac circuit. However, there is another factor, the resistance, which also opposes the flow of the current. Since in practice ac circuits containing reactance also contain resistance, the two combine to oppose the flow of current. This combined opposition by the resistance and the reactance is called the IMPEDANCE, and is represented by the symbol Z.

Since the values of resistance and reactance are both given in ohms, it might at first seem possible to determine the value of the impedance by simply adding them together. It cannot be done so easily, however. You know that in an ac circuit which contains only resistance, the current and the voltage will be in step (that is, in phase), and will reach their maximum values at the same instant. You also know that in an ac circuit containing only reactance the current will either lead or lag the voltage by one-quarter of a cycle or 90 degrees. Therefore, the voltage in a purely reactive circuit will differ in phase by 90 degrees from that in a purely resistive circuit and for this reason reactance and resistance are combined by simply adding them.

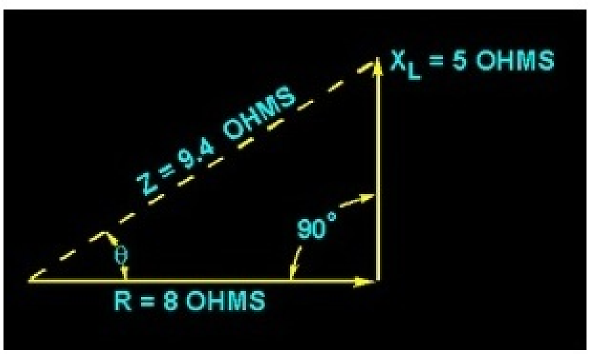
When reactance and resistance are combined, the value of the impedance will be greater than either. It is also true that the current will not be in step with the voltage nor will it differ in phase by exactly 90 degrees from the voltage, but it will be somewhere between the in-step and the 90-degree out-of-step conditions. The larger the reactance compared with the resistance, the more nearly the phase difference will approach 90°. The larger the resistance compared to the reactance, the more nearly the phase difference will approach zero degrees.

If the value of resistance and reactance cannot simply be added together to find the impedance, or Z, how is it determined? Because the current through a resistor is in step with the voltage across it and the current in a reactance differs by 90 degrees from the voltage across it, the two are at right angles to each other. They can therefore be combined by means of the same method used in the construction of a right-angle triangle.

Assume you want to find the impedance of a series combination of 8 ohms resistance and 5 ohms inductive reactance. Start by drawing a horizontal line, R, representing 8 ohms resistance, as the base of the triangle.

Then, since the effect of the reactance is always at right angles, or 90 degrees, to that of the resistance, draw the line X_L, representing 5 ohms inductive reactance, as the altitude of the triangle. This is shown in figure (5). Now, complete the hypotenuse (longest side) of the triangle. Then, the hypotenuse represents the impedance of the circuit.

Vector diagram showing relationship of resistance, inductive reactance, and impedance in a series circuit.



Vector diagram showing relationship of resistance, inductive reactance, and impedance in a series circuit.

One of the properties of a right triangle is:

$(\text{hypotenuse})^2 = (\text{base})^2 + (\text{altitude})^2$

Hypotenuse = $\sqrt{(\text{base})^2 (\text{altitude})^2}$

Applied to impedance, this becomes,

$(\text{impedance})^2 = (\text{resistence})^2 + (\text{reactance})^2$

Impedance = $\sqrt{(\text{resistance})^2 + (\text{reactance})^2}$

Impedance = $\sqrt{(\text{resistance})^{2+} (\text{reactrance})^2}$

$Z = \sqrt{R^2 + X^2}$

Now suppose you apply this equation to check your results in the example given above.

Given : $R = 8\Omega$, $X_L = 5\Omega$

Solution : $Z = \sqrt{R^2 + X_L{}^2}$

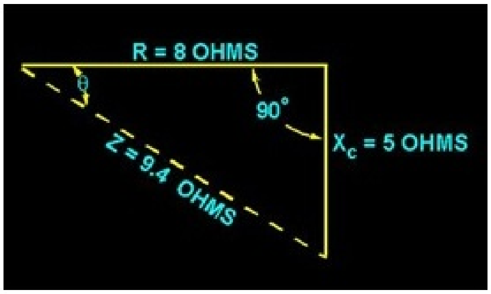
$= \sqrt{(8\Omega)^2 + (5\Omega)^2}$

$= \sqrt{64 + 25}$

$= \sqrt{89}\Omega = 9.4\Omega$

When you have a capacitive reactance to deal with instead of inductive reactance as in the previous example, it is customary to draw the line representing the capacitive reactance in a downward direction. This is shown in figure (6). The line is drawn downward for capacitive reactance to indicate that it acts in a direction opposite to inductive reactance which is drawn upward. In a series circuit containing capacitive reactance the equation for finding the impedance becomes:

$Z = \sqrt{R^2 + X_C{}^2}$



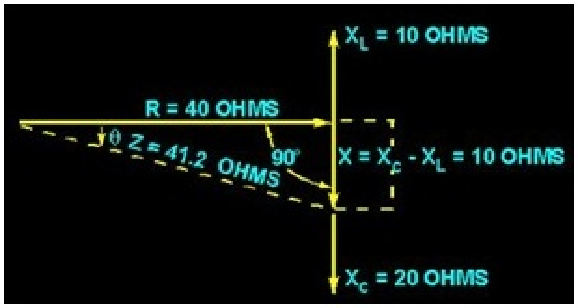
Vector diagram showing relationship of resistance, capacitive reactance, and impedance in a series circuit

In many series circuits you will find resistance combined with both inductive reactance and capacitive reactance. Since you know that the value of the reactance, X, is equal to the difference between the values of the inductive reactance, X_L, and the capacitive reactance, X_C, the equation for the impedance in a series circuit containing R,X_L, and $X_C$then becomes:

$Z = \sqrt{R^2 + (X_L - X_C)^2}$ or $Z = \sqrt{R^2 + X^2}$

Note: The formulas $Z = \sqrt{R^2 + X_C{}^2}$, $Z = \sqrt{R^2 + X^2}$ and $Z = \sqrt{R^2 + X_L{}^2}$ can be used to calculate Z only if the reisitance and reactance are connected in series.

In the above expressions, you will see the method which may be used to determine the impedance in a series circuit consisting of resistance, inductance, and capacitance.



Vector diagram showing relationship of resistance, reactance (capacitive and inductive), and impedance in a series circuit

Assume that 10 ohms inductive reactance and 20 ohms capacitive reactance are connected in series with 40 ohms resistance. Let the horizontal line represent the resistance R. The line drawn upward from the end of R, represents the inductive reactance, X_L. Represent the capacitive reactance by a line drawn downward at right angles from the same end of R. The resultant of X_L and X_C is found by subtracting $X_L$from X_C. This resultant represents the value of X.

Thus: $X = X_C – X_L = 20$ ohms – 10 ohms

= 10 ohms

Note: If inductive reactance is smaller than the capacitive reactance and is therefore subtracted from the capacitive reactance.

These two examples serve to illustrate an important point: when capacitive and inductive reactance are combined in series, the smaller is always subtracted from the larger and the resultant reactance always takes the characteristics of the larger.

Question

1. Given $X_L = 10\ \Omega$, $X_C = 20\ \Omega$, $R = 40\ \Omega$, calculate the impedance.

A. 40Ω B. 41.2Ω C. 56Ω D. 45.3Ω

Given f = 1MHz, L = 100µH, C = 0.0002µF

2. Calculate for inductive reactance

A. 352Ω B. 702Ω C. 628Ω D. 245Ω

3. Calculate for capacitive reactance

A. 801Ω B. 235Ω C. 769Ω D. 745Ω

4. Now assume you have a circuit containing a 100 – µH inductor in series with a .0002-µF capacitor, and operating at a frequency of 1 MHz. What is the value of the resultant reactance in this case?

A. 168Ω B. 200Ω C. 265Ω D. 158Ω

5. As with capacitive reactance, the voltage across the inductor is given by

A. IX_L B. IX_C C. IX_R D. IX_Z

Answer

1. B 2. C 3. C 4. A 5. A

WEEK 3

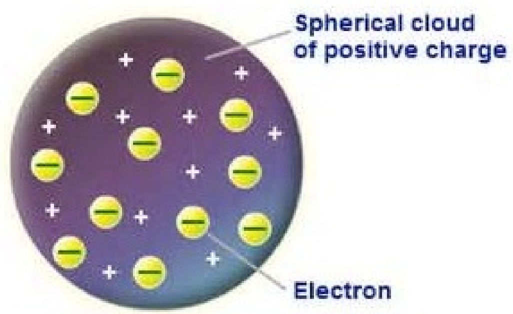
Physics SS3
Topic: MODEL OF ATOMS

Thompson Rutherford Model

Modern atomic theory provides a reasonably satisfactory explanation of the properties of matter, the mechanisms of chemical change and the interaction of matter and energy. Such a theory emerged from the synthesis of the work of several scientists, only a few of which will be discussed here.

John Dalton (1766-1844) is generally credited as the father of the atomic theory but the early Greek philosophers were the originators of the concept of atoms. They taught that matter is composed of atom and is therefore finitely divisible, John Dalton considered the atom as constituting the simplest component of matter. He viewed the atom as 'indestructible', tiny hard spheres. The discovery of radioactivity by Henry Becquerel (1891 showed that atoms are complex rather than 'indivisible' or 'indestructible' but can disintegrate forming atom of different elements. The discovery of cathode rays in electric discharge tubes (1895) by William Crookes revealed that negatively charged electrons were components of the atom.

By 1900, it was already established that matter consists of atom but nothing was known about the structure of the atom. It was only known that the atoms contain electrons but that on the whole the atom was electrically neutral. This neutrality means that there must exist within the atom enough positively charged components to balance the negatively charged electrons. This led Sir J. J. Thomson the English Physicist, to propose an atomic model which visualized the atom as a homogeneous sphere of positive charge inside of which are embedded negatively charged electrons as shown below



Thomson's Plum pudding model

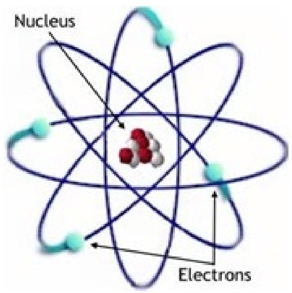
J. J. Thomson also determined the ratio of the charge to mass, e/m, of electrons, and found e/m to be identical for all cathode-ray particles, irrespective of the kind of gas the in the tube or the metal the electrodes are made of.

Rutherford Model – the Nuclear Idea and the Planetary Model

Ernest Rutherford (1911) and his co-workers performed experiments a beam of positively charged alpha (α) particles was directed at a pin sheet of a metal foil. It was found that most of the alpha-particles passed through the foil without deflection as if the foil were mostly empty space. Only a few of them were diverted from their paths. Some of these few actually rebounded backwards.

The scattering of alpha-particle by the metal foil was explained as a repulsion from a heavy positively charged nucleus present at the centre of the atom of the metal foil. This follows because an abrupt change in path (as noted for a few α-particles) of a relatively heavy and positively charged α-particle can result only from its hitting or from its close approach to another particle (the nucleus) with a highly concentrated, positive charge. All these contradicted Thomson's atomic model which supposed that the distribution of the charges was diffuse. Hence Rutherford proposed his model of the atom.

Rutherford proposed a planetary model of the atom which suggested that the atom consists of a positively charged heavy core called the nucleus which most of the mass of the atom was concentrated. Around this nucleus negatively charged electron circle in orbits much as planets move around the sun. Each nucleus must be surrounded by a number of electrons necessary to produce an electrically neutral atom.



This model was a major step toward how we view the atom today. It however had two problems. According to Newtonian Physics, such an atom as Rutherford's would collapse by spiraling into the nucleus, since there is an attractive force between the oppositely charged nucleus and electrons. Further experiments indicated that charged moving in a field of opposite charge lose energy by emitting radiation. But atoms in their normal state neither collapsed nor emitted radiation. Thus the two main difficulties of the Rutherford model are these:

1. It predicts that light of a continuous range of frequencies will be emitted, whereas experiment shows line spectra instead of continuous spectra.
2. It predicts that atoms are unstable – electrons quickly spiral into the nucleus – but we know that atoms in general are stable, since the matter around is stable.

Clearly Rutherford's model was not sufficient to explain experimental observations. Some sort of modification was needed and this was provided by Neils Bohr.

Assumption of Bohr's theory

In 1913 Niels Henrik Bohr published his new theory of the atoms constitution. Just like Rutherford he assumed that electrons rotate around the nucleus. But had the three completely new ideas:

1. There are some orbits called by him the stationery ones, where the moving electrons don't emit energy.
2. Each emission or absorption of radiation energy represents the electron transition from one stationery orbit to another. The radiation emitted during such transition is homogeneous and its frequency is given by the formula:

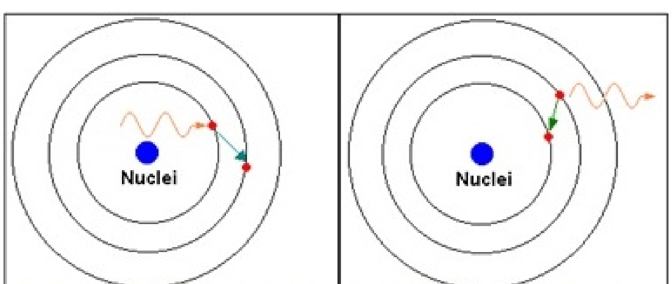
where h is the Planck constant, E_n and E_l are the energies in the two stationary states.

3. The laws of mechanics describe the dynamic equilibrium of electrons in stationery states but do not describe the situation of the electron transition from one stationery orbit to another.

Let's now think what each postulate means:

The first one says that electrons can't move on unlimited orbits around the nucleus. Only some orbits are permissible. Electrons moving on them don't lose energy for radiation. The postulate was in complete disagreement with other theories, and especially with the Maxwell theory of electromagnetism. Bohr formulated the postulate ad hoc. He didn't know what it might come from. But he was of the opinion that to properly understand the nature of the atom one has to accept his idea. The second postulate says that in an atom an electron can change orbits. On each orbit the electron has some defined energy. The energy of the electron is different on different orbits. The bigger the orbit is, the bigger the energy is. If the electron change a higher orbit into a lower one then it emits a quant of energy that is the same as the difference of energy of the higher and lower orbit. To change a lower orbit into a higher one the electron has to absorb an adequate quant of energy. The quant of energy is proportional to the frequency of the emitted radiation. The second postulate explains why the atom emit radiation of strictly defined wavelengths.

The third postulate is in complete disagreement with the classic theory. According to that postulate the laws of mechanics can only describe electrons moving on stationary orbits and not while changing their orbits.



An electron (the red one) absorbs a photon (the orange one) and jumps to a higher orbit (the green arrow).

An electron (the red one) emits a photon (the orange one) and jumps to a lower orbit (the green arrow).

The Electron-Cloud Model

The model visualizes the atom as consisting of a tiny nucleus of radius of the order of 10^{-15}m. The electron is visualized as being in rapid motion within a relatively large region around the nucleus, but spending most of its time in certain high-probability regions. Thus the electron is not considered as a ball revolving around nucleus but as a particle or wave with a specified energy having only a certain probability of being in are given region in the space outside the nucleus. The electron is visualized as spread out around the nucleus in a sort of electron-cloud. Some chemists prefer to consider the electron in terms of a cloud of negative charges (electron cloud), with a cloud being dense in regions of high electron probability and more diffuse in regions of low probability.

This model is therefore known as the electron-cloud model. Most of the time, but not always, the electron will be located inside spherical outline. In other words, the probability of finding the electron inside the spherical bonding is high. The probability then decreases rapidly as the distance of the thin shell from the nucleus increases.

Protons, Neutrons and Isotope

In considering the atom as made up of a tiny but massive nucleus at the centre and outside the nucleus is a cloud of electrons which move in wave-like orbits or shells around the massive nucleus . The nucleus consists of protons which carry positive charges and neutrons which carry negative charges. The proton and the neutron together constitute the nucleon. Practically all the mass of an atom is concentrated in the central nucleus. The protons, neutrons and electrons are the fundamental subatomic particles of the atom.

The electron is the lightest particle of an atom, with a mass (m_e) of 9.1 x 10^{-31} kg and an electronic charge e^- = 1.6 x 10^{-19} C.

The proton has a mass of 1.67 x 10^{-27} kg which is over 1836 times heavier than the mass of an electron. It carries a positive charge, e^+ = 1.6 x 10^{-19} C (i.e. e^+ = e^- = 1.6 x 10^{-19} C). There are the same number of protons in the atoms of a given element but different number of protons in the atoms of different elements. For example, there is only one proton in a hydrogen atom, but eleven proton in a sodium atom. In a neutral atom, the number of proton equals the number of electrons. The neutron has the same mass as the protons but it carries no charge.

We denote the atom of an element X by ${}^{A}_{Z}X$ where A is the mass number and Z is its atomic number.

The atomic number or proton number (Z) is the number of proton in the nucleus of an element. The mass number of Nucleon number (A) is the total number of protons and neutrons in an atom of an element. Thus the number of neutrons in the atom of an element. Thus the number of neutrons in the atom of an element equals the difference between the mass number and the atomic number (i.e. A-Z). The carbon atom denoted by ${}^{12}_{6}C$ has 6 protons and 6 neutrons and a mass number of 12; the nitrogen atom denoted by ${}^{14}_{7}N$ has 7 protons, and 7 neutrons and a mass number of 14. Since a neutral atom has the same number of electrons as the number of protons, ${}^{12}_{6}C$ has 6 protons and 6 electrons.

An important question to consider is if all atoms are composed of these same components, why do different atoms have different chemical properties? The answer lies in the number and arrangement of the electrons. The electrons comprise most of the atomic value and thus are the parts that 'intermingle' when atoms combine to form molecules. Therefore the number of electrons possessed by a given atom greatly affects its ability to interact with other atoms.

The properties of the elements are periodic functions of their atomic number. When arranged in order of increasing atomic numbers, the elements with similar chemical properties recur at definite intervals, i.e. periodically. In regard to the electronic structure of electronic configuration (i.e. the arrangement of electrons in shells around the nucleus), this suggests a periodicity in the number of electrons in the outer shells of the atoms of the elements. The electrons in the outermost shell or shells of an atom are called valence electrons. These valence electrons are largely responsible for the chemical behavior of the atom of an element. If elements having the same number of valence electrons are grouped together, the elements falling within each group have similar chemical properties. Thus we note that atomic mass, atomic number, valence and periodicity, all indicate particularity of matter.

Isotopes are atoms of the same elements which have the same atomic number (Z) but different mass number. Isotopes are thus atoms with the same number of protons but different number of neutrons. Isotopes have similar chemical properties because they have the same number of electrons round the nucleus. Chemical combination is due to an exchange of outer or valence electrons between elements.

In nature, elements are usually found as a mixture of isotopes. For example, two naturally occurring isotopes of chlorine are

i. $^{35}_{17}Cl$ (17 protons, 17 electrons, 18 neutrons)

ii. $^{37}_{17}Cl$ (17 protons, 17 electrons, 20 neutrons)

For Carbon we have:

i. $^{12}_{6}C$ (6 protons, 6 electrons, 6 neutrons)

ii. $^{13}_{6}C$ (17 protons, 6 electrons, 7 neutrons)

For Oxygen we have:

i. $^{16}_{8}O$ (8 protons, 8 electrons, 8 neutrons)

ii. $^{17}_{8}O$ (8 protons, 8 electrons, 9 neutrons)

iii. $^{18}_{8}O$ (8 protons, 8 electrons, 10 neutrons)

For Uranium we have:

i. $^{238}_{92}U$ (92 protons, 92 electrons, 146 neutrons)

ii. $^{235}_{92}U$ (92 protons, 92 electrons, 143 neutrons)

iii. $^{234}_{92}U$ (92 protons, 92 electrons, 142 neutrons)

Questions

1. Which of the following statements is not correct? Isotopes of an element have

A. the same number of electric charges on the nucleus B. the same chemical properties C. different nucleon numbers D. different atomic mass

2. Which of the following representations is correct for an atom X with 28 electrons and 30 neutrons?

A. $^{30}_{28}X$ B. $^{28}_{30}X$ C. $^{58}_{30}X$D. $^{58}_{28}X$

3. Which of the following statements are correct?

I The neutron has no charge II The electron is lighter than the proton III The proton is positively charged IV The algebraic sum of the charges on the protons and the charges on the neutrons in a neutral atom is zero

A. I, II, III, IV B. I, II, III only C. I, III and IV only D. II and IV only

4. Which of the following particles determine the mass of an atom?

A. Protons and neutrons B. Neutrons only C. Protons and electrons D. Neutrons and electrons

5. Which of the following names is not associated with models of the atom

A. Neils Bohr B. Ernest Rutherford C. J. J. Thomson D. Isaac Newton

Answers

1. C 2. D 3. A 4. A 5. D

WEEK 4

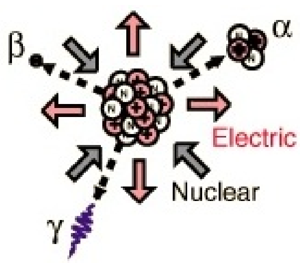
Physics SS 3

Topic: RADIOACTIVITY

RADIOACTIVITY

Radioactivity refers to the particles which are emitted from nuclei as a result of nuclear instability. Because the nucleus experiences the intense conflict between the two stronger forces in nature, it should not be surprising that there are many nuclear isotopes which are unstable and emit some kind of radiation. The most common types of radiation are called alpha, beta and gamma radiation, but there are several other varieties of radioactive decay.

Alpha Radioactivity

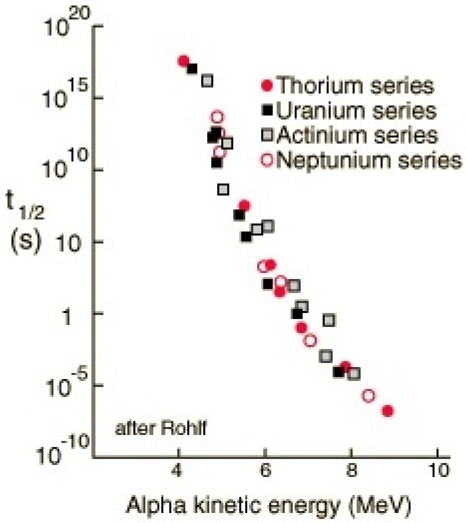


Composed of two protons and two neutrons, the alpha particle is a nucleus of the element helium. Because of its very large mass (more than 7000 times the mass of the beta particle) and its charge, it has a very short range. It is not sustainable for radiation therapy since its range is less than a tenth of a millimeter inside the body. Its main radiation hazard comes when it is ingested into the body; it has great destructive power within its short range. In contact with fast-growing membranes and living cells, it is positioned for maximum damage.

Alpha particle emission is modeled as a barrier penetration process. The alpha particle is the nucleus of the helium atom and is the nucleus of highest stability.

Alpha Barrier Penetration

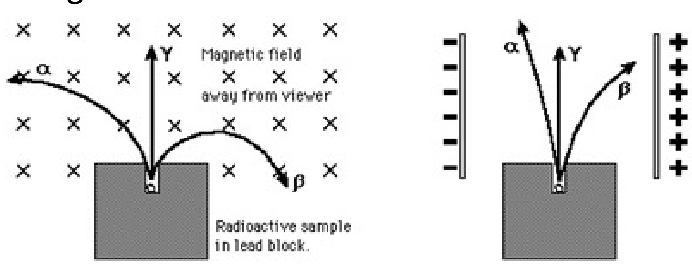
The energy of emitted alpha particles was a mystery to early investigators because it was evident that they did not have enough energy, according to classical physics, to escape the nucleus. Once an approximate size of the nucleus was obtained by Rutherford scattering, one could calculate the height of the Coulomb barrier at the radius of the nucleus. It was evident that this energy was several times higher than the observed alpha particle energies. There was also an incredible range of half lives for the alpha particle which could not be explained by anything in classical physics.



The resolution of this dilemma came with the realization that there was a finite probability that the alpha particle could penetrate the wall by quantum mechanical tunneling. Using tunneling, Gamow was able to calculate a dependence for the half-life as a function of alpha particle energy which was in agreement with experimental observations.

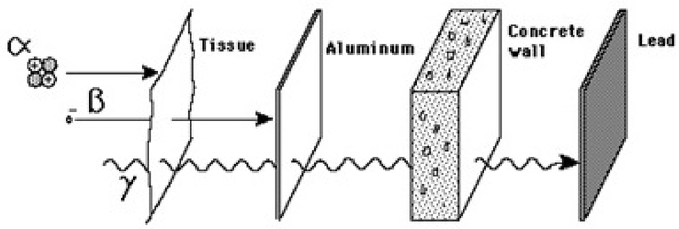
Alpha, Beta, and Gamma

Historically, the products of radioactivity were called alpha, beta, and gamma when it was found that they could be analyzed into three distinct species by either a magnetic field or an electric field.



Penetration of Matter

Though the most massive and most energetic of radioactive emissions, the alpha particle is the shortest in range because of its strong interaction with matter. The electromagnetic gamma ray is extremely penetrating, even penetrating considerable thicknesses of concrete. The electron of beta radioactivity strongly interacts with matter and has a short range.



Different radiations have different properties, as summarized below:

Properties of Radiation

Radiation	Alpha (α) –particles	Beta particles	Gamma (λ) rays
Nature	Helium nuclei $^{4}_{2}H$	High energy electrons	Electromagnetic waves of very short wavelength
Velocity			

Charge
5-7% speed of light
+2e(+3.2 x 10^{-19} C)Travel at approx speed of light –e(-1.6 x 10^{-19} C)Travel at speed of light
Electrically neutral Mass Relatively massive Relatively light Negligible Effect of magnetic field Slightly deflected in a magnetic field in a direction expected for a +ve charge Strongly deflected in a magnetic field in a direction expected for a -ve charge Small or no effect Ionizing power Large, cause heavy ionization Medium. About 0.1% of that of α-particles Small Penetrating power Little penetrating power e.g. by thin sheets of paper Good penetrating power in air, , several mm of aluminium High penetrating power in air and in solid e.g. many cm of lead Fluorescent Cause fluorescence in ZnS No fluorescence in ZnS

Peaceful Uses of Radiation

Although scientists have only known about radiation since the 1890s, they have developed a wide variety of uses for this natural force. Today, to benefit humankind, radiation is used in medicine, academics, and industry, as well as for generating electricity. In addition, radiation has useful applications in such areas as agriculture, archaeology (carbon dating), space exploration, law enforcement, geology (including mining), and many others.

1. Medical Uses
2. Academic and Scientific Applications
3. Industrial Uses
4. Nuclear Power Plants

Medical Uses

Hospitals, doctors, and dentists use a variety of nuclear materials and procedures to diagnose, monitor, and treat a wide assortment of metabolic processes and medical conditions in humans. In fact, diagnostic x-rays or radiation therapy have been administered to about 7 out of every 10 Americans. As a result, medical procedures using radiation have saved thousands of lives through the detection and treatment of conditions ranging from hyperthyroidism to bone cancer.

The most common of these medical procedures involve the use of x-rays — a type of radiation that can pass through our skin. When x-rayed, our bones and other structures cast shadows because they are denser than our skin, and those shadows can be detected on photographic film. The effect is similar to placing a pencil behind a piece of paper and holding the pencil and paper in front of a light. The shadow of the pencil is revealed because most light has enough energy to pass through the paper, but the denser pencil stops all the light. The difference is that x-rays are invisible, so we need photographic film to “see” them for us. This allows doctors and dentists to spot broken bones and dental problems.

X-rays and other forms of radiation also have a variety of therapeutic uses. When used in this way, they are most often intended to kill cancerous tissue, reduce the size of a tumor, or reduce pain. For example, radioactive iodine (specifically iodine-131) is frequently used to treat thyroid cancer, a disease that strikes about 11,000 Americans every year.

X-ray machines have also been connected to computers in machines called computerized axial tomography (CAT) or computed tomography (CT) scanners. These instruments provide doctors with color images that show the shapes and

details of internal organs. This helps physicians locate and identify tumors, size anomalies, or other physiological or functional organ problems.
In addition, hospitals and radiology centers perform approximately 10 million nuclear medicine procedures in the United States each year. In such procedures, doctors administer slightly radioactive substances to patients, which are attracted to certain internal organs such as the pancreas, kidney, thyroid, liver, or brain, to diagnose clinical conditions.

Academic and Scientific Applications
Universities, colleges, high schools, and other academic and scientific institutions use nuclear materials in course work, laboratory demonstrations, experimental research, and a variety of health physics applications. For example, just as doctors can label substances inside people's bodies, scientists can label substances that pass through plants, animals, or our world. This allows researchers to study such things as the paths that different types of air and water pollution take through the environment. Similarly, radiation has helped us learn more about the types of soil that different plants need to grow, the sizes of newly discovered oil fields, and the tracks of ocean currents. In addition, researchers use low-energy radioactive sources in gas chromatography to identify the components of petroleum products, smog and cigarette smoke, and even complex proteins and enzymes used in medical research.
Archaeologists also use radioactive substances to determine the ages of fossils and other objects through a process called carbon dating. For example, in the upper levels of our atmosphere, cosmic rays strike nitrogen atoms and form a naturally radioactive isotope called carbon-14. Carbon is found in all living things, and a small percentage of this is carbon-14. When a plant or animal dies, it no longer takes in new carbon and the carbon-14 that it accumulated throughout its life begins the process of radioactive decay. As a result, after a few years, an old object has a lower percent of radioactivity than a newer object. By measuring this difference, archaeologists are able to determine the object's approximate age.

Industrial Uses

We could talk all day about the many and varied uses of radiation in industry and not complete the list, but a few examples illustrate the point. In irradiation, for instance, foods, medical equipment, and other substances are exposed to certain types of radiation (such as x-rays) to kill germs without harming the substance that is being disinfected — and without making it radioactive. When treated in this manner, foods take much longer to spoil, and medical equipment (such as bandages, hypodermic syringes, and surgical instruments) are sterilized without being exposed to toxic chemicals or extreme heat. As a result, where we now use chlorine — a chemical that is toxic and difficult-to-handle — we may someday use radiation to disinfect our drinking water and kill the germs in our sewage. In fact, ultraviolet light (a form of radiation) is already used to disinfect drinking water in some homes.

Similarly, radiation is used to help remove toxic pollutants, such as exhaust gases from coal-fired power stations and industry. For example, electron beam radiation can remove dangerous sulphur dioxides and nitrogen oxides from our environment. Closer to home, many of the fabrics used to make our clothing have been irradiated (treated with radiation) before being exposed to a soil-releasing or wrinkle-resistant chemical. This treatment makes the chemicals bind to the fabric, to keep our clothing fresh and wrinkle-free all day, yet our clothing does not become radioactive. Similarly, nonstick cookware is treated with gamma rays to keep food from sticking to the metal surface.

The agricultural industry makes use of radiation to improve food production and packaging. Plant seeds, for example, have been exposed to radiation to bring about new and better types of plants. Besides making plants stronger, radiation can be used to control insect populations, thereby decreasing the use of dangerous pesticides. Radioactive material is also used in gauges that measure the thickness of eggshells to screen out thin, breakable eggs before they are packaged in egg cartons. In addition, many of our foods are packaged in polyethylene shrinkwrap that has been irradiated so that it can be heated above its usual melting point and wrapped around the foods to provide an airtight protective covering.

All around us, we see reflective signs that have been treated with radioactive tritium and phosphorescent paint. Ionizing smoke detectors, using a tiny bit of americium-241, keep watch while we sleep. Gauges containing radioisotopes measure the amount of air whipped into our ice cream, while others prevent spillover as our soda bottles are carefully filled at the factory.

Engineers also use gauges containing radioactive substances to measure the thickness of paper products, fluid levels in oil and chemical tanks, and the moisture and density of soils and material at construction sites. They also use an x-ray process, called radiography, to find otherwise imperceptible defects in metallic castings and welds. Radiography is also used to check the flow of oil in sealed engines and the rate and way that various materials wear out. Well-logging devices use a radioactive source and detection equipment to identify and record formations deep within a bore hole (or well) for oil, gas, mineral, groundwater, or geological exploration. Radioactive materials also power our dreams of outer space, as they fuel our spacecraft and supply electricity to satellites that are sent on missions to the outermost regions of our solar system.

Nuclear Power Plants

Electricity produced by nuclear fission — splitting the atom — is one of the greatest uses of radiation. As our country becomes a nation of electricity users, we need a reliable, abundant, clean, and affordable source of electricity. We depend on it to give us light, to help us groom and feed ourselves, to keep our homes and businesses running, and to power the many machines we use. As a result, we use about one-third of our energy resources to produce electricity.

Electricity can be produced in many ways — using generators powered by the sun, wind, water, coal, oil, gas, or nuclear fission. In America, nuclear power plants are the second largest source of electricity (after coal-fired plants) — producing approximately 21 percent of our Nation’s electricity.

The purpose of a nuclear power plant is to boil water to produce steam to power a generator to produce electricity. While nuclear power plants have many similarities to other types of plants that generate electricity, there are some significant differences. With the exception of solar, wind, and hydroelectric plants, power plants (including those that use nuclear fission) boil water to produce steam that spins the propeller-like blades of a turbine that turns the shaft of a generator. Inside the generator, coils of wire and magnetic fields interact to create electricity. In these plants, the energy needed to boil water into steam is produced either by burning coal, oil, or gas (fossil fuels) in a furnace, or by splitting atoms of uranium in a nuclear power plant. Nothing is burned or exploded in a nuclear power plant. Rather, the uranium fuel generates heat through a process called fission.

Nuclear power plants are fueled by uranium, which emits radioactive substances. Most of these substances are trapped in uranium fuel pellets or in sealed metal fuel rods . However, small amounts of these radioactive substances (mostly gases) become mixed with the water that is used to cool the reactor. Other impurities in the water are also made radioactive as they pass through the reactor. The water that passes through a reactor is processed and filtered to remove these radioactive impurities before being returned to the environment. Nonetheless, minute quantities of radioactive gases and liquids are ultimately released to the environment under controlled and monitored conditions.

Radioactive Decay; Half-Life; Decay Constant

Radioactivity is a spontaneous process. It goes on independent of external control. it is not affected by temperature or pressure or by chemical treatment . It is also a random process as no one can predict which atom will disintegrate at a given time.

Experiments have shown that each radioactive element has a definite rate of decay which can be characterized by its Half-life.

Half-Life of a radioactive element is the time taken for half of the atoms initially present in the element to decay.

If the half-life of an element is T years, it means that after T years, 1 gm of the element will have a mass of 1/2gm, after 2T years, the mass of the element will be 1/4gm (or ½ of ½ gm) and so on.

Thus if we have 1000 atoms of a radioactive element initially, whose half-life is 10 years, then after 10 years, 500 atoms will remain ; after 20 years, 250 atoms will be left and after 30 years, 125 atoms will be left undecayed and so on.

Decay Constant, λ

The rate of decay of radioactive elements is found to be proportional to the number of atoms of the material present. Suppose there are N atoms of a radioactive element present at a time, t, then the probable number of disintegrate per unit time or activity can be expressed by $-dN/dt$ (The minus sign arises from the fact that N is decreasing with time). Since the rate of disintegration is proportional to the number of atoms present at a given time, we have

$-dN/dt \propto N$ or $dN/dt = -\lambda N$

where λ is a constant of proportionality called the Decay Constant of the element. From the above equation we have

$\lambda = -1/N(dN/dt)$

Hence Decay Constant is defined as the instantaneous rate of decay per unit atom of a substance OR

no. of atoms disintegrating per second/no. of atoms in the source at the time = λ

Also by interpreting the first equation, we have that

$N = N_0e^{-\lambda t}$

where N_0 is the number of atom present at time t = 0 (i.e. at the time when observations of decay were begun) and N is the number of atoms present at time t.

We obtain the time required for half of the atoms to disintegrate (half-life) by substituting $N = 1/2N_0$ into this equation $N = N_0e^{-\lambda t}$ and eliminating N_0 we have

$N_0/2 = N_0e^{-\lambda t}$

$\frac{1}{2} = e^{-\lambda t}$

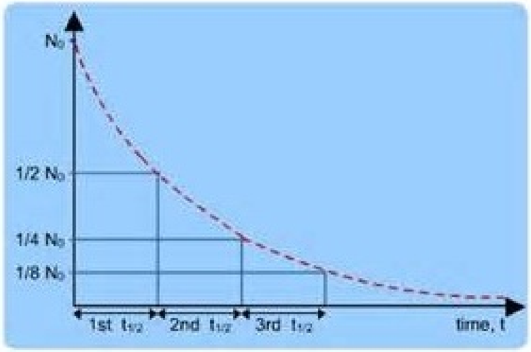
Taking the natural or Naperian logarithm of both

$\log_e \frac{1}{2} = -\lambda t$

But $\log_e \frac{1}{2} = \log_e 1 - \log_e 2 = 0 - \log_e 2 = -0.693$

Hence, $-0.693 = -\lambda t$

$t = 0.693/\lambda$



Questions

1. Radioactivity refers to the particles which are emitted from nuclei as a result of
A. nuclear stability B. nuclear instability C. Uranium decrement D. Proton Increment
2. Which of these is not part of radioactive emission?
A. Alpha   B. Gamma   C. Electron D. Beta
3. Complete this statement, The rate of decay of radioactive elements is found to be ………………………….. to the number of atoms of the material present.
A. Proportional B. Inverse C. indirect D. direct
4. Half-Life of a radioactive element is the time taken for half of the atoms initially present in the element to ………………………
A. form B. compose C. decay D. stimulate
5. A certain radioactive element has a half-life of 10 years. How long will it take to loose 7/8 of its atoms originally present.
A. 20 B. 30 C. 25 D. 35

Answer

1. B 2. C 3. A 4. C 5. B

WEEK 5

Physics SS 2

Topic: Transformation of Elements

Transformation of Elements

There are two types of radioactivity – natural and artificial radioactivity.

The phenomenon of radioactivity was first discovered by Henri Becquerel

Natural radioactivity is the spontaneous disintegration of the nucleus of an atom during which α-particle or β-particle, or gamma rays or a combination of any or all the three and heat (or energy) are released.

When a radioactivity element undergoes radioactive decay, it may emit either α-particle, β-particle or γ rays. This changes the atomic number of the element, hence a new element is formed. For example, Radium-226 decays by emitting an α-particles to turn into a new element Radon. Radium-226 has a mass number 226 and an atomic number 88 and hence it is denoted by ${}^{226}_{88}Ra$. The α-particles it emits is a Helium nucleus denoted by ${}^{4}_{2}He$. So when Radium 226 emits an α-particle. We can write a nuclear equation:

$${}^{226}_{88}Ra \rightarrow {}^{4}_{2}He + \underbrace{{}^{222}_{86}Rn}_{\text{(Radon – 222)}} + \text{energy}$$

Radon – 222 decays to Radium – 222 by emitting 2 β-particles. When the nucleus of an atom emits a β-particle (i.e. an electron), the atomic number of the atom increases by one unit, but its mass number remains unaltered. Hence since two β-particles are emitted from Radon 222 we can write the equation

$${}^{222}_{86}Rn \rightarrow 2\ {}^{0}\text{-}1^{e} + {}^{222}_{88}\ Ra + \text{energy}$$

Uranium-238 decays by emitting two α-particles and two β-particles to thorium -230. Hence we can write the nuclear equation thus:

$${}^{238}_{92}U \rightarrow \underbrace{2\ {}^{4}_{2}H}_{\text{2α-particles}} + \underbrace{2\ {}^{0}\text{-}1^{e}}_{\text{2β-particles}} + {}^{230}_{90}Th + \text{energy}$$

Thorium-234 decays by emitting a β-particle to the element Protactinium-234 thus:

$${}^{234}_{90}\ Th \rightarrow {}^{0}\text{-}1^{e} + {}^{234}_{91}Pa + \text{energy}$$

Generally we represent alpha (α) decay by

$^{A}_{B}X \rightarrow {}^{4}_{2}He + {}^{(A-4)}_{(z-2)}Y$

$\uparrow$

α-particle

and Beta (β)-decay by $^{A}_{Z}X \rightarrow {}^{0}-1^{e} + {}^{A}_{(Z+1)}Y$

$\uparrow$

B-particle

Gamma radiation (γ) is a form of light, emitted as photons of energy *hf*, and has zero mass number and zero charge (A = 0, Z = 0).

To balance a nuclear equation we ensure that the sun of the atomic numbers, Z (subscripts) must be the same on the two sides of the equation. Also the sum of the mass numbers A (superscripts) must be the same on the two sides of the equation.

Artificial Radioactivity

If the radioactivity is induced in an element by irradiation with, for example, neutrons, the process is known as artificial radioactivity. By irradiation we mean exposure to radiation either by accident or by intent.

Artificial disintegration was first achieved by Rutherford when he disrupted nitrogen nucleus with energetic α-particles, to produce first of all an isotope of fluorine. The fluorine nucleus being unstable disintegrates immediately into an oxygen nucleus and a proton thus:

$^{4}_{2}He + {}^{14}_{7}N \rightarrow {}^{18}_{9}F \rightarrow {}^{17}_{8}O + {}^{1}_{1}H$ + Energy

In artificial radioactivity an ordinary material, not normally radioactive is made radioactive by bombarding it with radioactive particles

$^{4}_{2}He + {}^{27}_{13}Al$

$\rightarrow {}^{30}_{15}P + {}^{1}_{0}n \rightarrow {}^{30}_{14}Si + {}^{0}_{1}e$ + Energy

Thus phosphorus nuclei which are not stable but radioactive can be produced by bombarding non-radioactive aluminium with α-particles. The radioactive phosphorus nuclei then disintegrates spontaneously into stable Silicon atoms.

The neutron, the proton and the α-particles have been found very effective bombarding projectiles for disintegrating the nuclei of elements. Examples of such bombardments are:

$^{1}_{0}n + {}^{6}_{3}Li \rightarrow {}^{3}_{1}H + {}^{4}_{2}He$ + Energy

$^{1}_{0}n + ^{24}_{12}Mg \rightarrow ^{24}_{11}Na + ^{1}_{1}P$ + Energy

$^{4}_{2}He + ^{9}_{4}Be \rightarrow ^{12}_{6}C + ^{1}_{0}n$ + Energy

$^{1}_{0}n + ^{59}_{27}Co \rightarrow ^{60}_{27}Co$ + Energy

Isotopes can be made artificially by bombarding neutrons or protons or deuterons at elements; e.g.

$^{34}_{10}S + ^{1}_{0}n \rightarrow ^{35}_{10}S$ + Energy

$^{79}_{35}Br + ^{1}_{0}n \rightarrow ^{80}_{35}Br$ + Energy

Such artificially produced isotopes are unstable and decay with the emission of α-particles, β-particles and γ-rays. They are therefore called radioisotopes.

Radioisotopes or radioactivity isotopes are isotopes that are made artificially by bombarding neutrons or protons or deuterons at elements.

Nuclear Energy

The protons and neutrons (nucleons) in the nucleus of each atom are held together by very powerful nuclear forces. An enormous amount of energy is therefore required to tear the nucleon apart. This energy is over 10^6 times more than that required to remove the electrons from an atom.

Nuclear Fusions

Fusion is a nuclear process in which two or more light nuclei combine or fuse to form a heavier nucleus with the release of a large amount of energy.

A typical example of nuclear fusion is the fusion of two hydrogen nuclei to form Helium, He.

$$\underset{\text{Deuterium}}{\underset{\uparrow}{^{2}_{1}H}} + \underset{\text{tritium}}{\underset{\uparrow}{^{3}_{1}H}} \rightarrow \underset{\text{Helium}}{\underset{\uparrow}{^{4}_{2}He}} + \underset{\text{neutron}}{\underset{\uparrow}{^{1}_{0}n}} + \text{Energy}$$

In the above fusion process the two isotopes of hydrogen, deuterium and tritium, combine to form the heavier nucleus of Helium. To bring the two light nuclei together in a fusion process, very high temperature of the order of 10^6 – 10^8 degrees Celcius are required to over-come the coulomb repulsive forces between the two nuclei. This poses severe technological problem due to the fact that materials to withstand this high temperature are difficult to come by. It is for this reason that fusion reactions have not been harnessed in nuclear power stations

on Earth to produce power, even though they theoretically offer the prospect of enormous quantities of cheap power.
The sun and stars produce an enormous output of energy through nuclear fusion. This is because of the presence of hydrogen isotope in the sun and the conditions of extremely high temperature and pressure are found in the interior of the sun and stars.

Advantages of Fusion over Fission

1. Fusion is more easily achieved with lightest elements such as hydrogen; nuclear repulsion is easily overcome as nuclei approach each other.
2. The draw materials required for fusion are more cheaply and readily available. For example, hydrogen can be obtained by the electrolysis of sea water which is cheap and in plentiful supply.
3. Fusion process produces less dangerous (i.e. non radioactive) by-products.
4. There is no upper limit to the mass of hydrogen that can be exploded in a nuclear fusion process, so very large energies can be obtained. Nuclear fusion has been used in the hydrogen bomb.

Nuclear Fission

As was first shown in 1934 by Enrico Fermi, the heavy nucleus of Uranium-235 can be split into two other elements, Krypton and Barium, by bombarding it with a slow neutron

${}^{1}_{0}n + {}^{235}_{92}U \rightarrow {}^{141}_{56}Ba + {}^{92}_{36}Kr + 3{}^{1}_{0}n$ + Energy

It was found that the total mass of the component products is less than the mass of the original Uranium. The difference in mass (mass defect) is a measure of the nuclear energy released. According to Albert Einstein

$E = \Delta mc^2$

where E is the energy released, Δm is the difference in mass and c is the velocity of light. The amount of energy released when 1 g of Uranium-235 undergoes fission is about 7.4×10^{10} J, a prodigious amount of energy. This energy released in the form of heat.
Nuclear Fission is the splitting up of the nucleus of a heavy element into two approximate equal parts with the release of a huge amount of energy and neutrons.

Fission can occur with most of the very massive nuclei (e.g. Plutonium and Uranium) and has been produced by slow neutron, high-energy alpha particles, protons, X-rays and gamma-rays.
In the bombardment of Uranium-235 by slow neutron, several neutrons are produced as by-products. These neutrons may cause the splitting of other Uranium nuclei which, in turn yield more neutrons which may further split other Uranium nuclei and so on. Thus a chain reaction is set in motion. A chain reaction is a multiplying and self-maintaining reaction. When the size of Uranium exceeds a certain critical mass, there is a rapid production of neutrons accompanied by a release of a tremendous amount of energy in a nuclear explosion. This is the principle of the atomic and nuclear fission bombs.
Fission is also the process used in the present day nuclear power stations.

Applications of Radioactivity

Radioactive substances find very much application in
(i) Agricultural and scientific research
(ii) Medical field and
(iii) Industrial field

Agricultural and scientific research

Radioactive elements are used in agriculture as radioactive tracers and to induce mutations in plants and animals to obtain new and improved varieties. Biologists used them as tracers or markers to help trace the paths of metabolic processes in plants and animals. Geologists and archaeologists use the measurement of half-life to estimate the age or rocks, and carbon-14 (radiocarbon) to date recent organic remains. This is called radioactive dating.

Medical Field

Gamma-rays from radioactive substances are used to treat cancer patients, to sterilize surgical equipment, foods etc.

In Industry

Radioactive elements are used in industry (a) to study the defects in metals and welded joints, and to check metal fatigue. (b) As radioisotopes, they are used to trace underground pipe leakages.

Radioactive Hazard and Safety Precautions

Radioactive substances emit continuously powerful radiations such as β particles, ϒ particles and α particles ray. These are high energy radiation and hence, harm to living tissues. Energy absorbed by the passage of radiation through human body gives rise to the structural change called radiation damage. This damage may lead to death of a person. M. curie and E. Fermi lost their lives because of the damage caused by these radiations. When radiations enter a living system, the cell and tissues gets damaged due to the interaction with radiations. The harmful effects on an organism caused by these radiations is called radiation hazard. The damaged cell and tissue hamper normal functioning of the living system living ultimately to the death of the organism.

Following types of the damages can be caused by the radiation hazards:

1. The exposure to radiation induces deleterious (harmful often in a subtle or an unexpected way) genetic effects. When the radiation passes through genetic cells, there occur mutations of the chromosome of the cellular nuclei. The mutations are transmitted from one generation to the next and so on. The genetic effects are irreversible.
2. The strong α-ray exposure can cause lung cancer.
3. The exposure to fast and slow neutron can cause blindness.
4. The exposure to neutrons, protons, and α particle can cause damage to red blood cells.
5. The strong exposures to protons and neutrons can cause serious damage to reproductive organs.

Following are the safety precautions for radiation hazards.

1. The radio isotopes should be transferred in thick wall lead containers and kept it rooms with thick walls of lead.
2. The radio isotopes are handled with the help of remote control device.
3. The workers are asked to wear lead aprons.

4. The radioactive contamination of the working area is avoided at all casts.
5. Nuclear explosions should be carried out far away from the public area.

Binding Energy

Nucleons are protons and neutrons in the nucleus of an atom. Binding energy is the energy required to take all the nucleons apart so that they are totally separated. When separated, it follows that the total mass is often less than the mass of the nucleus. Binding energy can also be referred t as the difference in mass.

Binding Energy = mass difference of nucleus and nucleons.

Questions

1. Which of these is not relevant in the applications of radioactivity?

A. Agricultural and scientific research B. Medical field C. Industrial field D. Carpentry field

2. What are Beta particles?

A. Protons B. neutrons C. electrons D. Helium nuclei

3. The phenomenon of radioactivity was first discovered by

A. Marie Curie B. Henri Becquerel C. Sir J. J. Thomson D. Enrico Fermi

4. Alpha particles are

A. not charged B. Highly penetrating C. electromagnetic radiation D. Helium nuclei

5. Gamma radiation (γ) is a form of light, emitted as photons of energy hf, and has

A. zero mass number and zero charge B. zero mass only C. zero charge only D. none is correct.

Answers

1. D 2. C 3. B. 4. D. 5.A

<u>WEEK 6</u>

Physics SS 3

Topic: **QUANTUM**

INTRODUCTION

Some of the important postulates of the Bohr's model of the atom were that the electrons moved around the nucleus in specified orbits. In these orbits they can move without radiating energy.

They can acquire or lose energy only in discrete units called quanta. Thus Bohr suggests that electrons in the atom exist in discrete (or quantized) energy states.

In 1902, Max Planck was able to show that experimental observations in the radiation emitted by substances could be explained on the basis that the energy from such bodies emitted in separate or discrete packets of energy known as energy quanta of value hf, where f is the frequency of radiation and h is a constant known as Planck's Constant. Thus the energy E of the quantum of radiation or photon is given by $E = hf$.

This is known as Planck's theory of radiation. The term quantum means amount fixed amount, or discrete or separate amount as distinguished from a continuous quantity.

Planck's quantum hypothesis thus suggests (and this is accepted today) that the energy of radiation can be $E = hf$, or $2hf$ or $3hf$, etc. but there cannot be vibrations or radiations whose energy lies between these values. That is energy radiated is not a continuous quantity but is rather quantized – i.e., it exists only in discrete amounts. This is the meaning of the concept of energy quantization.

The electrons in an atom can only have certain, specific amounts of energy. This is an unusual concept—if an electron is whirling around an atom with a given amount of energy.

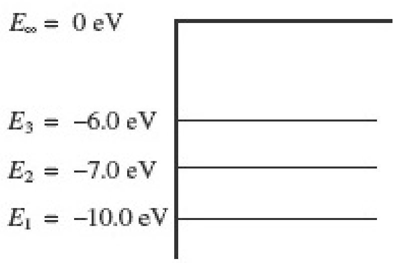


Figure 25.1 Energy levels in a hypothetical atom.

Energy Levels in an Atom

In Figure 25.1, we've drawn a few of the energy levels of a hypothetical atom. Let's start by looking at E1. This is the lowest energy level, and it is called the ground state energy of the atom. When an electron is sitting as close to the nucleus as possible, and when it's completely unexcited, it will have the energy of E1, -10 eV.

To increase the energy of the electron—to move the electron into a higher energy level—energy must somehow be transferred to the electron. This energy transfer is done by a photon. To jump up in energy, an electron absorbs a photon, and to drop down in energy, an electron emits a photon.

The energy diagram in Figure 25.1 tells us that to get to E2 from the ground state, an electron must absorb a photon carrying 3 eV—and only 3 eV!—of energy. If a photon with an energy of 2.9 eV comes along and knocks into the electron, nothing will happen. Similarly, if a photon with an energy of 3.1 eV comes along, nothing will happen. It's all or nothing; either the electron gets just the right amount of energy to go from one energy level to the next, or it doesn't.

How does the electron get from the ground state to E3? There are two ways. The electron could first absorb a photon with an energy of 3 eV, taking it from E1 to E2, and then it could absorb another photon with an energy of 1 eV, taking it from E2 to E3. Or, the electron could start in E1 and simply absorb a photon with an energy of 4 eV.

We've been talking about photons having certain energies, but we haven't yet told you how to figure out the energy of a photon. Here's the formula:

E = hf

This formula tells us that the energy of a photon is equal to Planck's constant, h, which is 6.63 × 10-34 J·s (this value is given to you on the constants sheet), multiplied by the frequency of the photon. You should remember from Chapter 23 that the frequency of a wave is related to the wavelength by the formula

$v = \lambda f$

For light, the velocity is c, or 3 . 108 m/s, so we can instead write

$c = \lambda f$

This means that we can rewrite the equation for the energy of a photon to read

$E = hc/\lambda$

These formulas tell us that a photon with a high frequency, and therefore with a small wavelength, is higher in energy than a photon with a low frequency and long wavelength. So gamma rays, for example, are a lot higher energy than radio waves because gamma rays have a higher frequency.

This is a simple plug-and-chug problem. Figure 25.1 gives us the energy levels, and we just need to use our formula to find the wavelength of the absorbed photon. But what value do we use for the energy? -10 eV? -7 eV? Wrong… the value we use is the energy of the jump, not the energy of one of the states. To get from E1 to E2, an electron must absorb 3 eV of energy, so that's the value we use in the formula:

At this point, the problem is simply a matter of converting units and using your calculator. However, solving it is a lot easier if you know that the quantity hc equals 1240 eV nm. This value is on the constants sheet, and it is very important to know! by knowing this value, we can solve for the photon's wavelength very quickly.

In other words, for an electron in Figure 25.1 to go from the ground state to E2, it must absorb light with a wavelength of exactly 410 nm, which happens to be a lovely shade of violet.

If you look back at Figure 25.1, you might wonder what's going on in the gap between E3 and E∞. In fact, this gap is likely filled with lots more energy levels—E4, E5, E6… However, as you go up in energy, the energy levels get squeezed closer and closer together (notice, for example, how the energy gap between E1 and E2 is greater than the gap between E2 and E3). However, we didn't draw all these energy levels, because our diagram would have become too crowded.

The presence of all these other energy levels, though, raises an interesting question. Clearly, an electron can keep moving from one energy level to the next if it absorbs the appropriate photons, getting closer and closer to E∞. But can it ever get beyond E_∞? That is, if an electron in the ground state were to absorb a photon with energy greater than 10 eV, what would happen?

The answer is that the electron would be ejected from the atom. It takes exactly 10 eV for our electron to escape the electric pull of the atom's nucleus—this minimum amount of energy needed for an electron to escape an atom is called the ionization energy[2]—so if the electron absorbed a photon with an energy of, say, 11 eV, then it would use 10 of those eV to escape the atom, and it would have 1 eV of leftover energy. That leftover energy takes the form of kinetic energy.

As we said above, it takes 10 eV for our electron to escape the atom, which leaves 1 eV to be converted to kinetic energy. The formula for kinetic energy requires that we use values in standard units, so we need to convert the energy to joules.

$1\text{eV}(1.6 \times 10^{-19}\text{J/eV}) = \frac{1}{2} mv^2$

If we plug in the mass of an electron for m, we find that v = 5.9 × 105 m/s. Is that a reasonable answer? It's certainly quite fast, but electrons generally travel very quickly. And it's several orders of magnitude slower than the speed of light, which is the fastest anything can travel. So our answer seems to make sense.

The observation that electrons, when given enough energy, can be ejected from atoms was the basis of one of the most important discoveries of twentieth century physics: the **photoelectric effect**.

Photoelectric Effect: Energy in the form of light can cause an atom to eject one of its electrons, but only if the frequency of the light is above a certain value.

This discovery was surprising, because physicists in the early twentieth century thought that the brightness of light, and not its frequency, was related to the light's energy. But no matter how bright the light was that they used in experiments, they couldn't make atoms eject their electrons unless the light was above a certain frequency. This frequency became known, for obvious reasons, as the cutoff frequency. The cutoff frequency is different for very type of atom.

A metal surface has a work has a work function of 10eV. What is the cutoff frequency for this metal?

Remembering that hc equals 1240 eV·nm, we can easily find the wavelength of a photon with energy of 10 eV:

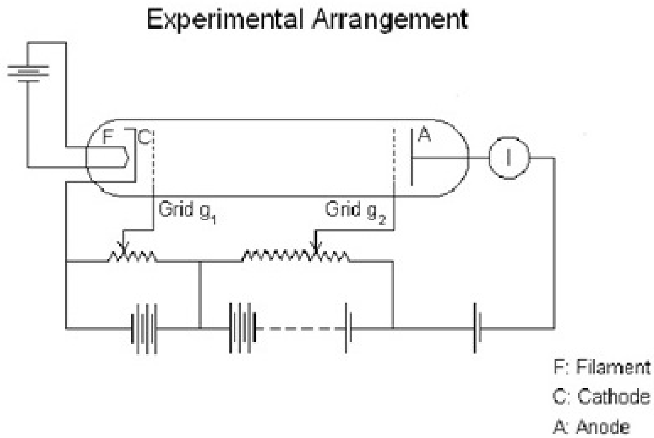
$\lambda = 1240\text{eV.nm}/10\text{eV}$

= 124nm

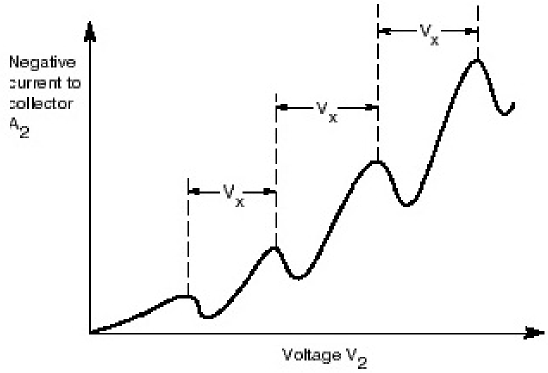
Using the equation c = λf, we find that f = 2.42 × 1015 Hz. So any photon with a frequency equal to or greater than 2.42 × 1015 Hz carries enough energy to eject a photon from the metal surface.

Franck-Hertz Experiment

The discreteness of atomic energy levels was first shown directly in the Franck-Hertz experiment. This experiment is one of the classic demonstrations of the quantization of atomic energy levels and was first performed by J. Frank and G. Hertz in 1914. Their goal was to verify the quantum theory assumptions about the existence of discrete energy levels in atoms and that quantized amounts of energy are transferred in emission and absorption. By accelerating a beam of electrons through a mercury vapor, they found that when the kinetic energy of the electrons reached about 4.9 eV, the vapor emitted a spectrum line at 254 nm. This experiment led to the detailed investigation of the atomic structure of many elements. When an electron encounters an atom and bounces off without losing any of its energy then such an event is called an “elastic scattering”. The electron will be elastically scattered unless it has sufficient energy to cause a change in the internal energy of the atom. Since atomic energy levels are quantized, this means that electrons flowing through a gas of atoms with less energy than the first excited state of the atoms will not lose any energy as they travel. On the other hand, an electron with enough energy to cause a transition to an excited state of the atom may induce such a transition with subsequent loss of kinetic energy. Such an event is called an “inelastic scattering” of the electron by the atom. In the Franck Hertz experiment electrons are emitted from a hot cathode into a tube filled with mercury vapor. The experimental arrangement is shown in figure below.



The driving potential Vg_1 reduces the space charge and causes a large current to flow in the tube. The electrons are then accelerated through a positive potential by the accelerating potential Vg_2. After being accelerated, the electrons are slowed by a potential drop in the opposite direction by the braking potential. Then the electrons are collected at a far end of the tube and the current is measured. In a vacuum tube that contains no gas the current would rise steadily as the accelerating voltage Vg_2, is increased. The presence of the gas changes this behavior because of collisions of the electrons with the gas atoms. At first the current does rise with the potential, but when the electrons get enough energy they inelastically collide with the gas atoms and excite higher energy levels in the gas. After these collisions the electrons will have lower energy and due to the opposing potential, they will not make it to the end of the tube. This will cause the current to decrease to a minimum. After this minimum, as the potential increases the current will again increase until the electrons get enough energy to excite the gas twice. This process continues with the electrons repeatedly exciting the gas atoms. The potential difference between the minima (or maxima) is equivalent to the energy of the excited level. The graph plotted between the accelerating potential and anode current is shown below.



At each of the critical potentials (V_1, V_2, or V_3 etc.) the electrons have just sufficient kinetic energy to raise the internal energy of the sodium atom by collision.

Spectra

Embedded in an atom which consists of a nucleus that is positively charged and which revolves round the orbit is an electron which is negatively charged.

When the atom or metal is heated and it absorbs energy to a particular intensity, the electrons starts to collide with one another and with the electron shell until the electrons are able to liberate themselves from the orbit. They create an atom with high energy level which teams up with the low energy level. When this occurs, radiation is emitted. Some of the particles and energies in the radiation are equal to the difference in energy of the atom between initial and final stage.

$E = E_0 - E_1$

The study of spectra is referred to as spectroscopy.

Types of Spectra

(i) Emission spectra:

These are spectra observed due to light emitted when the temperature of an atom, gas or metal is raised, e.g. incandescent solid or vapour , a spark discharge, arc discharge, the discharge of electricity through a gas or vapour contained in a discharged tube, etc.

(ii) Line Spectra

Line spectrum is obtained from atoms in gases such as hydrogen or neon at low pressure in a discharge tube. This spectrum mainly occurs in sodium, mercury vapours and hydrogen gas. When a light is incidented on a vapour-like mercury or hydrogen, a particular temperature is reached when or vapour starts to emit light of different wavelengths which depends on the nature of gas or metal. The emission can be viewed with the aid of a spectrometer in a slit. Differences are observed to be discontinuous but not overlapping, with equal distance between them, i.e. they are regularly spaced but do not form a series.

Line spectra occur when there is transmission of electron from one region to another usually from higher to lower energy level or vice-versa. Line spectra observed can be represented by:

$\Delta E = hf = E_n - E_o$

Where hf is the change of energy level (eV)

E_n = higher energy level or excitation energy level (eV)

E_o = lower energy level or ground state (eV)

Where h Planck's constant with value at 6.6×10^{-34}J

(iii) Continuous Spectrum

The spectrum formed from white light contains all colors, or frequencies, and is known as a continuous spectrum. Continuous spectra are produced by all incandescent solids and liquids and by gases under high pressure. A gas under low pressure does not produce a continuous spectrum but instead produces a line spectrum, i.e., one composed of individual lines at specific frequencies characteristic of the gas, rather than a continuous band of all frequencies. If the gas is made incandescent by heat or an electric discharge, the resulting spectrum is a bright-line, or emission, spectrum, consisting of a series of bright lines against a dark background. A dark-line, or absorption, spectrum is the reverse of a bright-line spectrum; it is produced when white light containing all frequencies passes through a gas not hot enough to be incandescent. It consists of a series of dark lines superimposed on a continuous spectrum, each line corresponding to a frequency where a bright line would appear if the gas were incandescent. The Fraunhofer lines appearing in the spectrum of the sun are an example of a dark-line spectrum; they are caused by the absorption of certain frequencies of light by the cooler, outer layers of the solar atmosphere. Line spectra of either type are useful in chemical analysis, since they reveal the presence of particular elements.

(iv) Band Spectrum

It consists of a number of bright bands with a sharp edge at one end but fading out at the other end. Band spectra are obtained from molecules. It is the characteristic of the molecule. Calcium or barium salts in a bunsen flame and gases like carbon-di-oxide, ammonia, and nitrogen in molecular state in the discharge tube give band spectra. When the bands are examined with high resolving power spectrometer, each band is found to be made of a large number of fine lines, very close to each other at the sharp edge but shaped out at the other end. Using band spectra the molecular structure of the substance can be studied.

(v) Absorption Spectrum

When a cool gas is placed in the path of a continuous spectrum of light, dark absorption lines will appear in the resulting spectrum. Conversely, if we observe the gas in an oblique angle, we can see emission lines produced by the gas too. The most interesting thing about the spectrum of an element is that, the wavelengths of absorption lines and emission lines produced by an element match exactly each other. This is the one of the evidences for electrons in an atom are situated in some kind of resonant columns.

When a photon falls on an electron and the resonant frequency of the shell (electron shell or transitory shell) in which the electron exists matches with the frequency of the photon, the electron will oscillate with the frequency of the photon and the photon will be efficiently absorbed by the atom. In the resultant spectrum, the absorbed photon will not be present. This is the way an atom produces its absorption lines.

When an electron oscillates, it will emit a radiation with the frequency of its oscillation. Therefore an atom produces emission lines with exactly matching wavelengths of the absorption lines that produced by the atom.

Questions

1. Calculate the energy in joules of ultraviolet light of wavelength 3×10^{7} m. Take the velocity of light as 3×10^{8}m/s and Planck's constant as 6.6×10^{-34} Js

A. 6.6×10^{-19} B. 9.8×10^{-2} C. 3×10^{-8} D. 4.5×10^{8}.

2. An electron jumps from one energy level to another in an atom radiating 4.5 x 10^{-19} Joules. If Planck's constant is 6.6 x 10^{34} Js, what is the wavelength of the radiation? Take velocity of light = (3 x 10^{8} m/s).

A. 3.4 x 10^{-7}m B. 4.4 x 10^{-7} C. 5.9 x 10^{-8} D. 2.5 x 10^{-6}

3. If an electron in the ground state were to absorb a photon with energy greater than 10 eV, what would happen?

A. the electron will stick to the atom B. the electron would be ejected from the atom C. the electron wil be neutralised D. there will be increase in the electron in the atom.

4. Which of the spectra Line spectra occur when there is transmission of electron from one region to another usually from higher to lower energy level or vice-versa?

A. Line spectra B. Continuous spectra C. Absorption spectra D. Emission spectra

5. Which of these is not correct?

A. Thus Bohr suggests that electrons in the atom exist in discrete (or quantized) energy states.

B. The discreteness of atomic energy levels was first shown directly in the Franck-Hertz experiment.

C. Franck-Hertz experiment is one of the classic demonstrations of the quantization of atomic energy levels

D. To increase the energy of the electron, to move the electron into a higher energy level—energy of the electron must be stable.

Answer

1. A 2. B 3. B 4. A 5. D

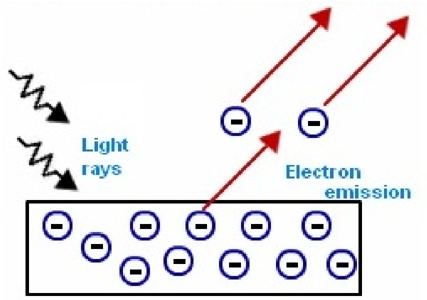
WEEK 7

Physics SS 3

Topic: PHOTOELECTRIC EFFECT

INTRODUCTION

In **1887**, photoelectric effect was invented by the scientist **H. Hertz**. When we passing a light into a material, the material should emit an electrons. This effect is called as Photoelectric effect. Some of the rays produced in **Photoelectric Effect** are x-rays, ultraviolet rays and *γ*-rays.



Einstein's says that every photon or a quantum of light has an energy value h*v*. Electrons are located inside the surface of the metal. There may be some amount of work W should be required for an electron to come out of the metal. If the transferred energy is high comparing the internal attractions then the electrons gets liberated. In this the one photon can release a one electron. This is nothing but the **Photoelectric Effect**. Let us study more about the Photoelectric effect in this section.

Photoelectric Effect Definition

It is the phenomenon of emission of electrons from the surface of metals when the radiations of suitable frequency and suitable wavelength fall on the surface of the metal.

The following parameters are related to Photoelectric effect:

Photoelectric Current

Stopping Potential

Threshold Energy

Work Function

The emitted electrons are called photo electrons and the current so produced is called as the **Photoelectric Current**. The intensities of photo electrons vary with material.

The **Stopping Potential** is the potential difference applied to stop the electrons from being ejected from the surface when the light falls on it.

The **Threshold Frequency** is defined as the minimum frequency of incident light required for the photo electric emission. It is denoted by v.

Work Function is the minimum amount of energy necessary for the photo electric emission to start. It is denoted by ϕ_0.

Photoelectric Effect Equation

The Einstein equation describes the phenomena of photoelectric effect but the validity of this equation is only up to two regions, one is visible light and second is for ultraviolet light.

The energy of photon is equal to energy required to remove electron + kinetic energy of the emitted electron.

E = W + K.E.

where E = Energy of photon

W = Work function

K.E = Kinetic Energy.

Energy of photon, E = h v …………………….. (a)

Where h = Planck's constant

v = frequency of photon incident light or

So **h v = W + K.E …………………………. (b)**

The work function **W** is defined as the minimum energy needed to remove an electron from the surface of a given metal. It is equal to h v0.

K.E= This is the kinetic energy (maximum) of emitted electrons.

K.E = 1/2 mv^2.

Here v_0 is the threshold frequency.

m = rest mass of the emitted electron

v = The speed or velocity of the emitted electron.

Work function or threshold energy (φ):

The minimum energy of incident photon below which no ejection of photon electron from a metal surface will take place is known as work function of threshold energy for that metal.

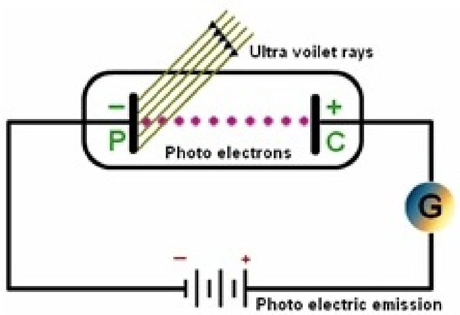
ϕ = h $v0$ = $hC/\lambda 0$

Work function is the characteristic of a given metal. If E is the Energy of incident photon then

1. If E < ϕ, no photo electric effect will take place.
2. If E = ϕ, photoelectric effect will take place but kinetic effect of ejected photo electron is zero.
3. If E > ϕ, photo electric effect will take place along with possession of kinetic energy by ejected electron.

Photoelectric Effect Experiment

The Phenomenon was discovered by *Hertz* and was experimentally proved by Thomsonand Millikan. Einstein added his photoelectric equation to support the phenomenon. The phenomenon is basically about light energy (photo) converting into electrical energy.



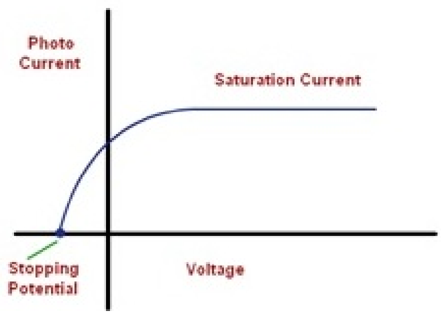
The Apparatus consists of an evacuated quartz tube having photosensitive plate called emitter **A** and collector **B**. Plate **A** is connected to the negative terminal and plate **B** is connected to the positive terminal of a battery via rheostat **R**. The potential difference between the plates **AB** can be adjusted by changing the value of the rheostat. When light of suitable wavelength is incident on the plate **A** then electrons are emitted and reaches plate **B**, thus measurable current flows through the circuit.

Factors on which Photoelectric effect depends :

The amount of current flow (number of electrons) and the kinetic energy of the emitted electrons depend upon :

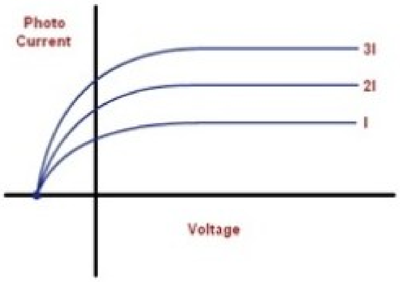
1. Applied Potential difference between the plates:

For a given photo metal if the frequency and intensity of incident light kept constant and if the potential difference between the plates is increased then photo current also increases until it reaches a maximum (saturated) value. If the terminals are reversed and we increase the potential difference gradually then photo current decreases and at one point it becomes zero. This point where the current is zero is termed as **Stopping Potential**.

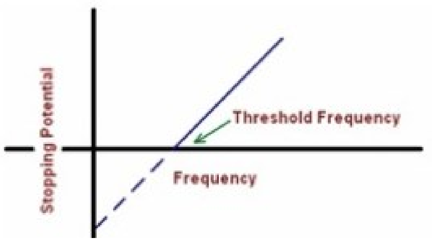


2. Intensity of incident radiation:

For a given photo metal if the frequency and applied voltage is kept constant with increasing intensity then we observe that photo current increases with increases in intensity with out change in the stopping potential.



3. Frequency of incident radiation:
The intensity and voltage across the plates is kept constant for a photo metal with change in frequency of the incident light. The stopping potential is measured for different frequencies. The graph is drawn between frequencies versus stopping potential. The frequency at which the photo current begins is called threshold frequency. This is the minimum frequency for photo electric effect to occur.



4. Material of the Photo metal:
The experiment is repeated for different photo metal and their threshold frequencies are noted. The variation of frequency versus stopping potential is graphed and noticed that threshold frequency varies from metal to metal.

Photoelectric Effect Applications
Below are the applications of Photoelectric effect:
1. The photoelectric effect is used in the photoelectric cell which converts a light energy into electrical energy.
2. In cinematography photoelectric effect has a application of reproducing the sound.
3. Photoelectric effect also has an application in street lights for automatic switch on and off.
4. Traffic signals are using this effect for automatic controls and for count the machines.
5. Working of burglar alarm uses this photoelectric effect.
6. Television transmission is one of the applications of this photoelectric effect.

Threshold Frequency

Threshold frequency is defined as the minimum frequency of incident light which can cause photo electric emission i.e. this frequency is just able to eject electrons without giving them additional energy.

It is denoted by V_0

Work Function

Minimum amount of energy which is necessary to start photo electric emission is called Work Function. If the amount of energy of incident radiation is less than the work function of metal, no photo electrons are emitted.

It is denoted by ϕ_0. Work function of a material is given by $\phi_o = hv_0$.

It is a property of material. Different materials have different values of work function. Generally, elements with low I.P values have low work function such as Li, Na, K, Rb, and Cs.

Einstein's Equation

The mass-energy relationship announced by Albert Einstein in 1905 in the form $E = \mathbf{mc}^2$, where E is a quantity of energy, m its mass, and c is the speed of light. It presents the concept that energy possesses mass. **2.** The relationship $E_{max} = \mathbf{hf} - W$, where E_{max} is the maximum kinetic energy of electrons emitted in the photoemissive effect, h is the Planck constant, f the frequency of the incident radiation, and W the work function of the emitter. This is also written $E_{max} = \mathbf{hf} - \phi e$, where e is the electronic charge and ϕ a potential difference, also called the work function. (Sometimes W and ϕ are distinguished as **work function energy** and **work function potential**.) The equation can also be applied to photoemission from gases, when it has the form: $E = \mathbf{hf} - I$, where I is the ionization potential of the gas.

X-ray Production

When fast-moving electrons slam into a metal object, x-rays are produced. The kinetic energy of the electron is transformed into electromagnetic energy. The function of the x-ray machine is to provide a sufficient intensity of electron flow from the cathode to anode in a controlled manner. The three principal segments of an x-ray machine – a control panel, a high-voltage power supply, and the x-ray tube are all designed to provide a large number of electrons focused to a small spot in such a manner that when the electrons arrive at the target, they have acquired high kinetic energy.

Kinetic energy is the energy of motion. Stationary objects have no kinetic energy; objects in motion have kinetic energy proportional to their mass and the square of their velocity.

The equation used to calculate kinetic energy is:

$$\mathbf{KE = \tfrac{1}{2} mv^2}$$

where m is the mass in kilograms, v is the velocity in meters per second, and KE is the kinetic energy in joules. In determining the magnitude of the kinetic energy of a projectile, the velocity is more important than the mass.

In a x-ray tube, the projectile is the electron. As its kinetic energy is increased, both the intensity (number of x-rays) and the energy (their ability to penetrate) of the created x-rays are increased.

The x-ray machine is a remarkable instrument. It conveys to the target an enormous number of electrons at a precisely controlled kinetic energy. At 100 mA, for example, 6×10^{17} electrons travel from the cathode to the anode of the x-ray tube every second.

The distance between the filament and the target is only about 1 to 3 cm. Imagine the intensity of the accelerating force required to raise the velocity of the electrons from zero to half the speed of light in so short a distance.

The electrons traveling from the cathode to anode in a vacuum tube comprise the x-ray current and are sometimes called projectile electrons. When the projectile electrons impinge on the heavy metal atoms of the target, they interact with these atoms and transfer their kinetic energy to the target. These interactions occur within a very small depth of penetration into the target. As they occur, the projectile electrons slow down and finally come nearly to rest, at which time they can be conducted through the x-ray anode assembly and out into the associated electronic circuitry.

The projectile electron interacts with either the orbital electrons or the nuclei of target atoms. The interactions result in the conversion of kinetic energy into thermal energy and electromagnetic energy in the form of x-rays.
By far, most of the kinetic energy of projectile electrons is converted into heat. The projectile electrons interact with the outer-shell electrons of the target atoms but do not transfer sufficient energy to these outer-shell electrons to ionize them. Rather, the outer-shell electrons are simply raised to an excited, or higher, energy level. The outer-shell electrons immediately drop back to their normal energy state with the emission of infrared radiation. The constant excitation and restabilization of outer-shell electrons is responsible for the heat generated in the anodes of x-ray tubes.
Generally, more than 99% of the kinetic energy of projectile electrons is converted to thermal energy, leaving less than 1% available for the production of x-radiation. One must conclude, therefore, that, sophisticated as it is, the x-ray machine is a very inefficient apparatus.
The production of heat in the anode increases directly with increasing tube current. Doubling the tube current doubles the quantity of heat produced. Heat production also varies almost directly with varying kVp.
The efficiency of x-ray production is independent of the tube current. Regardless of what mA is selected, the efficiency of x-ray production remains constant. The efficiency of x-ray production increases with increasing projectile-electron energy. At 60 kVp, only 0.5% of the electron kinetic energy is converted to x-rays; at 120 MeV, it is 70%.

Characteristic Radiation

If the projectile electron interacts with an inner-shell electron of the target atom rather than an outer-shell electron, characteristic x-radiation can be produced. Characteristic x-radiation results when the interaction is sufficiently violent to ionize the target atom by total removal of the inner-shell electron. Excitation of an inner-shell electron does not produce characteristic x-radiation.
When the projectile electron ionizes a target atom by removal of a K-shell electron, a temporary electron hole is produced in the K shell. This is a highly unnatural state for the target atom and is corrected by an outer-shell electron falling into the hole

in the K shell. The transition of an orbital electron from an outer shell to an inner shell is accompanied by the emission of an x-ray photon. the x-ray has energy equal to the difference in the binding energies of the orbital electrons involved.

Discrete X-ray Spectrum

We saw earlier that characteristic x-rays have precisely fixed, or discrete energies and that these energies are characteristic of the differences between electron binding energies of a particular element. A characteristic x-ray from tungsten, for example, can have one of fifteen energies and no others.

Bremsstrahlung Radiation

The production of heat and characteristic x-rays involves interactions between the projectile electrons and the electrons of target atoms. A third type of interaction in which the projectile electron can lose its kinetic energy is an interaction with the nucleus of a target atom. In this type of interaction, the kinetic energy of the projectile electron is converted into electromagnetic energy.

A projectile electron that completely avoids the orbital electrons on passing through an atom of the target may come sufficiently close to the nucleus of the atom to come under its influence. Since the electron is negatively charged and the nucleus is positively charged, there is an electrostatic force of attraction between them. As the projectile electron approaches the nucleus, it is influenced by a nuclear force much stronger than the electrostatic attraction. As it passes by the nucleus, it is slowed down and deviated in its course, leaving with reduced kinetic energy in a different direction. This loss in kinetic energy reappears as an x-ray photon. These types of x-rays are called bremsstrahlung radiation, or **bremsstrahlung x-rays**. Bremsstrahlung is the German word for slowing down or braking; bremsstrahlung radiation can be considered radiation resulting from the braking of projectile electrons by the nucleus.

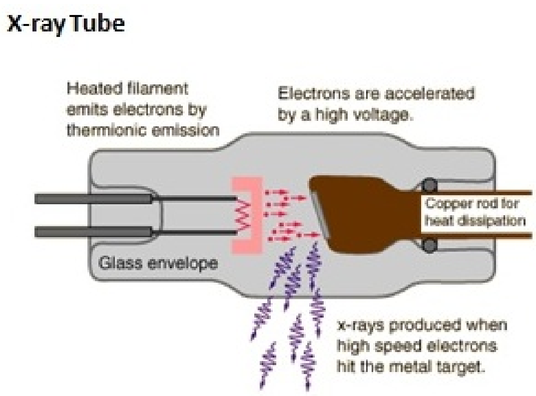
A projectile electron can lose any amount of its kinetic energy in an interaction with the nucleus of a target atom, and the bremsstrahlung radiation associated with the loss can take on a corresponding range of values. For example, an electron with kinetic energy of 70 keV can lose all, none, or any intermediate level of that kinetic energy in a bremsstrahlung interaction; the bremsstrahlung x-ray produced can have an energy in the range of 0 to 70 keV. This is different from the production of characteristic x-rays that have specific energies.

Continuous X-ray Spectrum

If it were possible to identify and quantify the energy contained in each bremsstrahlung photon emitted from an x-ray tube, one would find that these energies extend from that associated with the peak electron energy all the way down to zero. In other words, when an x-ray tube is operated at 70 kVp, bremsstrahlung photons with energies ranging from 0 to 70 keV are emitted. Thus, creating a typical continuous, or bremsstrahlung, x-ray emission spectrum.

This emission spectrum is sometimes called the continuous emission spectrum because, unlike in the discrete spectrum, the energies of the photons emitted may range anywhere from zero to some maximum value. The general shape of the continuous x-ray spectrum is the same for all x-ray machines. The maximum energy that an x-ray can have is numerically equal to the kVp of operation. The greatest number of x-ray photons is emitted with energy approximately one-third of the maximum photon energy. The number of x-rays emitted decreases rapidly at very low photon energies and below 5 keV nearly reaches zero.

X-ray Tube



X-rays for medical diagnostic procedures or for research purposes are produced in a standard way; by accelerating electrons with a high voltage and allowing them to collide with a metal target. X-rays are produced when the electrons are suddenly decelerated upon collision with the metal target; these x-rays are commonly called brehmsstrahlungor “braking radiation”. If the bombarding electrons have sufficient energy, they can knock an electron out of an inner shell of the target metal atoms.

Then electrons from higher states drop down to fill the vacancy, emitting x-ray photons with precise energies determined by the electron energy levels. These x-rays are called characteristic x-rays.

Radiation and radioactive materials are naturally part of our environment. Radiation is energy that moves in a very high speed. It can move through space or through matter. It is common knowledge that radiation can cause adverse effects to humans. **Radiological safety hazards** are potential harmful threats to human health that must be regulated by safety controls and precautions. There are several health problems radiation poses to humans. It starts by breaking chemical bonds that hold molecules together. This then starts the cells of the body to change. The adverse effects of radiation depend upon the dosage and time of exposure of the person to radiation. The most dangerous will probably be getting large doses of radiation in a shorter period of time. A cell can instantaneously die at high radiation doses; therefore, a person can instantaneously die. On lower doses, the cells can repair themselves first and the person will recuperate. However, if the person has a preexisting disease that causes malfunctions with the cell repairs, the person can be in grave danger.

Threshold effects are those immediate noticeable effects due to radiation exposure. These include radiation sickness, cataracts, sterility and fetal effects. Signs of radiation sickness are nausea, vomiting, headache and loss of white blood cells. If the person does not receive medical treatment, there is a chance that he or she will die within 60 days. Hair loss can also be a sign of radiation sickness.

Radiation is highly beneficial in the medical, research, and industrial fields and can even be used in communications as well. Different types of radiation are used in different applications. The two main classifications of radiation are the ionizing and non-ionizing radiations.

Radiological safety hazards are much more evident in the ionizing type of radiation. This is because it carries more amount of energy than the non-ionizing radiation. This type of radiation exists in two forms: the electromagnetic rays or the particles. X-rays, gamma rays, alpha and beta particles are just a few examples of ionizing radiation.

Non-ionizing radiation, on the other hand, refers to two main regions of the electromagnetic spectrum. These are the optical radiation and electromagnetic fields. This type of radiation is used extensively in the manufacturing and telecommunications industry. There also have been no significant related health problems. Despite this, the ultra-violet light, which can be natural or man-made, can still cause health problems like the cancer of the skin.

Radiological safety hazards call for different precautions and safety measures. Different types of radiation call for different types of shielding protection. The amount and intensity of radiation also matters in this. For example, one can be shielded from alpha particles with just a sheet of paper while aluminum sheets can protect you from beta particles. Also, the thicker the shielding protection is, the lesser the intensity of the radiation becomes.
There are three types of radiation protection: occupational radiation protection, medical radiation protection, and public radiation protection. Occupational radiation protection specifically deals with the workers in the different industries that make use of radiation. Medical radiation protection is the protection of the patients and the radiographer. For example, when a patient undergoes an X-ray testing, both the patient and the radiologist take preliminary precautions. Lastly, public radiation protection is directed towards the public, the population.

Questions:
1. In Photoelectric Effect, when we passing a light into a material, the material should emit
A. Electrons B. Protons C. Positrons D. Neutrons
2. ……………………….. amount of energy which is necessary to start photo electric emission is called Work Function
A. Maximum B. Average C. Minimum D. Littlest
3. Threshold frequency is defined as the ……………………. frequency of incident light which can cause photo electric emission
A. Maximum B. Minimum C. Average D. highest
4. Which is not an example of ionizing radiation?
A. X-Rays B. Gamma Rays C. Projectile Gamma Rays D. Alpha and Beta Particles

5. The following parameters are related to Photoelectric effect, except
A. Photoelectric Current B. Threshold Energy C. Work Function D. Frictionless Electrons

Answers:
1. A 2. C. 3. B 4. C 5. D

WEEK 8

Physics SS3

Topic: CONDUCTION OF ELECTRICITY IN GASES

INTRODUCTION

Electrons: When a high tension discharge passes between electrodes sealed into a partially evacuated vessel, the gas becomes luminous showing a series of highly colored glows which are often very beautiful. If the pressure is sufficiently reduced, a series of streams appears, proceeding in straight lines from the cathode. These streams are known as “cathode rays,” and are found to be independent of the position of the anode, and often penetrate regions occupied by other glows in the tube. The researches of modern physics have shown that these rays are streams of discreet particles of negative electricity, called “electrons.” Their properties do not depend upon the material of the electrodes nor the nature or presence of the gas through which the discharge takes place. They may be produced from all chemical substances, and consequently must play an important part in the structure of matter. The velocities with which they move through the tube vary from one-thirtieth to one-third that of light. The ratio of the charge of an electron to its mass is constant and is equal to 1.77×10^{7}electromagnetic units per gram. The charge of an electron is 1.5×10^{-20} electromagnetic units and Themass is about 1/1800 thatof the hydrogen atom. The radius of an electron is estimated, at 1.9×10^{-13}c.m.s, which is about 1/50 000 that of the atom. Formany years the mass has been regarded as purely electromagnetic in character; that is, while exhibiting inertia, it shows no gravitational attraction in the sense possessed by ordinary matter. Recently, however, certain experimental and theoretical evidence has been produced which makes it appear likely that this cannot be entirely the case. Many attempts have been made to discover evidence of quantities of electricity smaller or larger than the electron, but none smaller have ever been found. In fact, when quantities comparable to the electron have been isolated, they have always proved to be exact integral multiples of it. The evidence points to the conclusion that electricity is atomic in structure and that the smallest possible element is the electron, which thus constitutes our natural unit of electricity. Electric currents through conductors, as we know them in every day practice, are simply streams of electrons through or between the atoms and molecules making up the conducting body.

Conductivity of Gases: A gas in its normal state is one of the best insulators known. This may be shown by mounting a gold leaf electroscope inside an enclosed space, and allowing only a small rod carrying a polished knob, for the purpose of charging, to project out. If the support carrying the electroscope is well insulated from the container, the electroscope will remain charged for a long time, showing that the air or whatever gas surrounds the electroscope is a poor conductor of electricity. If, however, X-rays are allowed to shine through the enclosure, or if a small quantity of some radioactive substance such as thorium or radium is placed inside it, or again if the products of combustion of a flame are drawn through it, it is then found that the gold leaves collapse quite rapidly, indicating that the gas has lost its insulating properties. That the leakage has taken place through the air and not across the insulating support may be shown by using a second chamber connected with the electrometer enclosure by a glass tube, and introducing the X-rays, the radioactive substance or other agent into this, and then drawing the air thus acted upon into the first chamber. The same effects are observed. However, if glass wool is introduced in the connecting tube, or if the air is passed between two insulated plates connected to a battery before entering the electrometer chamber, it is found that its insulating properties are restored. Experiments of this sort as well as many others of an entirely different nature have shown that the conduction of electricity through gases is due to carriers of electricity, and that the carriers are of two distinct types, positive and negative; the former are similar to the carriers of electricity through solutions and are called positive ions, while the latter are either negative ions or electrons.

Cathode Rays

When the pressure in the discharge tube is less than 10^{-4} mm of Hg, the discharge tube starts showing fluorescence. When this fluorescence was investigated, it was found that the fluorescence consisted of beams of negatively charged electrons.

These electrons emanate normally from the cathode. As these emanate from the cathode, the rays are called the Cathode Rays.
We know that hydrogen atom is the lightest atom. In 1869, Sir J.J. Thomson found that the electrons had mass far less than of even the hydrogen atom. Experiments showed that mass of the electron is approximately 1/1860 of mass of hydrogen atom.
The cathode, in the discharge tubes used by Sir Thomson, was cold cathode. It needed a high voltage of the order of 30,000 volts. Modern cathode ray tubes have hot cathodes which require much less voltage, nearly 3000 volts.

Cathode Ray Oscilloscope

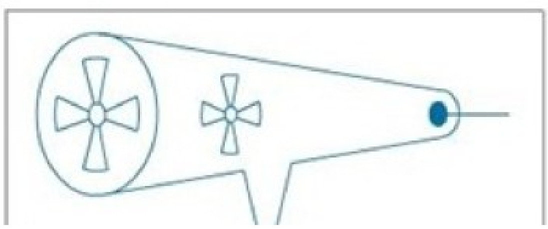
The discovery of cathode rays led to a vast field of practical application in Electronics. The glow produced by the fast moving electrons on a fluorescent screen led to its use in radar and television.
The oscilloscope has many points in common with the discharge tube. In the oscilloscope, the electrons are emitted by a hot cathode which is situated in a highly evacuated tube. At a short distance from the cathode is an anode having a central hole in it. A potential difference of some hundreds of volts is applied between cathode and anode. As a result, the electrons accelerate across the gap between the electrodes and a narrow stream of the electrons emerges from the hole in the anode.
Such arrangement of electrodes where a stream of electrons is produced is often known as the electron gun. On leaving the gun, the electron stream passes across the tube and eventually hits the screen at the far side. The screen is coated with phosphorus.
If necessary the stream of emerging electrons can be deflected in its passage between the gun and the screen. This deflection is produced by two pairs of parallel plates arranged at right angles. Usually, the potential difference applied to the X-plates makes the spot move across the screen at a uniform speed. If we alter the potential difference between the Y-plates, the beam is deflected upwards or downwards on the screen.

Oscillating voltages applied across the X and Y plates cause deflection of the spot on the fluorescent screen.

The fluorescent screen at the end of tube is coated with a mixture of fluorescent material and phosphorescent material. The phosphorescent material is responsible for the persistence of the image on the screen.

Properties of Cathode Rays

Cathode rays travel in straight lines from the cathode to anode and cast shadow of the object placed in their path.



(a) Cathode rays travel in straight lines.

Cathode rays consist of an invisible stream of negatively charged particles.

When the plates X and Y are given +ve and -ve potentials respectively, the cathode ray beams gets attracted to the positive plate. This proves that they are made up of negatively charged particles.

(b) Cathode rays get deflected in the electric field.

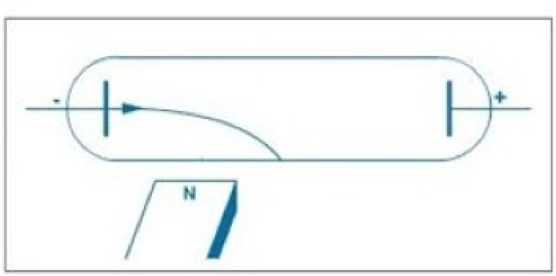
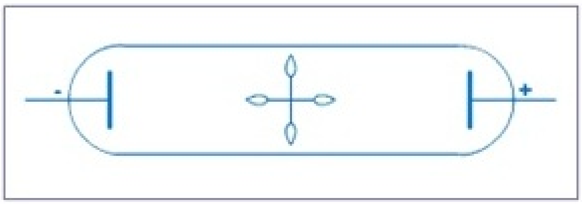


Figure (c) above shows a beam of cathode rays being deflected by the magnetic field. By applying ‘Fleming’s left hand rule’ we can also prove that they are negatively charged particles.
Cathode rays travel with a great velocity nearly 9/10th of the speed of light and hence they possess great kinetic energy. When cathode rays are made to fall on a mica paddle wheel, the wheel starts rotating. This experiment proves that the rays possess great amount of kinetic energy.



(d) Due to the impact of cathode rays, the paddle wheel rotates with great speed.

Cathode rays can ionize gases.
Cathode rays can penetrate through thin sheets of aluminium and silver without perforating them.
They can produce fluorescence in many substances.
When they impinge on a metal of high atomic weight, X-rays are produced.
Measurements of their deflection by electric and magnetic fields show that they consist of charged particles whose specific charge is (charge to mass ratio) $1.76 \times 10^{11}\ C\ kg^{-1}$. In fact they are electrons which are constituent of all matter.

Uses of Cathode Ray Tubes
They are widely used in science research laboratories by scientists for converting electrical signals into visual signals and television tubes.
Doctors use them for converting electrical impulses corresponding to heart beats into visual signals.

Thermionic Emission and its Applications
Thermionic emission, also known as thermal electron emission, is the process by which charge carriers, such as electrons or ions, move over a surface or some sort of energy barrier by the induction of heat. Charge carriers naturally restrain activity; however, in thermionic emission, thermal energy is introduced to the carriers,

causing them to overcome these forces. The reason behind the charge carriers' ability to perform this action is because electrons and ions are mobile and unbound to the normal chains of atomic structure that affect other particles. Traditionally, these charge carriers were referred to as "thermions."

One property of the thermionic emission theory is that the emitting region is sustained with a charge opposite to the original but equal in magnitude. This means that the location of the charge carrier prior to emission will generate a positive charge in the case of electrons. However, this can be altered using a battery. The emission is neutralized when the carriers are farther away from the region, resulting in no change to the original state.

Historically, the primary example of thermionic emission is that used in the Edison effect. Electrons are emitted from a hot metal cathode, which uses a polarized electrical device to cause electrical current to flow out into a vacuum tube. This allows a device to maintain control over the movement of the electrons and amplify or modify the electrical signal.

Anything used for either cooling or generating power utilizes the concept of thermionic emission theory. As temperature increases, the magnitude of the flow increases. Besides the traditional use of vacuum tubes for electronics, solid-state devices can also be used to create the thermionic movement of electrons, allowing modern technology to function.

Thermionics was first reported by Frederick Guthrie in 1863. He was able to identify an alteration in the positive charge of a highly heated iron sphere that did not occur if the object was negatively charged. However, it wasn't until 1880 that the science was readily harnessed by Thomas Edison. When working with his incandescent light bulbs, he noticed that certain areas remained darkened. This allowed him to identify the flow of electrons due to heat, resulting in the creation of the diode.

Richardson's law describes the reason electrons are able to flow in this manner. Specifically, metals contain two electrons in the atomic structure that are able to move from atom to atom. In 1928, Sir Owen Willans Richardson, a British physicist, found that some electrons were able to leave the atom without returning. This process requires a certain amount of energy depending on the metal. The term for this effect is *work function*.

Questions:

1. When a metal is heated to a high temperature, electrons are emitted from its surface. This process is known as

A. photoelectric emission B. field emission C. artificial radioactivity D. thermionic emission

2. Which of the following is not true of a discharge tube?

A. The pressure of gas must be very low B. The glass tube must have two electrodes at the end C. The gas in the tube must be carbon dioxide D. Air is gradually pumped out of the tube.

Answers:

1. D. 2. C

WEEK 9

SS3 Physics Second Term

Topic: WAVE NATURE OF MATTER

Electron Diffraction

The wave nature of X-rays was established by X-ray diffraction experiments. In the same way the Davisson and Germer experiment established the wave nature of electrons.

In the Davisson and Germer experiment, a beam of electrons emitted from a heated filament was made to impinge on a layer of a thin metal film or crystal at C. The electrons were diffracted and the diffraction rings were produced on a photographic plate placed behind the thin metal film as shown below.

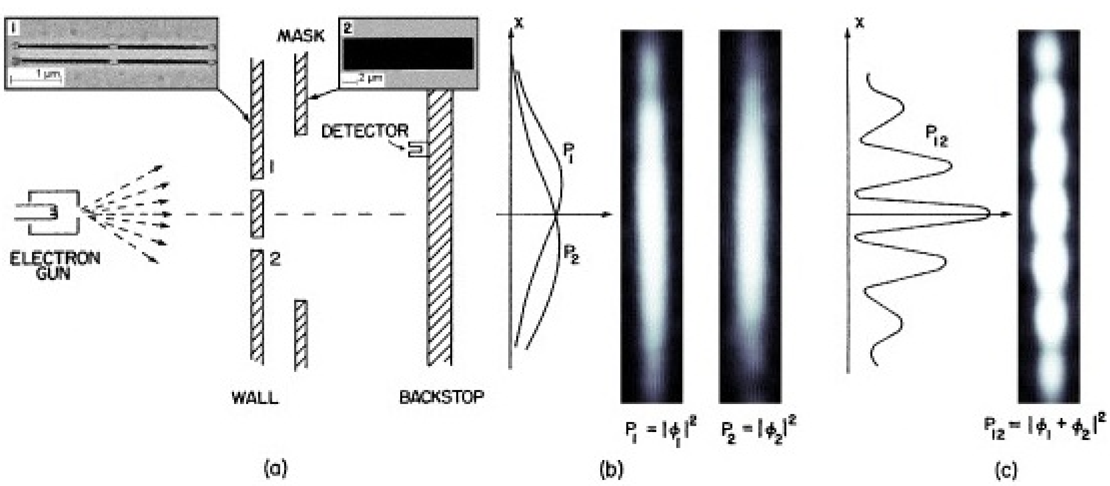
If the voltage V on the anode was increased, the velocity, v, of the electrons was increased. The rings were then seen to become narrower. Hence the wavelength, λ, of the electron waves decreases with increasing electron velocity.

The wavelength of a material object is given by

$\lambda = h/mv$

where mv is the momentum of the object and h is the Planck's constant.

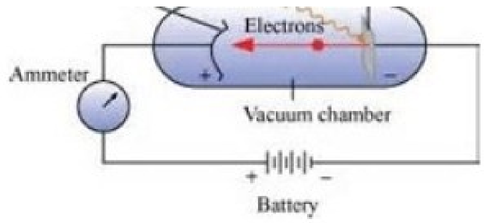
Later experiments showed that protons, neutrons and other particles also have the wave property of diffraction.



Particle Nature of Matter

a. Photoelectric Effect

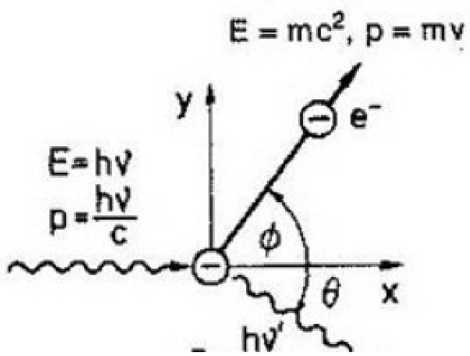
In the photoelectric effect, it was shown that when light falls on a surface, electrons are emitted from the surface. Similarly, when X-ray is allowed to fall on the surface of a thin sheet of metal like gold, the X-ray not only produces diffraction pattern but also acting like particles they may collide with the atoms of the metal and eject electrons as in the photoelectric effect.



Photoelectric effect

b. The Compton Effect

When a single X-ray photon collides with a free electron recoils off as though it were a perfectly elastic sphere.



This is the Compton Effect. In this effect, the scattered photon has a slightly lower frequency than the incident X-ray photon. In
this phenomenon, matter in form of X-rays is shown to behave as a particle. The recoiling photon and electron are able to conserve energy and momentum.

Wave Particle Duality

From the foregoing discussions, we seem to be in a dilemma as to the nature of matter. Some experiments e.g. electron diffraction indicate that matter behaves like a wave, but other phenomena such a s photoelectric and Compton effect experiments indicate that matter behaves like a stream of particles or photons.

These two theories seem to be incompatible but both have shown to have validity. Thus matter appears to have a dual nature. This is referred to as the wave particle duality or the wave particle paradox.

As in matter, so is it with light. Some observable phenomenon in the nature of light, such as reflection, refraction, diffraction, interference and polarization can be interpreted or explained by assuming that light (or matter) behaves like waves. But other observable phenomena such as emission and absorption of light, photoelectricity, radiation of energy from heated bodies, thermionic emission can only be understood by assuming the particle nature of matter.

This dual nature of matter is known as the wave-particle duality or the wave particle paradox.

The wave particle duality refers to the idea that light and matter 9such as electrons) have both wave and particle properties, that is, light behaves either as a wave or as a particle but not as both simultaneously.

The Uncertainty Principle

In general, the process of making a measurement tends to alter the quantity being measured. This is very much pronounced in very small scales such as the atomic and the nuclear scales. As was first pointed out by Heisenberg, It is impossible in principle to make precise measurement of both the position (x) and momentum (p) of a particle simultaneously. In fact the more precisely the location of the particle *has* to be specified, the more uncertainty is introduced into the determination of its momentum and conversely too. Any such measurements have inbuilt uncertainties Δx in the position and Δp in the momentum. Such measurements can only be expressed as probabilities. Heisenberg showed that:

$\Delta x \,.\, \Delta p \geq h$

$\Delta x \,.\, \Delta v \geq h$

$\Delta E \,.\, \Delta t \geq h$

here ΔE, Δt, Δp and Δx are the uncertainties in the energy, time, momentum and position measurements.

1. Momentum and position,
2. Energy and time,
3. Position and velocity, are known as complementary variables.

Heisenberg Uncertainty Principle states that it is impossible to know accurately the exact position and momentum of a particle simultaneously. The uncertainty in the momentum multiplied by the uncertainty in the position approximately equals the Planck's constant, *h*.

Because of the extremely small value of h (=6.63×10^{-34} Js) the uncertainty principle is of no consequence for objects above atomic sizes.

Question

A photon with a wavelength of 6.00×10^{-12} m collides with an electron. After the collision the photon's wavelength is found to have been changed by exactly one Compton wavelength (2.43×10^{-12} m).

1. What is the photon's wavelength after the collision?

A. 3.57×10^{-12} m B. 8.43×10^{-12} m C. 6.00×10^{-12}D. It could be either one of the above

Answer

1. A

<u>WEEK 10</u>

Physics SS 3

Topic: Radioactive Decay

Introduction

Radioactive decay is the spontaneous radioactive disintegration of an atomic nucleus, resulting in the release of energy. Some atoms are stable. Others are unstable and ‘decay’, emitting radiation to achieve a stable state. The emissions from an unstable atom’s nucleus, as it decays, can be in the form of alpha, beta or gamma radiation.

When an atom decays, it changes into another isotope, or form, of the same element or into a completely different element, in a process called transmutation. Different isotopes of the same element differ in the number of neutrons in their nuclei. Some elements reach stability via a series of steps through several isotopes, or ‘daughter products’.

One example is uranium-238 (U-238), which, through the process of radioactive decay, will eventually become a stable isotope of lead. However, this process takes billions of years. Along the way, as the U-238 isotope’s initial energy declines, it will transmute via a series of elements, each more stable than the last – thorium, radium, radon, polonium and bismuth – before it stabilizes as lead.

Alpha decay

In alpha decay, a positively-charged particle is emitted from the nucleus of an atom. This alpha particle consists of two protons and two neutrons (the same structure as a helium-4 nucleus). Although alpha particles are normally highly energetic, they travel only a few centimeters in air and are stopped by a sheet of paper or the outer layer of dead skin.

Beta decay

In beta decay, a particle is emitted from the nucleus of an atom. This beta particle is an electron with either negative or positive electric charge. Beta particles may travel metres in air and several millimetres into the human body. Most beta particles may be stopped by a small thickness of a light material such as aluminium or plastic.

Gamma decay

Gamma decay occurs because the nucleus of an atom is at too high an energy state. The nucleus 'falls down' to a lower energy state, emitting a high energy photon known as a gamma particle in the process. Gamma particles travel in a wave-like pattern at the speed of light. They can only be stopped by a dense material such as lead, steel, concrete or several metres of water.

Half Life of Radioactive Elements

The half-life of a radioactive element is the time that it takes for one half of the atoms of that substance to disintegrate into another nuclear form. The decay of an isotope can be measured by its half life. These can range from mere fractions of a second, to many billions of years.

Element	Most Stable Isotope	Half-life of Most Stable Isotope
Polonium	Po-209	102 years
Astatine	At-210	8.1 hours
Radon	Rn-222	3.82 days
Radium	Ra-226	1600 years
Thorium	Th-229	7.54×10^4 years
Uranium	U-236	2.34×10^7 years
Protactinium	Pa- 234	1.18 minutes

Example Rate of Radioactive Decay Problem

$^{226}_{88}Ra$, a common isotope of radium, has a half-life of 1620 years. Knowing this, calculate the first order rate constant for the decay of radium-226 and the fraction of a sample of this isotope remaining after 100 years.

Solution

The rate of radioactive decay is expressed by the relationship: $k = 0.693/t_{1/2}$

Where k is the rate and $t_{1/2}$ is the half-life.

Plugging in the half-life given in the problem: $k = 0.693/1620$ years $= 4.28 \times 10^{-4}$/year

Radioactive decay is a first order rate reaction, so the expression for the rate is:

$\log_{10} X_0/X = kt/2.30$

Where X_0 is the quantity of radioactive substance at zero time (when the counting process starts) and X is the quantity remaining after time *t*. *k* is the first order rate constant, a characteristic of the isotope that is decaying. Plugging in the values:

$\log_{10} X_0/X = (4.28 \times 10^{-4}/\text{year})/2.30 \times 100 \text{ years} = 0.0186$

Taking antilogs: $X_0/X = 1/1.044 = 0.958 = 95.8\%$ of the isotope remains

Nuclear Reactions

Rutherford in 1919 transmitted nitrogen isotope into an oxygen isotope. The nitrogen was subjected to the action of swift alpha – particles derived from radium salt.

Transmutation is the process by which radioactive elements change into different elements.

Nuclear reaction is a process in which two nuclei or nuclear particles collide, to produce different products than the initial particles.

Nuclear fission and nuclear fusion both are nuclear phenomena that release large amounts of energy, but they are different processes which yield different products. Learn what nuclear fission and nuclear fusion are and how you can tell them apart.

Nuclear Fission

Nuclear fission takes place when an atom's nucleus splits into two or more smaller nuclei. These smaller nuclei are called fission products. Particles (e.g., neutrons, photons, alpha particles) usually are released, too. This is an exothermic process releasing kinetic energy of the fission products and energy in the form of gamma radiation. Fission may be considered a form of element transmutation since changing the number of protons of an element essentially changes the element from one into another.

Nuclear Fission Example:
$^{235}_{92}U + ^{1}_{0}n \rightarrow ^{90}_{38}Sr + ^{143}_{54}Xe + 3^{1}_{0}n$

Nuclear Fusion

Nuclear fusion is a process in which atomic nuclei are fused together to form heavier nuclei. Extremely high temperatures (on the order of 1.5 x 10^7°C) can force nuclei together. Large amounts of energy are released when fusion occurs.

Nuclear Fusion Examples

The reactions which take place in the sun provide an example of nuclear fusion:

$^{1}_{1}H + ^{2}_{1}H \rightarrow ^{3}_{2}He$

$^{3}_{2}He + ^{3}_{2}He \rightarrow ^{4}_{2}He + 2^{1}_{1}H$

$^{1}_{1}H + ^{1}_{1}H \rightarrow ^{2}_{1}H + ^{0}_{+1}\beta$

Comparison between Nuclear Fission and Fusion

	Nuclear Fission	Nuclear Fusion
Definition:	Fission is the splitting of a large atom into two or more smaller ones.	Fusion is the fusing of two or more lighter atoms into a larger one.
Natural occurrence of the process:	Fission reaction does not normally occur in nature.	Fusion occurs in stars, such as the sun.
Byproducts of the reaction:	Fission produces many highly radioactive particles.	Few radioactive particles are produced by fusion reaction, but if a fission "trigger" is used, radioactive particles will result from that.
Conditions:	of the substance and high-speed neutrons are required.	High density, high temperature environment is required.
Energy Requirement:	Takes little energy to split two atoms in a fission reaction.	Extremely high energy is required to bring two or more protons close enough that nuclear forces overcome their electrostatic repulsion.

	Nuclear Fission	Nuclear Fusion
Energy Released:	The energy released by fission is a million times greater than that released in chemical reactions; but lower than the energy released by nuclear fusion.	The energy released by fusion is three to four times greater than the energy released by fission.
Nuclear weapon:	One class of nuclear weapon is a fission bomb, also known as an atomic bomb or atom bomb.	One class of nuclear weapon is the hydrogen bomb, which uses a fission reaction to “trigger” a fusion reaction.

Uses of Radioactivity

1. Radioactivity tracers are commonly used in the medical field and also in the study of plants and animals.
2. Radiation is used and produced in nuclear reactors, which controls fission reactions to produce energy and new substances from the fission products.
3. Radiation is also used to sterilize medical instruments and food.
4. Radiation is used by test personnel who monitor materials and processes by non-destructive methods such as x-rays.

Comparison of Nuclear Reaction and Ordinary Chemical Reaction

Nuclear Reaction	Ordinary Chemical Reaction
During nuclear reactions, the nuclei of atoms undergo change and therefore new elements are formed as a result of such reactions.	During chemical reactions, elements do not lose their identity. In these reactions, only the electrons in the outermost shell of atoms participate whereas the nuclei of atoms remain unchanged.
Reactivity of an element towards nuclear reactions is nearly independent of oxidation state of the element. For	Reactivity of an element towards chemical reactions depends upon the oxidation state of the element. In

example, Ra element or Ra^{2+} ion in RaC_2 behave s similarly during nuclear reactions.	ordinary chemical reactions, Ra and Ra^{2+} behave quite differently.
In nuclear reactions, isotopes behave quite differently. For example, U-235 undergoes fission quietly readily but U-238 does not.	Different isotopes of an element have nearly same chemical reactivity.
Rate of a nuclear reaction is independent of temperature and pressure.	Rate of a chemical reaction is largely affected by temperature and pressure.
A nuclear reaction cannot be reversed.	A chemical reaction can be reversed.
Nuclear reactions are accompanied by large energy changes.	Chemical reactions are accompanied by relatively small energy changes.

ASSESSMENT

1. Activity of one decay per second is equal to
   (a) 1 Bq
   (b) 1 atm
   (c) 1 mol
   (d) 1 Cd
2. Greater decay constant
   (a) less activity
   (b) greater activity
   (c) greater size
   (d) less size

- Radioactive decay is
  (a) a physical change
  (b) a chemical change

(c) a nuclear change
(d) none of the above4

- The adiabatic throttling process of a perfect gas is one of constant enthalpy in which
  (a) there is no drop in temperature
  (b) there is a reduction in temperature
  (c) there is an increase in temperature
  (d) none of the above
- The extinction coefficient for different qualities of glasses varies in the range of
  (a) 5-25 m-1
  (b) 5-25 cm-1
  (c) 5-25 mm-1
  (d) 5-25 km-1

ANSWERS

1. a
2. b
3. c
4. a
5. a