ENVIROMENTAL PHYSICS

(i) Agriculture physics

- Influence of solar radiation on plant growth.
- Influence of wind, humidity, rainfall and air temperature on plant growth.
- Soil environmental component which influence plant growth.

(ii) Energy from the environment

Photovoltaic energy

Wind energy

Geothermal energy

Wave energy

(iii) Geophysics (Earth quakes)

Elastic rebound theory

Types of seismic waves

Propagation of seismic waves

Seismology

(iv) Environmental pollution

Types of pollutant in the atmosphere

Transport mechanisms of atmospheric pollutant

Nuclear waste and their disposal

Effects of pollution on visibility and optical properties of materials.

INTRODUCTION

Environmental physics is an interdisciplinary subject that integrates the physics processes in the following disciplines: the atmosphere, the biosphere, the hydrosphere, and the geosphere.

Environmental physics can be defined as the response of living organisms to their environment within the framework of the physics of environmental processes and issues.

It is structures within the relationship between the atmosphere, the oceans (hydrosphere), land (lithosphere), soils and vegetation (biosphere).

It embraces the following themes:

(i) Human environment and survival physics,

(ii) Built environment

(iii) Renewable energy

(iv) Remote sensing

(v) Weather, climate and climate change, and

(vi) Environmental health.

The environment may be defined as the medium in which any entity finds itself, For example, for a cloud its environment may be the region of the atmosphere in which it is formed.

AGRICULTURE PHYSICS

Agriculture physics is concerned with physics environment in relation to plant growth.

(a) Influence of Radiation Environment on Plant Growth

Radiation environments. Refer to radiations present in the atmosphere, commonly coming from the sun.

Components of solar radiation

The main components of solar radiation are:

(i) Visible light

(ii) Infrared radiation, and

(iii) Ultraviolet radiation.

HEATING EFFECT OF SOLAR RADIATION ON PLANTS

Positive effect

An optimum amount of heat on plant favours the process of photosynthesis. This enables a plant to make its own food and hence provide its growth.

Negative effects

(i) Excessive solar radiation (ultraviolet light) on plants leads to bleaching of green pigment (chlorophyll). This lowers the amount of food produced by photosynthesis to plant and hence a plant may die.

(ii) Excessive solar radiation on plants leads to excessive water loss in the form of water vapour commonly on plant leaves (transpiration). Hence wilting (drying) of plants may occur.

(b) Influence of Aerial Environment on Plant Growth

Aerial environments refer to the atmospheric condition resulting from a series of processes occurring in the atmosphere. These include air temperature, wind, humidity and rainfall.

WIND EFFECT ON PLANT GROWTH

Positive effects

(a) Wind acts as pollinating agent for some plants and hence favours plant productivity.

(b) Wind also favours evaporation of water from plant leaves and thus maintains water balance for proper plant growth.

Negative effects

(a) Excessive wind on environments leads to plant breaking or cutting of tree branches. This may lead to the death of plant.

(b) As the wind speed increases further, cell and Cuticular damage occurs, followed by death of plant tissue, and a gnarled appearance becomes more apparent.

(c) At low wind speeds, the effect seems to be an increase in transpiration, which results in water stress. This stress causes the plant to adapt by decreasing leaf area and internodes length, while increasing root growth and stem diameter.

(d) Strong wind may also cause shade off flowers; this lowers plant productivity.

Effect of Rainfall on Plant Growth

Positive effect

An optimum amount of rainfall on plants favours its growth. Water is a raw material for the process of photosynthesis from which plants obtain their food and hence their growth.

Negative effect

Excessive rainfall leads to water logging in soil which in turn leads to root spoil and hence the death of plant.

Effect of Humidity on Plant Growth

Positive effect

Favourable humidity on plants help plants to conserve water for various activities and in seeds helps the development of new leaves.

Negative effect

Low humidity results into a greater rate of transpiration and hence may result into plant drying.

Effect of Air Temperature of Plant Growth

Positive effect

An Optimum temperature on plants enhances enzymic activities which in turn gives favourable conditions for plant growth.

Negative effect

(a) High temperature denature enzymes commonly for photosynthesis and hence the death of plant.

(b) Low temperature inactivates the plant growth enzymes, hence low growth rate.

Wind Belts

Wind belts are seasonal strong wind moving in a specified direction in a certain region of the earth.

The global wind belts are formed by two main factors:

(i) The unequal heating of the earth by sunlight and

(ii) The earth's spin.

Here is a simple explanation of the process

The unequal heating makes the tropical regions warmer than the Polar Regions. As a result, there is generally higher pressure at the poles and lower at the equator. So the atmosphere tries to send the cold air toward the equator at the surface and send warm air northward toward the pole at higher levels.

Unfortunately, the spin of the earth prevents this from being a direct route, and the flow in the atmosphere breaks into three zones between the equator and each pole.

These form the six global wind belts: 3 in the Northern Hemisphere (NH) and 3 in the Southern (SH). They are generally known as:

(1) The Trade winds, which blow from the northeast (NH) and southeast (SH), are, found in the sub tropic regions from about 30 degrees latitude to the equator.

(2) The Prevailing Westerlies (SW in NH in SH) which blow in the middle latitudes.

(3) The Polar Easterlies which blow from the east in the Polar Regions.

Effects of wind belts to plant

1. Wind belts because the loss of plant leaves and flowers hence lower plant productivity and growth. Loss of leaves lowers the rate of photosynthesis.

2. Wind belts sometimes cause plants to lean in direction of moving wing. This changes their direction of growth

3. Trees are broken by the strong wind.

(c) Soil Environment Components Which Influence Plant Growth

Soil is composed of both rock particles and organic matter (humus) – the remains of plants and animals in various stages of decomposition. The humus serves as food for many living organisms. Within the soil is a large population of animals, plants. These break down the humus into soluble substances that can be absorbed by the roots of large plants.

Components of a soil

Soil is composed of:

((a) Air, 25% by volume which supports life of soil organisms,

(b) Water, 25% which dissolves minerals so that are easily absorbed by plants,

(c) Organic matter (humus), 5% by volume,

(d) Inorganic matter (minerals), 45% by volume,

(e) Biotic organisms, micro – organisms like earth worm, centipedes, millipede, bacteria which decompose organic matter.

Types of soil

(i) Sandy soil

(ii) Silt soil,

(iii) Clay soil, and

(iv) Loamy soil (sand + silt + clay soil mixture)

Water Movement in the soil

Two forces primarily affect water movement through soils, (a) gravity and (b) capillary action.

Capillary action refers to the attraction of water into soil pores – an attraction which makes water move in soil. Capillary action involves two types of attraction – adhesion and cohesion.

Adhesion is the attraction of water to solid surfaces.

Cohesion is the attraction of water to itself.

Speed of water in a particular soil type depends on:

(i) How much water is in the soil, and

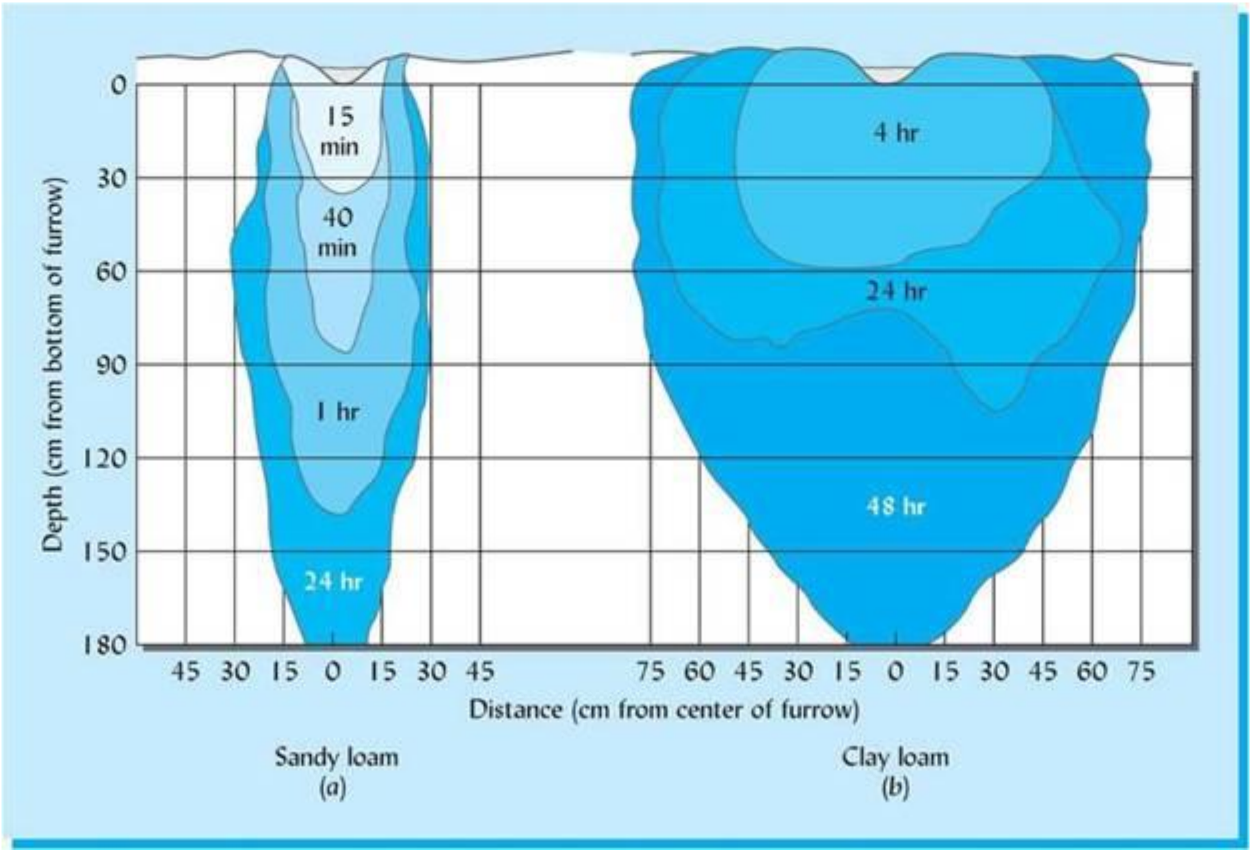
(ii) Porosity of the soil.

The movement of water in the solid is mainly due to gravity. The porosity gives a measure of how much water the soil can hold and the rate at which water flows through the soil. Large pore spaces give a faster rate and vice versa.

An experiment to study water movement in soil

An experiment to demonstrate the rate of flow of water in the soil is done using a glass tube and sand type filled in it. Water is poured into the tube and the time taken for water to reach the bottom of the tube in notes.

Unsaturated soil		**Saturated soil**	
Soil type	Water speed	Soil type	Water speed
Sandy	Fastest	Sand	Slowest
Loamy	Moderate	Loamy	Moderate
Clay	Slowest	Clay	Highest



i. Sand soil have large pore spaces thus allows water to travel downwards through it at a fastest rate.

ii. Clay soil can hold water as has very fine pore spaces.

iii. Loamy soil allows water movement at a medium rate.

Heat transfer in the soil

Within the soil heat is transferred by a conduction process. Since soil is poor conductor of heat most of the heat from the atmosphere appears at the surface of the earth.

An optimum soil temperature favours plants growth but a high temperature can lead to the rotting of plant roots.

(d) Techniques for the Improvement of the Plant Environment

Plant environment can be improved by using wind breaks, shading and mulching.

Shading

Shading is the process of obstructing plants from excessive solar radiation.

Positive Impacts of Shading

1. Prevents excessive loss of water by plants through transpiration. This enhances plant productivity.

2. Preserve moisture in the soil and hence water supply to plant.

Mulching

Mulching is the process of covering the soil by dry leaves, grasses and or papers.

Benefits (Advantages) of Mulching

1. **Improve soil moisture**. Bare soil is exposed to heat, wind and compaction loses water through evaporation and is less able to absorb irrigation or rainfall. Using mulches, the soil has greater water retention, reduced evaporation, and reduced weeds. Mulch can also protect trees and shrubs from drought stress and cold injury

2. **Reduce soil erosion and compaction**. Mulches protect soils from wind water, traffic induced erosion and compaction that directly contribute to root stress and poor plant health.

3. **Maintenance of optimal soil temperatures**. Mulches have shown to lower soil temperatures in summer months. Extreme temperatures can kill fine plant roots which can cause stress and root rot. Mulches protect soils from extreme temperatures, either cold or hot.

4. **Increase soil nutrition**. Mulches with relatively high nitrogen content often result in higher yields, but low nitrogen mulches, such as straw, sawdust and bark, can also increase soil fertility and plant nutrition.

5. **Reduction of salt and pesticide contamination**. In arid landscapes, evaporating water leaves behind salt crusts. Because mulches reduce evaporation, water is left in the soil and salts are diluted. Organic mulches can actively accelerate soil desalinization and help degrade pesticides and other contaminants.

6. **Improve plant establishment and growth**. Mulches are used to enhance the establishment of many woody and herbaceous species. Mulches improve seed germination and seed survival, enhance root establishment, transplant survival, and increase plant performance.

7. **Reduction of disease**. Mulches will reduce the splashing of rain or irrigation water, which can carry spores of disease organisms to stems and leaves of plants. Populations of beneficial microbes that reduce soil pathogens can be increased with mulches. Mulches can combat disease organisms directly as well.

8. **Reduction of Weeds**. Using mulches for weed control is highly effective. Mulches can reduce seed germination of many weed species and reduce light, which stresses existing weeds.

9. **Reduce pesticide use**. Mulches reduce weeds, plant stress, and susceptibility to pests and pathogens which translates to reduced use of herbicides, insecticides, and fungicides.

Mulch Problems (disadvantages of mulching)

1. **i. Acidification.** Some types of mulches can increase soil acidity.

2. **ii. Disease.** Many mulches made from diseased plant materials can be composted or treated at temperatures that kill pathogens that can be transmitted to healthy plants.

3. **iii.Pests.** Many organic mulches, especially wood – based mulches, have the reputation as being "pest magnets".

4. **iv. Weed contamination.** Improperly treated crop residues and composts as well as bark mulches are often carriers of weed seed. Mulch must be deep enough to suppress weeds and promote healthy soils and plants. Weed control and enhanced plant performance are directly linked to mulch depth.

v. Wind Breaks

Wind breaks are long rooted strong plants (trees) that are used to obstruct the path of wind or to slow down the wind.

Windbreaks provide many benefits to soil, water, plants, animals and man. They are an important part of the modern day agricultural landscape. Windbreaks come in many different sizes and shapes to serve many different conservation purposes.

In agriculture, wind breaks protect small growing plants from strong blowing wind

Advantages of Windbreaks to Plant Environment

1. **i. Control soil erosion.** Windbreaks prevent wind erosion from causing loss of soil productivity. This eliminates plant roots stresses and thus favours plant growth condition.

2. **ii. Increase plant yield.** Windbreak research substantiates that field windbreaks improve crop yields which offsets the loss of production from the land taken out of cultivation.

3. **Pesticide sprays.** Windbreaks control pesticide spray drift and provide buffers to delineate property lines and protect neighbors.

EXAMPLES: SET A

Example 01

(a) What is agriculture physics? (02 marks)

(b) What are the components of a soil? How do they support the life of a plant? (06 marks)

(c) Explain briefly how soil temperature affects plant growth. (02 marks)

Example 02

(a) What do you understand by the word environmental physics? (01 marks)

(b) Explain how the following climatic factors influence plant growth: air temperature, humidity, rainfall and wind. (06 marks)

(c) What are wind belts? Explain the effect of wind belts on plant productivity. (03 marks)

Example 03

(a) What is mulching? (02 marks)

(b) Give two advantages and two disadvantages of mulching. (04 marks)

(c) Discuss the heating effect of solar radiation to plant growth. (04 marks)

Example 04

(a) Explain two factors that primarily affect water movement in the soil (03 marks)

(b) Explain the soil environment that favours high crop yield (04 marks)

(c) What is shading and what is its purpose? (03 marks)

Example 05

(a) (i) Mention the components of solar radiation.
(ii) How do those components affect plant growth? (04½ marks)

(b) What are wind breaks? (02 marks)

(c) What are the advantages of wind breaks to plant environment? (03½ marks)

ENERGY FROM THE ENVIRONMENT

ENERGY

Energy is defined as the capacity to do work Or is defined as ability to do work.

Energy is measured in **Joules** (symbol J)

Types of energy according to their usefulness

(i) High grade energy

(ii) Low grade energy

i. High grade energy is the energy that is easily transformed into other forms of energy and is more suitable for doing works.

Examples are chemical and electrical energy.

ii. Low grade energy is the one that is not easily transformed into anything else.

Examples are the kinetic energy of molecules due to their randomness and the potential energy due to the forces between molecules.

ENERGY SOURCES

There are two types of energy sources, namely:

(i) Primary energy sources,

(ii) Secondary energy sources.

i. Primary energy sources

Primary energy sources are sources of energy that are used in the form in which they occur naturally.

Primary energy sources fall into two groups:

(a) Finite energy sources,

(b) Renewable sources.

a. Finite energy sources are those energy sources that last after a number of years when exploited.

Examples are coal, oil, natural gas, and nuclear fuels.

b. Renewable energy sources: these cannot be exhausted. Examples are solar energy, biofuels, hydroelectric power, wind power, wave power, tidal and geothermal power, wind power, wave power, tidal and geothermal power.

ii. Secondary energy sources

Secondary energy sources are used in the non – natural form.

SOLAR ENERGY

Nature of solar energy

The sun's energy is produced by thermonuclear fusion.

Not all of the solar radiation arriving at the edge of the Earth's atmosphere reaches the Earth's surface.

About 30% is reflected back into space by atmospheric dusts and by the polar ice caps.

About 47% is absorbed during the day by the land and sea and becomes internal energy (i.e. heats the Earth). At night this is radiated back into space as infrared.

23% causes evaporation from the oceans and sea to form water vapour. This results into rain and hence **hydroelectric power**.

-0.2% causes convection currents in the air, creating wind power which in turn causes **wave power.**

-0.02% is absorbed by plants during photosynthesis and is stored in them as chemical energy. Plants are sources of **biofuels**

Solar constant

Solar constant is defined as the solar energy falling per second on a square meter placed normal to the sun's rays at the edge of the Earth's atmosphere, when the Earth is at mean distance from the sun.

Its value is about **1.35 kWm^2**

The amount of solar radiation received at any point on the earth's surface depends on:

(i) The geographical location,

(ii) The season, (summer or winter)

(iii) The time of the day, the lower the sun is in the sky the greater is the atmospheric absorption.

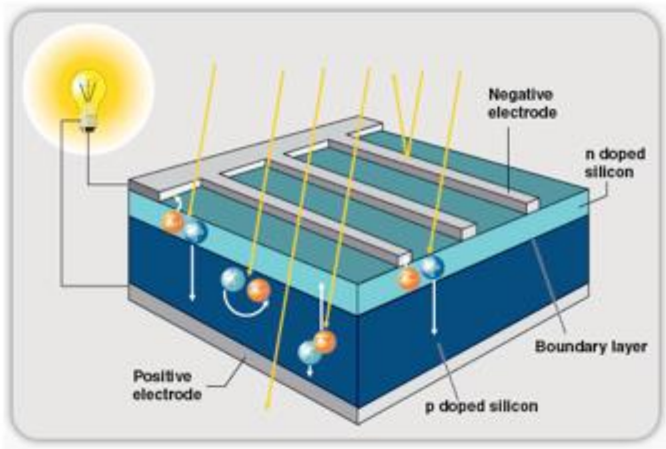
(iv) The altitude; the greater the height above sea level the less is the absorption by the atmosphere, clouds and pollution

PHOTOVOLTAIC DEVICES (SOLAR CELLS)

A solar cell (PV, cells) is a PN junction device which converts solar energy directly into electrical energy.

How it Works

PV cells are made of at least two layers of semiconductor material. One layer has a positive charge (p – type material), the other negative (n-type material). When light enters the cell, some of the photons from the light are absorbed by the semiconductor atoms, freeing electrons from the cell's negative layer to flow through an external circuit and back into the positive layer. This flow of electrons produces electric current.



Uses of the solar cell

1. (i)Are used to power electronics in satellite and space vehicles.
2. (ii)Are used as power supply to some calculators.
3. (iii)Are used to generate electricity for home, office and industrial uses.

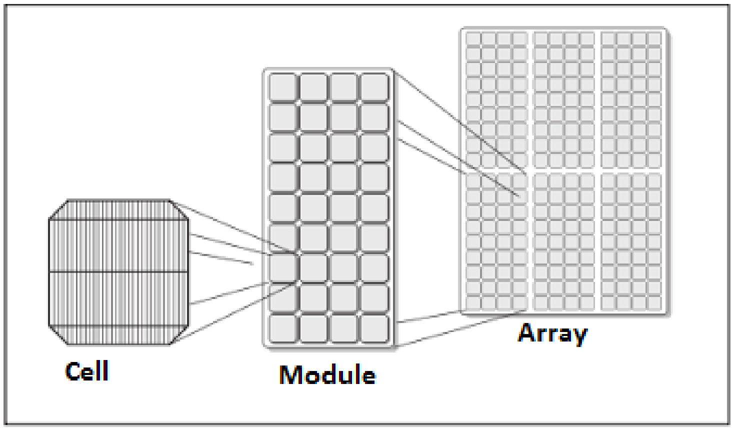
Series arrangement of solar cells

Solar panel (module) is a sealed, weatherproof package containing a number of interconnected solar cells so as to increase utility of a solar cell.

When two modules are wired together in series, their voltage is doubled while the current stays constant.

When two modules are wired in parallel, their current is doubled while the voltage stays constant.

To achieve the desired voltage and current, modules are wired in series and parallel into what is called a PV array.
The flexibility of the modular PV system allows designers to create solar power systems that can meet a wide variety of electrical needs, no matter how large or small.



Efficiency of a photovoltaic system

The output power of a solar cell depends on:

(i) The amount of light energy from the sun falling on a solar panel (the intensity of light).

(ii) The orientation of the solar panel. More electricity is produced if light falls perpendicular to panels.

(iii) The surface area of the panel. Large area collects more solar energy and hence greater electricity.

The best designed solar cell can generate 240 Wm^{-2} in bright sun light at an efficiency of about 24%.

Advantages of photovoltaic systems

1. Solar cells can produce electricity without noise or air pollution.

2. A photovoltaic system requires no fuels to purchase.

3. Panels of photovoltaic cells are used for small – scale electricity generation in remote areas where there is sufficient sun.

4. Net metering: This has the potential to help shave peak loads, which generally coincide with maximum PV power production.

5. The electricity from a PV system is controllable.

Disadvantages of photovoltaic systems

1. They require an inverter to convert the d.c output into a. c for transmission.

2. They produce electricity only when there is sunlight. Hence they need backup batteries to provide energy storage.

3. Suitable in areas which receives enough sunlight.

4. Photovoltaic large scale power generation is cost effective. This is due to large surface area of cells required for generating high power outputs and the need to convert d.c to a.c for transmission.

5. Compared to other energy sources, PV systems are an expensive way to generate electricity.

6. The available solar resource depends on two variables: The latitude at which the array is located and the average cloud cover.

WIND ENERGY

Winds are due to conventional currents in the air caused by uneven heating in the earth's surface by the sun.

Wind energy is extracted by a device called **wind turbine.**

Wind speed increases with the height; it is greatest in hilly areas. It is also greater over the sea and coastal areas where there is less surface drag.

Wind turbines are also called **aerogenerator** or **wind mills** (old name)

Types of wind turbines

There are two types of wind turbines;

(i) Horizontal axis wind turbines (HAWT)

(ii) Vertical axis wind turbines (VAWT)

Horizontal axis wind turbine (HAWT)

HAWT has two or more long vertical blades rotating about a horizontal axis. Modern HAWTs usually feature rotors that resemble aircraft propellers, which operate on similar aerodynamic principles, i.e. the air flow over the airfoil shaped blades creates a lifting force that turns the rotor. The nacelle of a HAWT houses a gearbox and generator (alternator).

Advantage of HAWT

1. HAWTS can be placed on towers to take advantage of higher winds farther from the ground.

Disadvantages of HAWT

1. The alternator (generator) is paced at the top of the supporting tower.

2. Can produce power in a particular wind direction.

Vertical axis wind turbine (VAWT)

In vertical axis, the blades are long and vertical and can accept wind in any direction. The blades are propelled by the drag force on the blades as the wind flows.

Advantages of VAWT

1. It can harness wind from any direction

2. Typically operate closer to the ground, which has the advantage of allowing placement of heavy equipment, like the generator and gearbox, near ground level rather than in the nacelle.

Disadvantages of VAWT

1. Winds are lower near ground level, so for the same wind and capture area, less power will be produced compared to HAWT.

2. Time varying power output due to variation of power during a single rotation of the blade.

3. The need for guy wires to support the tower.

4. Darrieus VAWTS are not self starting like HAWTS. (More colorful picture and videos during lecture)

Power of a Wind Turbine

Consider a wind turbine with blades of length, r (area A), the wind speed is v and the air density is ρ. Assuming that the air speed is reduced to zero by the blades.

Kinetic energy of the wind, K.E = $\frac{1}{2}mv^2$

Kinetic energy per unit volume

K.E per volume = $\frac{1}{2}mv^2$ ÷ volume = $\frac{1}{2}\rho v^2$

The blades sweeps out an area A in one turn, so the volume of air passing in one second is Av.

Kinetic energy per second

= K.E per unit volume x volume per second

K.E per second = $\frac{1}{2}\rho v^2 \times Av$ = $\frac{1}{2}\rho Av^3$

The available wind power is P = $\frac{1}{2}\rho Av^3$

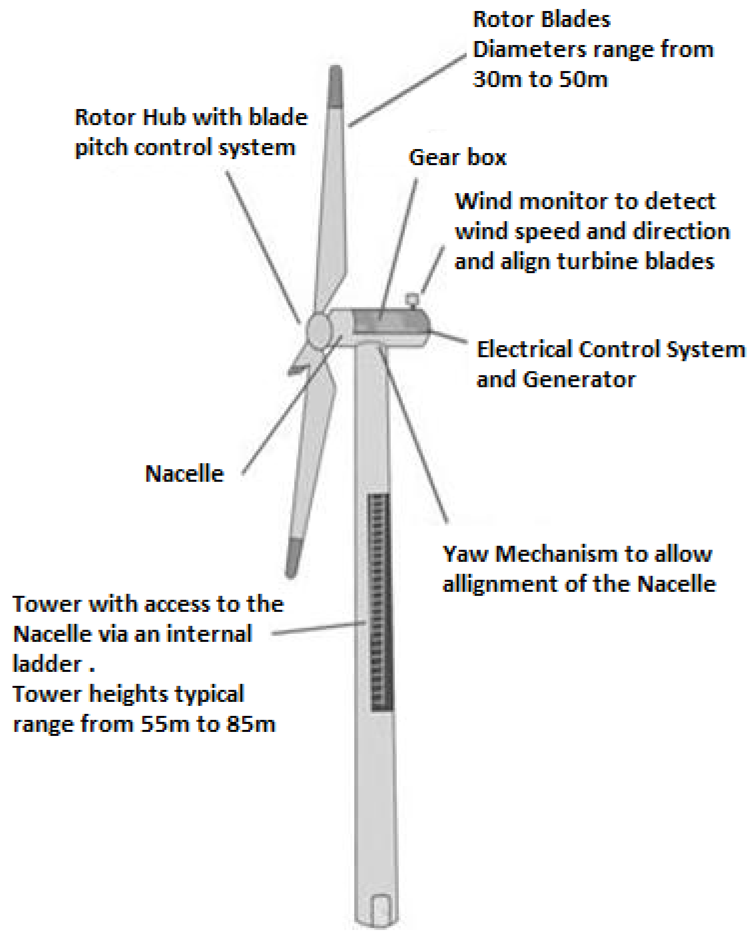


Diagram of a typical wind turbine

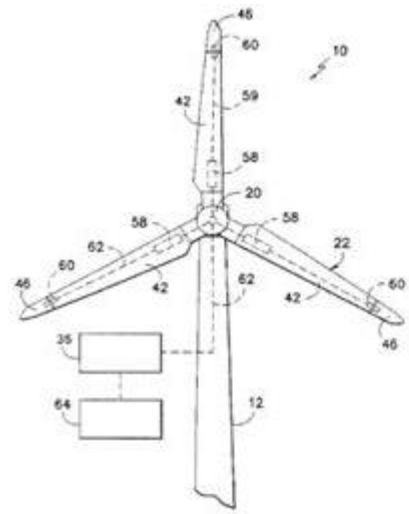


FIG. −11−

Extractable power

The power extracted by the rotating blades is much less than the available wind power. This is because:

(i) The velocity of the wind is not reduced to zero at the blades

(ii) Losses due to friction at the turbine and alternator

(iii) Due to losses in both the gear train and generator.

The power actually captured by the wind turbine rotor, $\mathbf{P_R}$, is some fraction of the available power, defined by the coefficient of performance, **Cp**, which is essentially a type of power conversion efficiency:

$$Cp = \frac{P_R}{P}$$

$$C_{P=}\frac{Power\ capture\ by\ rotor, P_R}{Available\ wind\ power, P}$$

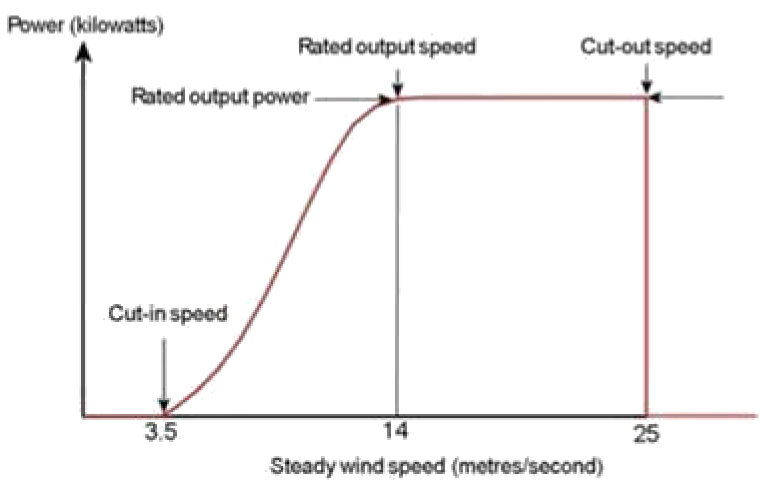
The extractable power (electrical power output) is given by

$$P_R = c_p n_s n_b ({}^1/_2 APV^3)$$

Where n_s and n_b are efficiencies (power output over power input) for the generator and the gearbox.

Variations of power with wind speed

The power curve for a wind turbine shows this net power output as a function of wind speed.



i. Cut in wind speed: This is the lowest speed at which the wind turbine will start generating power.

Typical cut – in wind speeds are 3 to 5 m/s.

ii. Nominal wind speed: This is the lowest speed at which the wind turbine reaches its nominal power output.

Above this speed, higher power outputs are possible, but the rotor is controlled to maintain a constant power to limit loads and stresses on the blades.

iii. Cut – out wind speed: This is the highest wind speed which the turbine will operate at.

Above this speed, the turbine is stopped to prevent damage to the blades.

Advantages of Wind Energy

1. Wind Energy is an inexhaustible source of energy and is virtually a limitless resource.
2. Energy is generated without polluting environment
3. This source of energy has tremendous potential to generate energy on large scale.
4. Like solar energy and hydropower, wind power taps a natural physical resource,
5. Windmill generators don't emit any emissions that can lead to acid rain or greenhouse effect.
6. Wind Energy can be used directly as mechanical energy
7. In remote areas, wind turbines can be used as great resource to generate energy
8. In combination with Solar Energy they can be used to provide reliable as well as steady supply of electricity.
9. Land around wind turbines can be used for other uses, e.g. Farming.

Disadvantages of Wind Energy

1. Wind energy requires expensive storage during peak production time.
2. It is unreliable energy source as winds are uncertain and unpredictable.
3. There is visual and aesthetic impact on region
4. Requires large open areas for setting up wind farms.
5. Noise pollution problem is usually associated with wind mills.

6. Wind energy can be harnessed only in those areas where wind is strong enough and weather is windy for most parts of the year.

7. Usually places, where wind power set-up is situated, are away from the places where demand of electricity is there. Transmission from such places increases cost of electricity.

8. The average efficiency of wind turbine is very less as compared to fossil fuel power plants. We might require many wind turbines to produce similar impact.

9. It can be a threat to wildlife. Birds do get killed or injured when they fly into turbines.

10. Maintenance cost of wind turbines is high as they have mechanical parts which undergo wear and tear over the time.

NB: Even though there are advantages of wind energy, the limitations make it extremely difficult for it to be harnessed and prove to be a setback

GEOTHERMAL ENERGY

Geothermal energy is the energy from nuclear energy changes deep in the earth, which produces hot dry rock.

Geothermal energy originates from the heat retained within the Earth since the original formation of the planet, from radioactive decay of minerals, and from solar energy absorbed at the surface.

Harnessing Geothermal Energy

Most high temperature geothermal heat is harvested in regions close to tectonic plate boundaries where volcanic activity rises close to the surface of the Earth. In these areas, ground and groundwater can be found with temperatures higher than the target temperature of the application.

Geothermal energy is extracted by using two methods:

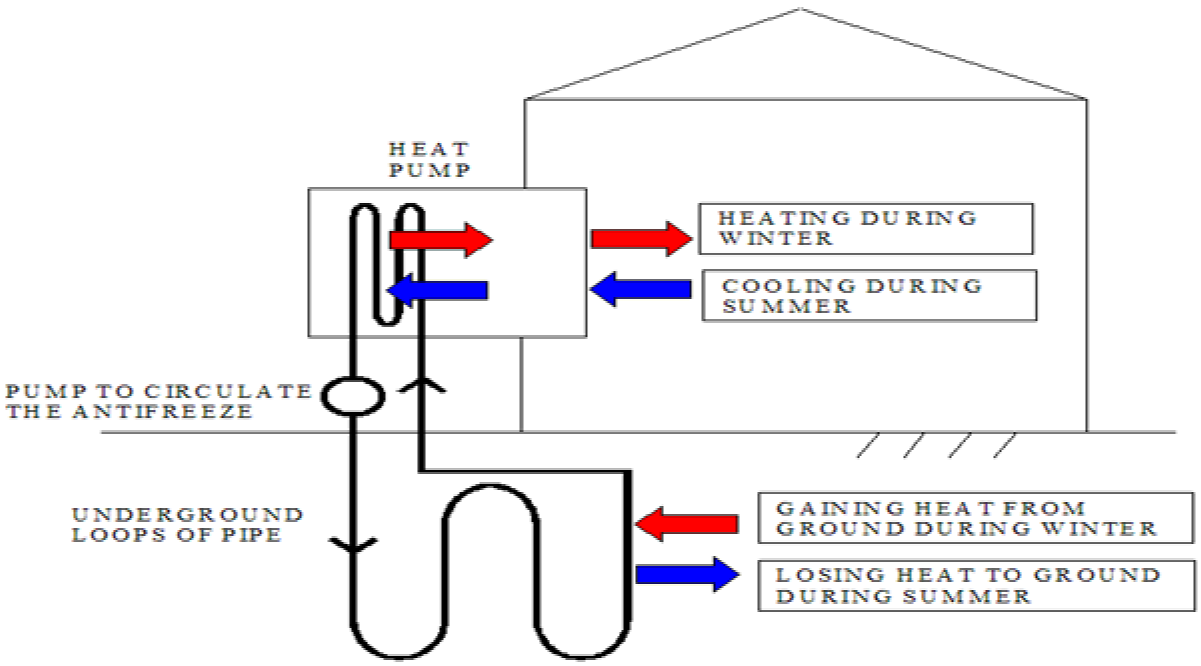
(i) A heat pump system

(ii) Hot dry rock conversion

The heat pump system

Hot aquifers are layers of permeable (porous) rock such as sandstone or limestone at a depth of 2 – 3 km which contains hot water at temperatures of 60 – 100^{0}C.

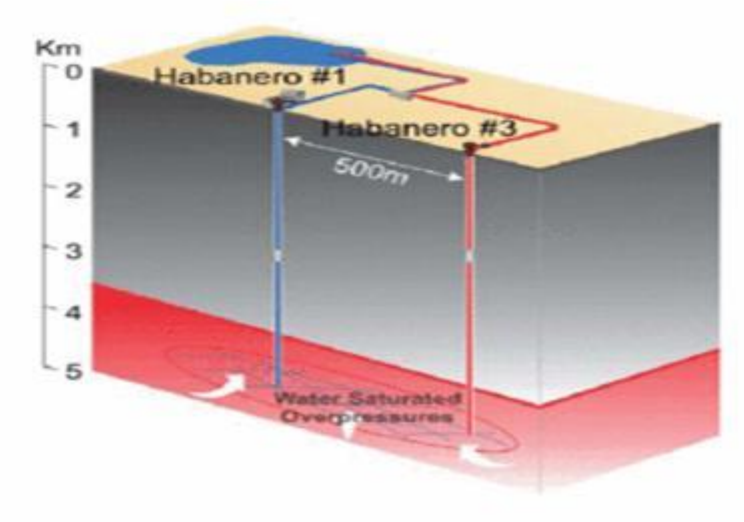
A shaft is drilled to aquifer and the hot water pumped up it to the surface where it is used for district space and water heating schemes or to generate electricity. A second shaft may be drilled to return the cool water to the rock.



The hot dry rock conversion

These are impermeable hot dry rocks found at depth of 5 – 6 km, have temperature of 200^{0}C or more.

Two shafts are drilled and terminate at different levels in the hot rock about 300 m apart. The rocks near the end are fractured by explosion or by methods to reduce the resistance
to the flow of cold water which is pumped under very high pressure (300 atm) down the injection shaft and emerges as steam from the top of the shallower shaft.



Uses of geothermal energy

Geothermal energy can be used for electricity production, for direct use purposes, and for home heating efficiency (through geothermal heat pumps).

Advantages of geothermal energy

1. Geothermal power plants provide steady and predictable base load power.
2. New geothermal power plants currently generate electricity at low cost.
3. Responsibly managed geothermal resources can deliver energy and provide power for decades.
4. Geothermal power plants are reliable, capable of operating about 98 percent of the time.
5. Power plants are small, require no fuel purchase and are compatible with agricultural land uses.
6. Geothermal plants produce a small amount of pollutant emissions compared to traditional fossil fuel power plants.

Disadvantages of geothermal energy

1. Many of the best potential resources are located in remote or rural areas, often of federal or state lands
2. Although costs have decreased in recent years, exploration and drilling for power production remain expensive

3. Using the best geothermal resources for electricity production may require an expansion or upgrade of the transmission system.

4. The productivity of geothermal wells may decline over time. As a result, it is crucial that

developers manage the geothermal resources efficiently.

WAVE ENERGY

Wave energy is the energy extracted from the ocean surface wave. Energy that comes from the waves in the ocean sounds like a boundless, harmless supply.

Machinery able to exploit wave power is generally known as a **wave energy converter** (WEC)

Wave power

Waves in the sea have kinetic energy and gravitational potential energy as the rise and fall.

Consider a sine wave of wave length λ spread over a width d the amplitude of the wave is a and the time period is T.

The power in a wave come from the change in potential energy of the water as it rotates on the circuit paths beneath the surface. It can be shown that the power carried forward by a wave is given by:

$$\left\{Power = \left(\frac{pa^2g^2T}{8\pi}\right)d\right\}$$

Wave Energy Flux

The mean transport rate of the wave energy through a vertical plane of unit width , parallel to a wave crest, is called **wave energy flux**.

From above,

$$\left\{Power = \left(\frac{pa^2g^2T}{8\pi}\right)d\right\}$$

$$\left\{Wave\ energy\ flux = \left(\frac{pa^2g^2T}{8\pi}\right)\right\}$$

Harvesting wave energy

There are two type of system:

1. i. Offshore systems in deep water more than 141 feet deep. **The Salter duck method**.

    (a) Pumps that use bobbing motion of waves.

    (b) Hoses connected to floats on surface of waves. As float rises and falls , the hose stretches and relaxes, pressurizing the water which then rotates a turbine

2. ii. Onshore systems are built along shorelines and harvest energy from braking waves.

    (a)Oscillating water columns are of concrete or steel and have an opening to the sea below the waterline. It uses the water to pressurize an air column that is drawn through the turbine as waves recede.

    (b)A Tapchan is a tapered water system in sea cliffs that forces waves through narrow channels and the water that spills over the walls is fed through a turbine.

    (c)A Pendulor device is a rectangular box with a hinged flap over one side that is open to the sea .Waves cause the flap to swing back and forth and this powers a hydraulic pump and generator.

Advantages of Wave energy

1. Renewable: It will never run out.

2. Environment friendly: Creating power from waves creates no harmful byproducts such as gas, waste, and pollution.

3. Abundant and widely available: Another benefit to using this energy is its nearest to places that can use it.

4. Variety of ways to harness: Current gathering method range from installed power plant with hydro turbine to seafaring vessels equipped with massive structures that are laid into the sea to gather the wave energy.

5. Easily predictable: The biggest advantage of wave power as against most of the other alternative energy source is that it is easily predictable and can be used to calculate the amount that it can produce.

6. Less dependency on foreign oil cost.

7. Non damage to land.

Disadvantages of wave energy

1. Suitable to certain locations: The biggest disadvantage to getting your energy from the wave is location. Only power plants and town near the ocean will benefit direct from it.

2. Effect on marine ecosystem: Large machine have to be put near and in the water gather energy from waves .These machines disturb the seafloor, changes the habitat of near-shore creatures (like crabs and starfish) and create noise that disturb the sea life around them.

3. Wavelength: Wave power is highly dependent on wavelength i.e. wave speed, wave length, and wavelength and water density.

4. Weak performance in Rough Weather: The performance of wave power drops significantly during rough weather.

5. Noise and Visual pollution: Wave energy generators may be unpleasant for some who live close to coastal regions. They look like large machines working in the middle of the ocean and destroy the beauty of the ocean. They also generate noise pollution but the noise is often covered by the noise of waves which is much more than that of wave generators.

6. Difficult to convert wave motion into electricity efficiently.

7. Difficult to design equipment that can withstand storm damage and saltwater corrosion.

8. Total cost of electricity is not competitive with other energy sources.

9. Pollution from hydraulic fluids and oils from electrical components.

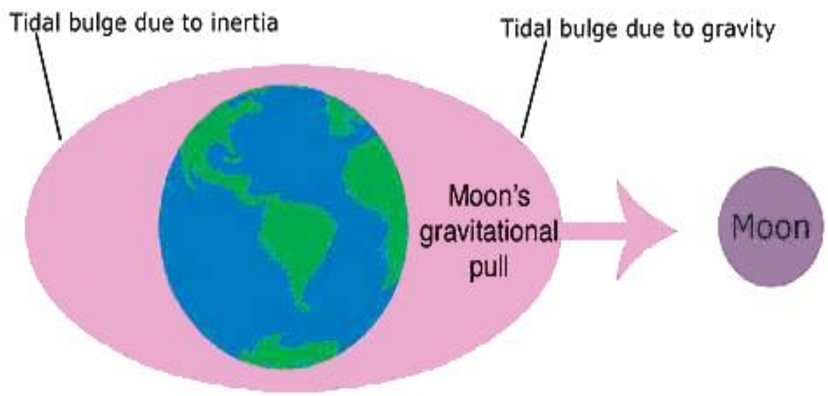
TIDAL ENERGY

Tidal Power is the power of electricity generation achieved by capturing the energy contained in moving water mass due to tides.

Two types of tidal energy can be extracted: **Kinetic energy** of currents between ebbing and surging tides and **potential energy** from the difference in height between high and low tides.

Causes of Tides

Tides are caused by the gravitational pull of the moon, and to a lesser extent the sun, on the oceans. There is a high tide places near the moon and also opposite on the far side.



i. High (spring) tide: Occurs when there is full moon. The moon, sun and earth are in line the moon being between earth and sun. The pulls of the moon and sun reinforce to have extra high tides.

ii. Lowest (neap) tide: Occurs when there is half moon and the sun and moon pulls are at right angles to each other.

iii. Harnessing Tidal Energy

Tidal energy can be harnessed by building a barrage (barrier), containing water turbines and sluice gates, across the mouth of river. Large gates are opened during the incoming (flood) tide, allowing the water to pass until high tides, when they are closed.

On the outgoing tide, when a sufficient head of water has built up, small gates are opened, letting the potential energy of the trapped water drive the turbines and generate electricity.

Advantages of Tidal Energy

1. Decrease reliance on coal driven electricity so less CO_2 emissions.

2. Changing technology allowing quicker construction of turbines, which in turn increases likelihood of investment with a shorter return.

3. Once constructed very little cost to run and maintain.

4. Tidal energy is renewable and sustainable.

Disadvantages of Tidal Energy

1. Intermittent energy production based around tides creates unreliable energy source.

2. High construction costs

3. Barrages can disrupt natural migratory routes for marine animals.

4. Barrages can disrupt normal boating pathways.

5. Turbines can kill up to 15% of fish in area, although technology has advanced to the point that the turbines are moving slow enough not to kill as many.

Tidal Power

If the tidal height (level) is h and the estuary area is A, then the mass of water trapped being the barrier is ρAh and the centre of gravity is h/2 above the low tide level.

The maximum energy per tide is therefore $= mgh$

Potential Energy of tide = $\rho Ah \times g \times \frac{h}{2} = \frac{1}{2}\rho g A h^2$

Averaged over a tidal period of *T* (approx. 12 hours a day), this gives a mean power available of.

Average tidal power = $\frac{\rho g A h^2}{2T}$

Note that the efficiency of the turbines (generator) will determine how much of this tidal power will be harnessed.

EXAMPLES: SET B

Example 01

The power output p of a windmill can be expressed as $P = kA\rho v^3$ where A is the area swept out by the windmill blades (sails), ρ is the density of air, *v* is the wind speed and *k* is a dimensionless constant

(a) Show that the units on both sides of this expression are the same

(b) Sketch a graph to show how the power increases with wind speed as v rises from zero to 15ms^{-1}

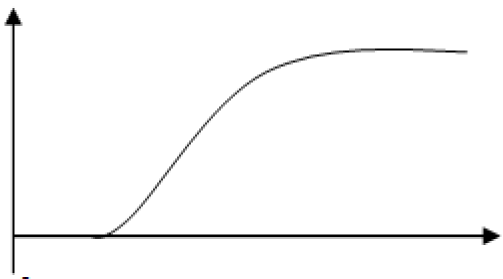
Solution

(a) Units on L.H.S = Nms^{-1}

Unit on R.H.S. = m^2 (kgm^{-3}) x (ms^{-1})^3

= (kgms^{-2}) ms^{-1}=Nms^{-1}

(b) Variation of power with speed



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Example 02

The radiation received from the sun at the earth's surface in a certain country is about 600 Wm^{-2} averaged over 8 hours in the absence of cloud.

(a) What area of solar panel would be needed to replace a power station of 2.0 GW output, if the solar panels used could convert solar radiation to electrical energy at an efficiency of 20%

(b) What percentage is this area of the total of the country (which is about 3 x $10^{11} m^2$)?

(c) If the total power station capacity is about 140 GW, what percentage of the surface of the country would be covered by solar panels if all the power stations were replaced?

Solution

(a) Output of a solar panel

$$P_{out} = 2x10^9 = \frac{20}{100} \times 600 \times A$$

$$\therefore A = 1.67 \times 10^7 m^2$$

(b) Percentage area to the country

$$= \frac{1.67 \times 10^7}{3.0 \times 10^{11}} \times 100\% = 0.0056\%$$

(c) Area of solar panels required

$$140\times10^{9} = \frac{20}{100}\times600\times A$$

$$\therefore A = 1.167\times10^{9}\,m^{2}$$

Percentage area to the country

$$= \frac{1.167\times10^{9}}{3.0\times10^{11}}\times100\% = 0.39\%$$

Example 03

(a) What are aerogenerators?

(b) Estimate the maximum power available from $10m^2$ of solar panels and calculate the volume of water per second which must pass through if the inlet and outlet temperatures are 20^0C and 70^0C. Assume the water carries away energy at the same rate as the maximum power available. The specific heat capacity of water is 4200 Jkg^{-1} and solar constant is 1.4 kWm^{-2}.

Solution

(a) Aerogenerators are devices that convert the kinetic energy of wind into electrical energy. E.g. windmill.

(b) Maximum power available from solar panel

$$P_{out} = \text{solar constant} \times \text{area of panel} = 14\text{kW}$$

$$\therefore A = 1.67\times10^{7}\text{m}^{2}$$

Volume of water per second used is given by

$$P_{out} = \frac{m}{t}\times c\Delta\theta = {}^{V}/_{t}\times(\rho c\Delta\theta)$$

$$1400 = {}^{V}/_{t}\times(1000\times4200\times(70-20))$$

$$\therefore {}^{V}/_{t} = 6.67\times10^{-5}m^{2}$$

Example 04

A coal – fired power station has an output of 100mW. Given that its efficiency is 45%, how much coal must be supplied each day? Assume 1 tonne of coal gives 3×10^{10} of energy.

Solution

Input power of the station is given by

$$\text{efficiency} = \frac{\text{output power}}{\text{input pwer}} \times 100\%$$

$$45\% = \frac{100 \times 10^{6}}{\text{input power}} \times 100\%$$

$$\text{input power} = 222.222 \times 10^{6} W$$

Total input energy in a day is

$$\text{input energy} = 222.222 \times 10^{6} \times 24 \times 3600$$
$$= 1.92 \times 10^{13}$$

The amount of coal required is

$$= \frac{1.92 \times 10^{13}}{3 \times 10^{10}} = 640 \textit{ tonnes}$$

Example 05

Calculate the energy required transport1000 tones of oil along a 100km pipeline; given that 0.05 kW hours of energy is used to shift each tone of oil along each km of pipeline. Given that 1 tonne of oil releases 4.2×10^{10} J if burned, what percentage of the total energy available from 1000 tonnes of oil is used to shift the oil along the pipeline? (Ans: 18GJ, 0.043%)

Example 06

A hydroelectric power station has efficiency of 25%. The water driving the turbines falls through a height of 300m before reaching the turbines. Calculate the volume of water that must pass through the turbines each second to give a power output of 2mW. Assume the density of water is $1000kg^{-3}$.

Solution

Power of the falling water

$$25\% = \frac{2 \times 10^6}{P} \times 100\%$$

$$\therefore P = 8.0 \times 10^6 J$$

But,

$$P = \left(\frac{m}{t}\right) gh = {V}/{t} \times \rho gh$$

$$\Rightarrow 8.0 \times 10^6 = {V}/{t} \times 10^3 \times 10 \times 300$$

$$\therefore {V}/{t} = 2.67 m^3 s^{-1}$$

Example 07

The solar energy flux near the Earth is 1.4W m^{-2}. A solar power station consists of concave mirrors that focus sunlight onto a steam boiler. What must be the minimum mirror area to given an output 1 mW, assuming 100% efficiency? Why in practice, should the mirror area be greater?

Solution

Minimum mirror area is given by

$$A = \frac{1 \times 10^6}{1.4 \times 10^3} = 714 m^2$$

The mirror area should be greater to achieve such a power output because part of the incident energy is absorbed by the mirror.

Example 08

A solar panel attached to the roof of a house is used to heat water from 5^0C to 40^0C. If the water flows through the panel at a rate of $0.012kgs^{-1}$ Calculate the heat gained per second by the water. Assume the specific heat capacity of water is $4200Jkg^{-1}K^{-1}$. **(Ans. 1764 Q)**

Example 09

An aerogenerator has a power output that is proportional to (wind speed) ^{2} and its efficiency varies with wind speed. On a day when there is a steady wind of speed 9 ms^{-1}, the power output is 40kW operating at an efficiency of 20%. If the wind speed on next day is 13.5 ms^{-1} and the efficiency increases to 25% what is the new power output?

Solution

Power output α efficiency × (wind speed)^2

$$\frac{P_1}{P_2} = \left(\frac{e_1}{e_2}\right)\left(\frac{V_1}{V_2}\right)^2$$

$$\Rightarrow P_2 = 40\times\left(\frac{25}{20}\right)\times\left(\frac{13.5}{9}\right)^2 = 112.5kW$$

Example 10

Estimate the energy released from a tidal power station if 100 km^3 of water raised to height of 1.5m by the tide behind a tidal barrier. What would be the mean power output of such a station if its efficiency is 25% and there are two tides per day?

Solution

The tidal power is given by

$$\text{Energy} = mgh = PVgh = 10^3 \times 100 \times 10^9 \times 10 \times \frac{1.5}{2}$$

$$\therefore \text{Energy} = 7.5 \times 10^{14}J$$

Note that the centre of gravity of water mass is at the half height up.

$$\text{Average tidal power} = \frac{\rho g A h^2}{2t} = \frac{\rho g V h}{2t}$$

$$= \frac{10^3 \times 10 \times 100 \times 10^9 \times 1.5}{2 \times 12 \times 3600} = 1.736 \times 10^{10}\text{W}$$

Mean power output is

$$= \frac{25}{100} \text{x}\, 1.736 \text{x}\, 10^{10} = 4.34 \text{x}\, 10^9 \text{W}$$

Example 11

An open boat of width 1.0 m has a total weight of 3000N.Used near a beach, it bobs up and down through 0.5 m once every 5s. Calculate the losses of P.E. every time it drops from a crest to a through. Hence estimate the mean power available per meter of beach waves.

Solution

Loss in P.E. is given by

$$= \text{mgh} = \text{Fh} = 3000\text{x}\frac{0.5}{5} = 300\text{W}$$

The mean power available per meter is 300 W

Example 12

(a) If energy is conserved, why is there energy crisis?

(b) Explain the terms high grade and low grade energy and give examples.

(c) Draw an energy flow diagram for a hydroelectric power station. Why does such a station have a much greater efficiency than a thermal power station?

Refer Advanced Physics by Tom Duncan fifth edition for more problems on energy.

GEOPHYSICS

Geophysics is the branch of physics which deals with the study of seismic waves and the Earth's magnetic and gravity fields and heat flow.

Because we cannot directly observe the Earth's interior, geophysical methods allow us to investigate the interior of the Earth by making measurements at the surface. Without studying these things, we would know nothing of the Earth's internal structure.

STRUCTURE OF THE EARTH

Major zones of the earth

The earth is divided into two major zones, namely;

(a) Outer zone, and

(b) Inner zone.

a) Outer zone: the earth's outer zone consists of;

(i) The hydrosphere – water bodies,

(ii) The atmosphere – gaseous envelope

(iii) The biosphere – living organisms, plant and animals

b) Inner zone: the earth's inner zone consists of;

(i) The crust – lithosphere

(ii) The mantle – mesosphere,

(iii) The core – barysphere

Atmosphere is the envelope of gases that surround the Earth (oxygen, nitrogen, carbon dioxide, etc)

Hydrosphere is the water bodies filling the depressions in the Earth. Examples are rivers, oceans, seas, oasis, *etc.*

Lithosphere is the solid outer most part of the earth.

EARTH'S LAYERS

Layers defined by composition

Layers are defined by composition because of density sorting during an early period of partial melting, Earth's interiors not homogeneous.

- **Crust** – the comparatively thin outer skin that ranges from 3 kilometers at the oceanic ridges to 70 kilometers in some mountain belts. It makes up 1% of the Earth's volume.

Continental crust (SIAL, Silicon and aluminium)

Average rock density about 2.7 g/cm^3

Its density varies between 2.0 to 2.8 g/cm^3

Composed of silicon and aluminium

Floats higher on the mantle forming the land masses and mountains. It is 30 to 70 km thick.

Oceanic crust (SIMA), silicon and magnesium)

Oceanic crust ranges from 3 to 15 km thick

Density vary between 3.0 to 3.1 g/cm^3

Floats lower on the mantle forming the oceanic basins. It is about 8 km thick.

□□□□ **Mantle** – a solid rocky (silica-rich) shell that extends to a depth of about 2900 kilometers. It makes up 83% of the Earth's volume

The mantle can further be dived into:

(i) Upper layer of mantle (Asthenosphere)

(ii) Transition layer and,

(iii) Lower layer of mantle (Mesosphere)

Upper mantle is a rigid layer of rock with average density 3.3kgm^{-3}

Transition layer is the layer that separates upper and lower mantle.

Lower mantle plays an important role in tectonic plate movement which creates earthquakes and volcanoes.

Its density is about 5.7 kgm^{-3}

Note:

The mantle rocks are said to be in a plastic state.

The upper part of a mantle has a temperature of about 870^0C. The temperature increases downwards through the mantle to about 2200^0C near the core.

- **Core** - an iron – rich sphere having a radius of 3486 kilometers making up 16% of the Earth's volume

The core is divided into two parts:

(i) Outer core

(ii) Inner core

i. Outer core is a liquid of molten iron and nickel alloys. The Earth's magnetic field is generated within the outer core due to convective. It is 2270 kilometers thick.

ii. Inner core is a solid iron and nickel alloys. The temperature within the inner core is higher than the outer core but the inner core is solid, this is because higher pressure in this region causes the melting point to rise. It is a sphere of radius of 1216 kilometers.

Average density is nearly 11 $gcm^{-3}$and at Earth's center.

Layers defined by physical properties

Lithosphere (sphere of rock)

Earth's outermost layer

Consists of the crust and uppermost mantle

Relatively cook, rigid shell

Averages about 100 kilometers in thickness, but may be 250 kilometers or more thick beneath the older portions of the continents

Asthenosphere (weak sphere partially molten)

Beneath the lithosphere, in the upper mantle to a depth of about 660 kilometers

Small amount of melting in the upper portion mechanically detaches the lithosphere from the layer below allowing the lithosphere to move independently of the asthenosphere i.e. allows tectonic plate movement.

Mesosphere or lower mantle

Rigid layer between the depths of 660 kilometers and 2900 kilometers

Earth's major boundaries

Discontinuity is the name given to any surface that separates one layer from another layer of the Earth.

The Moho (Mohorovicic discontinuity)

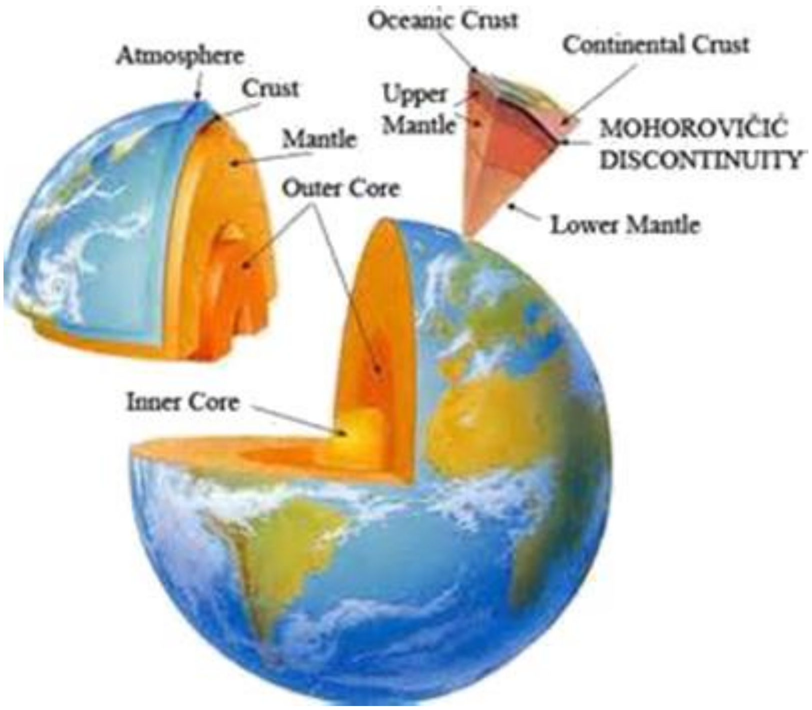
Discovered in 1909 by Andriaja Mohorovicic

Separates crustal materials (crust) from underlying mantle.

Gutenberg discontinuity

- Discovered in 1914 by Beno Gutenberg
- Is the boundary between the outer and inner core.

The Earth's Structure



TEMPERATURE INSIDE THE EARTH

Earth's temperature gradually increases with an increase in depth at a rate known as the geothermal gradient.

Temperature varies considerably from place to place

Averages between about 20^0C and 30^0C per kilometer in the crust (rate of increase is much less in the mantle and core)

The rate of heat flow within the Earth depends on:

(i) The thermal conductivity of the rock,

(ii) Temperature gradient of the rock

Sources of heat Energy within the Interior of the Earth

Major processes that have contributed to Earth's internal heat include:

1. Heat emitted by radioactive decay of isotopes of uranium (U), thorium (Th), and potassium (K).
2. Heat released as iron crystallized to form the solid inner core.
3. Heat released by colliding particles during the formation of Earth.
4. Gravitational work done by the Earth due to its rotation through its own axis.
5. Electron motion in the core behaves like an electric current.

Heat Lost by the Earth

Heat in the earth is transferred by the process of;

(i) Convection and

(ii) Conduction

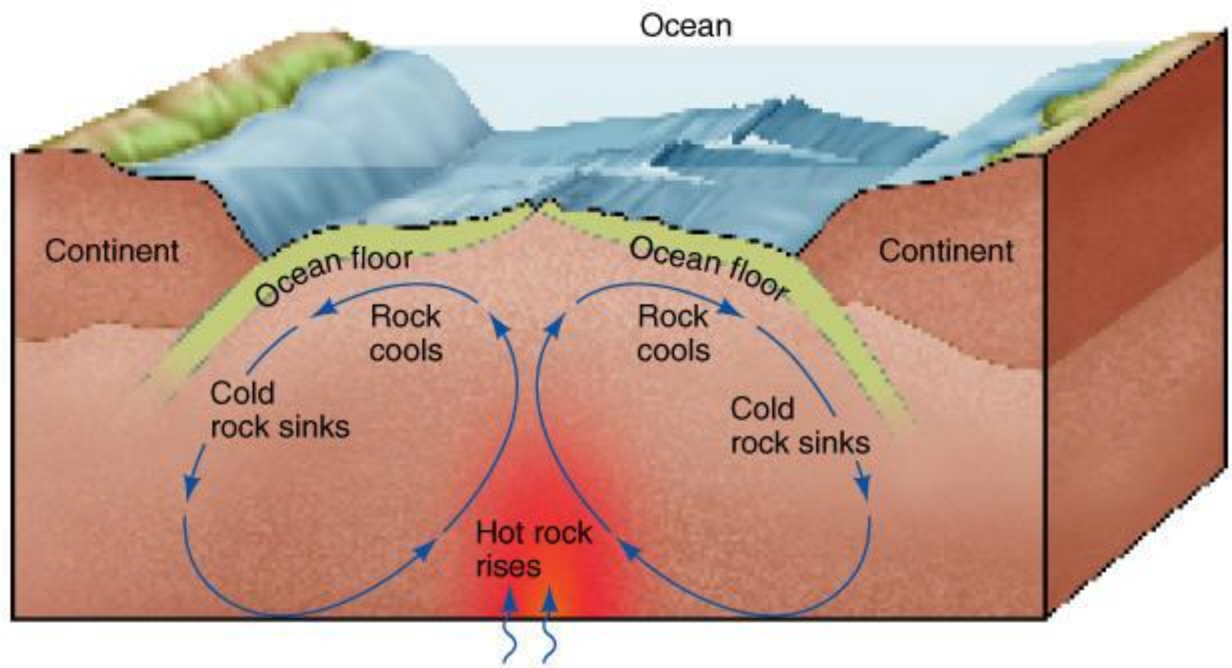
In the solid inner core and in the Earth's crust heat is transmitted by conduction process. Rates of heat flow in the crust vary.

In the Mantle heat is transmitted by conduction process. Rates of heat flow in the crust vary.

In the Mantle heat is transmitted by convection process. There is not a large change in temperature with depth in the mantle.

Mantle must have an effective method of transmitting heat from the core outward.

Transfer of heat in the Earth by mantle convection



Uses of the Mantle

1. The mantle transfers heat by convection from the earth's crust to the out regions of the earth and thus help it to regulate its temperature

2. The upper part of the mantle is molten, this allows tectonic plates movements.

EARTHQUAKES

An earthquake is a sudden motion or shaking of the earth caused by a sudden release of energy that has accumulated within or along edges of the earth's tectonic plates.

Earthquakes occur within the Earth's crust along faults that suddenly release large amounts of energy that have built up over long periods of time.

The shaking during an **earthquake** is caused by seismic waves.

Seismic waves are propagating vibrations that carry energy from the source of the shaking (earthquake) outward in all directions.

Seismic waves are generated when rock within the crust breaks, producing a tremendous amount of energy. The energy released moves out in all directions as waves, much like ripples radiating outward when you drop a pebble in a pond.

CAUSES OF EARTHQUAKES (SEISMIC WAVES)

The main causes of the Earthquakes and so seismic waves are:

1. Movement of tectonic plate.
2. Volcanic activity.
3. Landslide and avalanches.
4. Rebound of the crust.
5. Human activities.

Movement of tectonic plate

The Earth's crust is made up of segment (layers) called tectonic plates which are slowly drifting in various directions. Tectonic plates may create a fault.

A boundary is a line where two tectonic plates meet.

A geologic fault is a fracture in the earth's crust causing loss of cohesion and accompanied by displacement along the fracture.

How an earthquake is formed

Tectonic plates grind past each other, rather than slide past each other smoothly. As the plates move past each other they can become locked together due to friction. For some time, they don't move and strain energy builds up. Stresses builds between them until fractional force holding the plates together give away. The plates move suddenly, releasing the energy and then held again. This sudden jerk is what is felt as an earthquake.

Note

(a) The Earth's crusts near tectonic plate edges are forced to bend, compress, and stretch due to the internal forces within the earth, causing earthquakes.

(b) Nearly all earthquakes occur at plate boundaries.

Volcanic activity

Molten rock “magma” from the mantle is forced through a weak point in the Earth’s crust creating a volcanic eruption. When magma reaches the Earth’s surface it is known as **“Lava”.** Successive eruptions leads to the buildup of lava on the sides of the vent creating the familiar “cone – shape” of a volcanoes

Earthquakes may be created by the violent explosions which occur if there are sudden movements of the magma.

Landslides and avalanches

A landslide occurs when a large mass of land slips down a slope. An Avalanche occurs when a large mass of snow pours down a mountain side. Both of these effects can start an earthquake

Rebound of the crust

Elastic rebound theory state that "as tectonic plates move relative to each other, elastic strain energy builds up along their edges in the rocks along fault planes". Since fault planes are not usually very smooth, great amount of energy can be stored (if the rock is strong enough) as movement is restricted due to interlock along the fault. When the shearing stresses induced in the rocks on the fault planes exceed the shear strength of the rock, rupture occurs.

It follows from this that if rocks along the fault are of a certain strength, the fault is a certain length, and the plates are slipping past each other at a defined rate, it is possible to calculate the amount of time it will take to build up enough elastic strain energy to cause an earthquake and its probable magnitude.

When a fault breaks it release elastic strain energy it stored, and hence earthquake.

Human activities

Human activities such as those caused by nuclear bombs can create earthquake, together with mine actives.

<u>EARTHQUAKE TERMS</u>

Energy released by an earthquake moves outwards from the origin in the form of concentric waves.

Focus (Hypocenter) is the point in the Earth where seismic waves originate. **Epicenter** is the point on the earth's surface vertically above the focus.

Hypocentral distance is the distance between the focus and the seismic detection station.

Epicentral distance is the distance between the epicentral and the seismic station.

S = Seismic station

E = Epicenter

ES = Epicentral distance

TYPE OF SEISMIC WAVES

i. Seismic waves are elastic waves that propagate within the earth.

There are two type of seismic waves:

1. **ii. Body waves,** spread outward from the focus in all directions.
2. **iii. Surface waves (Long, L – waves)** spread outward from the epicenter to the Earth's surface along the crust, similar to ripples on a pond. These waves can move rock particles in a rolling motion that very few structures can withstand. These waves move slower than body waves.

BODY WAVES

There are two types of Body Waves

(1) Primary P – wave and

(2) Secondary, S – waves

1. **1. Primary Wave (P** – wave): Are longitudinal (compression) wave (travels in the same direction the waves move)

Characteristics of P – waves

1. Are the fastest seismic waves (7 – 14 km/second). Arrives at recording station first, hence the name primary means first.
2. Can pass through solid, gas and liquid, hence can pass through crust, mantle and the cores.

3. Are longitudinal compression waves. The rocks that transmit the P – waves are alternately compressed and expanded.

Velocity of P – waves

The velocity of primary waves depends on the density, ρ- bulk modulus B and the shear modulus η-

In solid, = $V_p = \sqrt{\frac{B + \frac{4}{3}\eta}{\rho}}$

In liquid = $V_p = \sqrt{\frac{B}{\rho}}$

A fluid cannot support shear stresses hence $\eta = 0$

2. **Secondary Wave (S – wave):** Are transverse (shear) wave (travels perpendicular to the wave movement).

Characteristics of S – waves

1. i. Slower moving (3.5 – 7 km/second) hence are detected after primary waves.
2. ii. Caused by a shearing motion
3. iii. Cannot pass through a fluid (gas or liquid) because they are transverse. Hence are unable to pass through the liquid outer core.

Velocity of S – waves

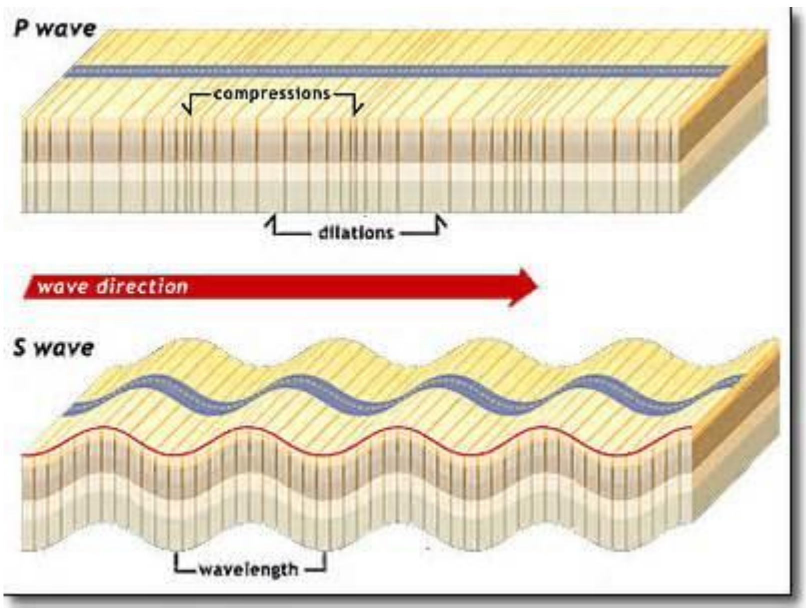
The velocity of shear waves depends on the density ρ, and the shear modulus η

In solid, = $V_s = \sqrt{\frac{\eta}{\rho}}$

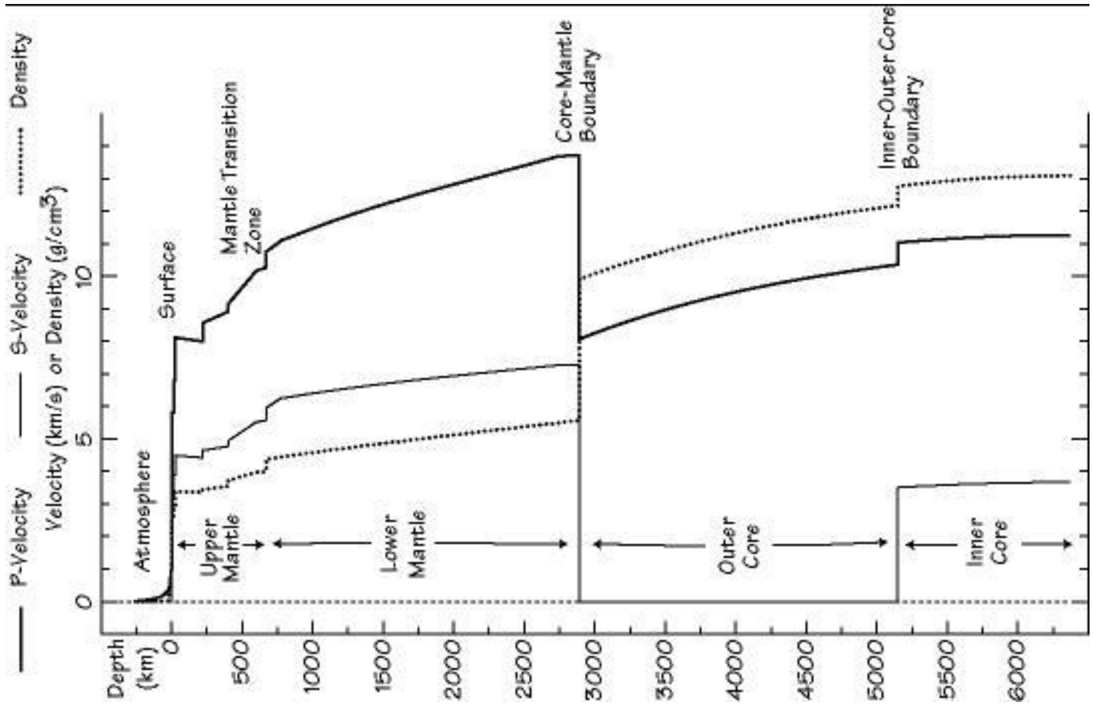
In liquid = $V_p = 0$, since $\eta = 0$

Note: Since the density and states of the earth layers varies, the speed of the seismic waves also vary from layer to layer, the solid part showing greater speed and the liquid ones lower speed.

Primary wave and secondary wave



Variation of speed of body waves with depth



SURFACE WAVES/LONG WAVES

Surfaces waves are produced when earthquake energy reaches the Earth's surface.

These are the slowest moving waves, but are the most destructive for structures on earth

There are two types of L – Waves:

(i) Love long waves

(ii) Rayleigh long waves

i. Love Waves

Love waves are Transverse horizontal motion, perpendicular to the direction of propagation and generally parallel to the Earth's surface.

They are formed by the interaction of S waves with Earth's surface and shallow structure and are dispersive waves. The speed at which a dispersive wave travels depends on the wave's period.

Characteristics of Love Waves

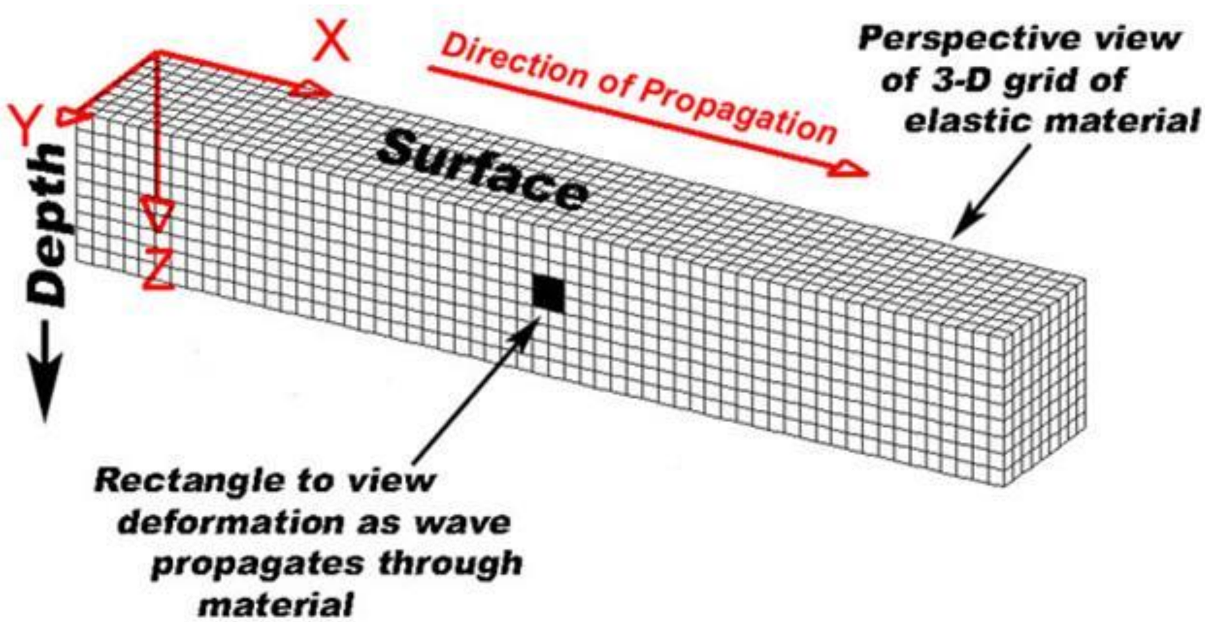
1. i. Love waves are transverse and restricted to horizontal movement (horizontally polarized).

2. ii. The amplitude of ground vibration caused by a Love wave decrease with depth. The rate of amplitude decrease with depth also depends on the period/frequency.

3. iii. Loves wave are dispersive, i.e. wave velocity is dependent on frequency; low frequency – higher velocity.

4. iv. Speed of love waves is between 2.0 and 4.4 km/s

5. v. Love waves travels within the earth's crust only.



LOVE WAVE

Rayleigh Waves

Rayleigh waves are vertically polarized long waves. The slowest of all the seismic wave types and in some ways the most complicated.

Characteristics of Rayleigh Waves

1. Rayleigh waves are transverse and restricted to vertical movements (vertically polarized).

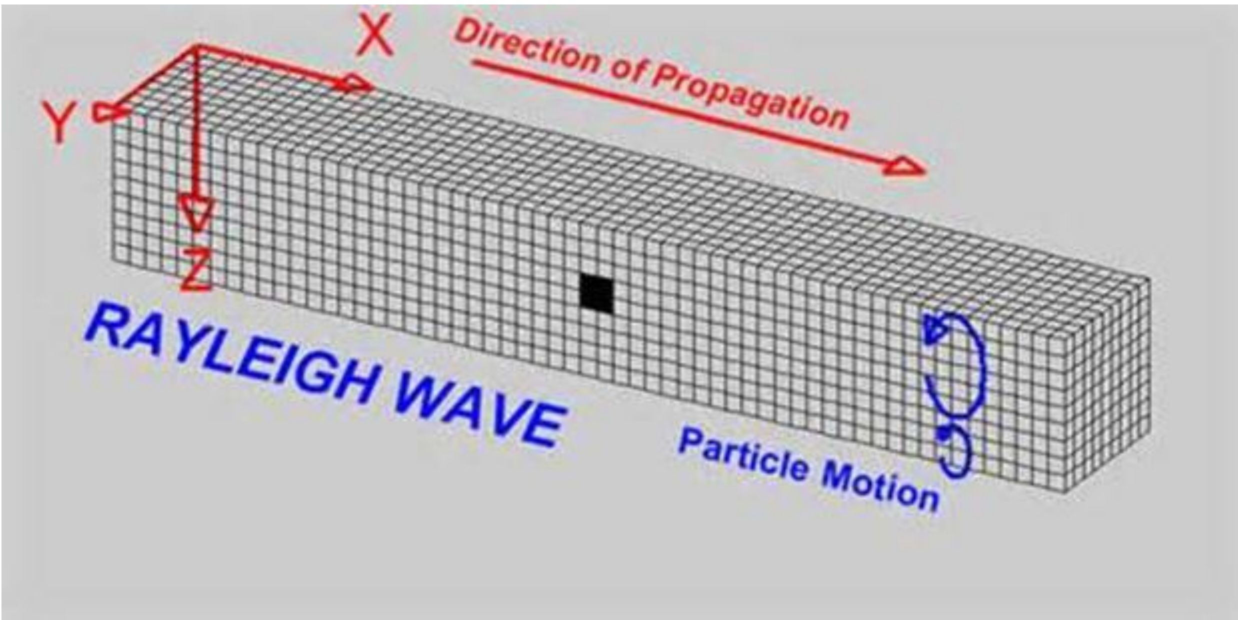
2. The amplitude of Rayleigh wave decreases with depth. The rate of amplitude decrease with depth depends on the period/frequency

3. Rayleigh wave are dispersive, i.e. wave velocity dependent on frequency; low frequency – high velocity

4. Speed of love waves is between 1.0 and 4.2 km/s slowest of all waves.

5. Travels within the earth's crust only.

6. Depth of penetration of the Rayleigh waves depend frequency, with lower frequencies, penetrating greater depth.



PROPAGATION OF SEISMIC WAVES

Like all other types of waves, seismic waves may undergo,

(i) Reflection, (ii) Refraction, (iii) Dispersion, (iv) Diffraction, (v) Attenuation.

Seismic reflection: Seismic waves bounce (reflect) rock boundaries of different rock type (density).

Seismic refraction:
Waves change velocity and direct (refract) when they enter a medium of different density it the one they just passed through.

Seismic Dispersion:
surface waves are dispersive which means that different periods travel at different velocities. The effects of dispersion become more noticeable with increasing distance because the long travel distance spreads the energy out (it disperses to energy).

SEISMIC WAVE PATHS

By comparing the data recorded by many stations all over the world the nature, speed and the paths of the seismic waves can be determined. This information can be used to tell us about the earth's interior such as density sand state in each layer.

L – Waves travel within the Earth's crust only

P and S waves travel through the earth in a curve path. The waves are refracted because their speeds a constantly changing with depth due to continue increase in density. Waves are also strongly refracted the Mantle – Core boundary.

Surface waves travels through the Earth crust only

Shadow zone is the region on the Earth's surface where no S or P waves are present.

This lies between 105^0 and 140^0. Only surface waves may be detected in this region.

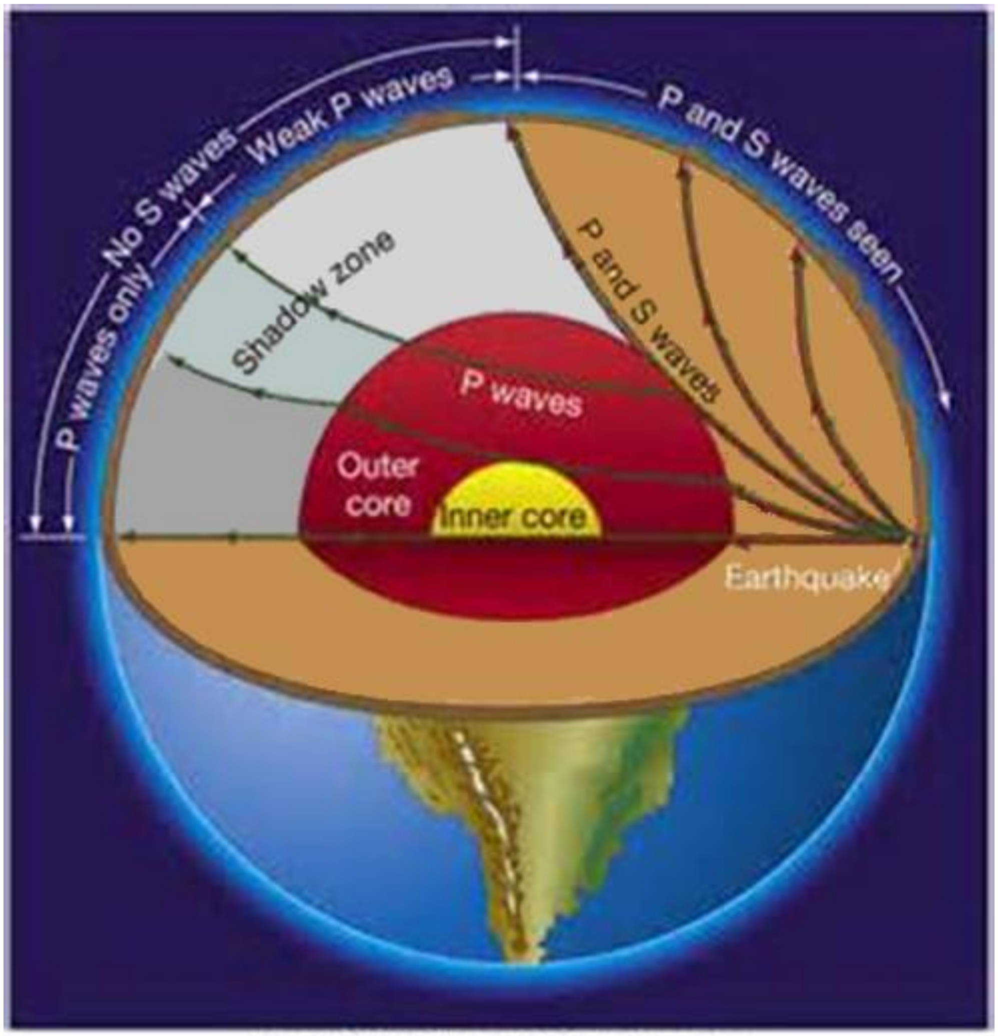
Shadow zone occurs because:

(i) P – Waves are strongly refracted at the liquid outer core.

(ii) S – Waves can't travel through the liquid outer core.

Seismic waves can also be used to locate the discontinuities in the earth's crust. A change in density or crack would affect the propagation of the waves.

This alteration in the wave's path or speed would indicate the discontinuity.



The fact that S waves do not travel through the core provides evidence for the existence of a liquid layer beneath the rocky mantle.

The change in the velocity of P waves at crust – Mantle boundary reveals the presence of Mohorovicic discontinuity

P waves passing through the inner core show increased velocity suggesting that the inner core is solid.

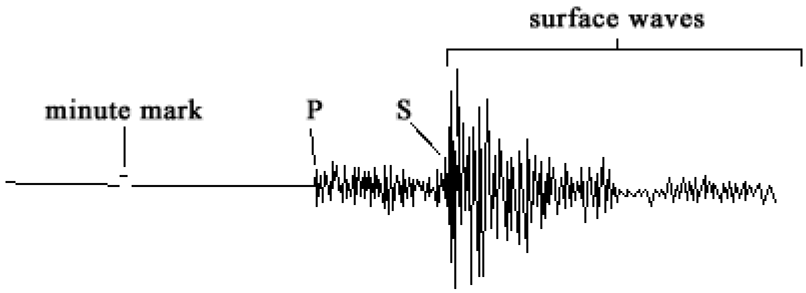
Both P and S – Waves slow down when they reach the **asthenosphere**. Because of this scientists know that the **asthenosphere** is partially liquid

MEASUREMENTS OF EARTHQUAKES

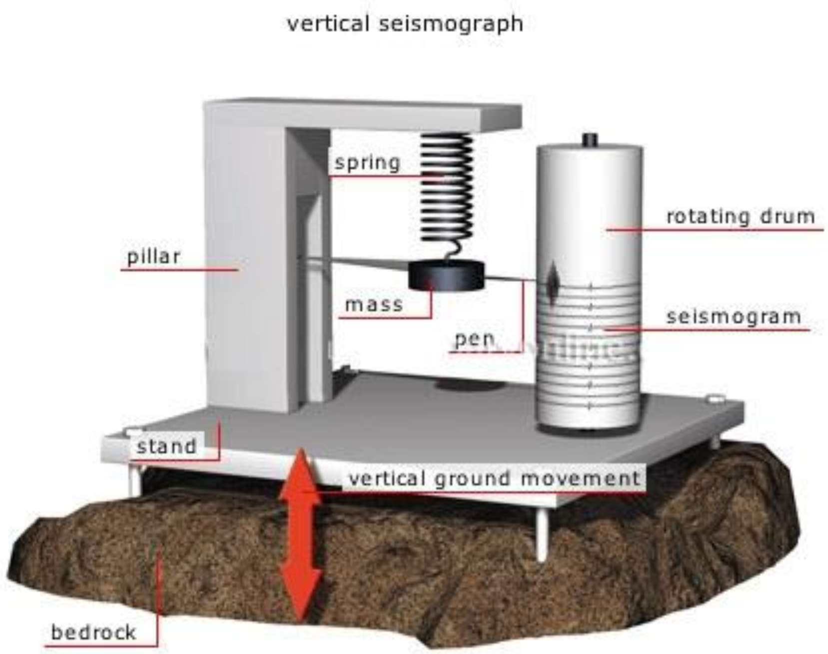
i. Seismology is the scientific study of earthquakes (seismic waves) and artificially produced vibrations in the earth.**Seismograph** is a sensitive instrument that is used to record earthquakes and seismic waves (i.e. ground movements).

ii. Seismogram is the record of ground movement drawn by a seismograph.

The arrival of seismic waves at a station



Seismograph consists of a heavy weight suspended from a frame fixed into the ground. When the earth vibrates the frame moves but the heavy weight remains stationary due to great inertia. A pen attached to weight plots the earth's movements on a chart recorder to produce a seismogram. To obtain a complete record of the earthquake measurements must be taken in all three planes (x, y and z).



The recording of the motion caused by seismic waves can be done by using;

(a) Mechanical method, as in the drawing above.

((b) Optical method, where light is used to write the motion on a photosensitive paper instead of using a pen.

(c) Electronic method, where a coil is fixed to the mass of the pendulum and moves in a magnetic field. This induces a voltage which is amplified so that they can be easily interpreted.

Seismometers record both the magnitude and intensity of the earthquake.

LOCATING THE EPICENTRE

Although S – waves, P – waves and surface waves all start out at the same time, they travel at different speeds. The speed of a traveling seismic wave can be used to determine the location of an earthquake epicenter.

A seismograph records the arrival time and the magnitude of horizontal and vertical movements caused by an earthquake. The arrival time between different seismic waves is used to calculate the travel time and the distance from the epicenter.

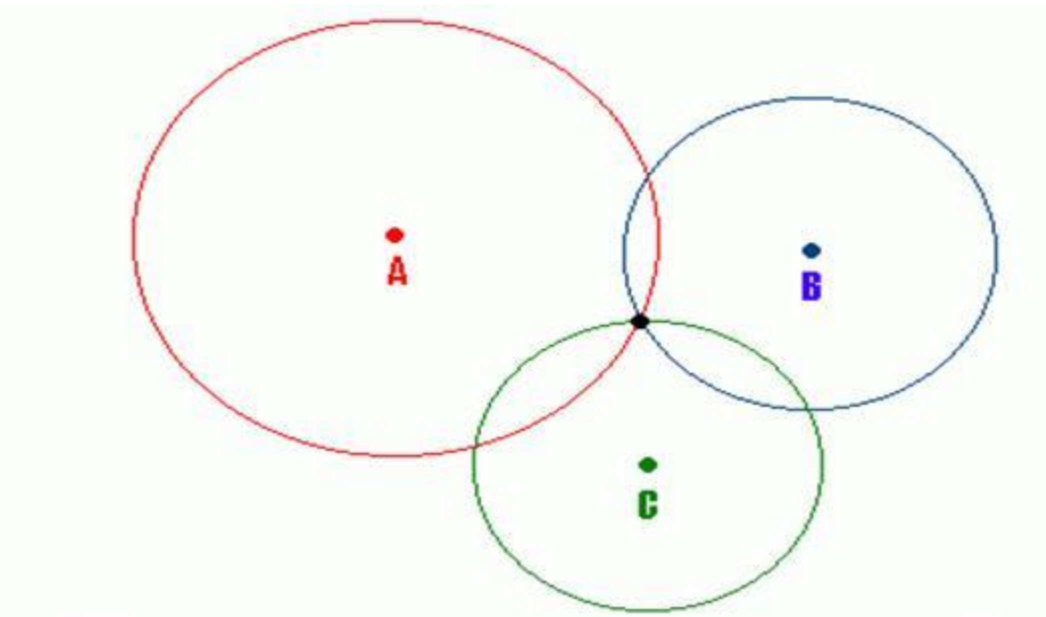
The difference in arrival time between primary waves and secondary waves is used to calculate the distance from the seismograph station to the epicenter.

It is crucial that seismic waves are recorded by three different seismograph stations in order to estimate the location of the epicenter.

(i) Locate at least 3 stations on a map that recorded the seismic waves.

(ii) Calculate the time difference between arrival of P – waves and arrival of S – waves from a seismogram. The time difference is proportional to the distance from the epicenter. Because the direction to the epicenter is unknown, the distance defines a circle around the receiving station. The radius of each circle equals that station's distance from the earthquake epicenter.

(iii) The epicenter is where the circles intersect.



Earthquake epicenter requires at least 3 recording stations to localize.

Radius = distance from epicenter

SIZE OF AN EARTHQUAKE

The size of an earthquake can be measured in terms of its intensity (Mercalli/Wood Neumann scale) or its magnitude (Richter scale).

Mercalli Intensity Scale

The Mercalli scale measures the intensity of how people and structures are affected by the seismic event. In essence, it measures damage. It is much more subjective and uses numbers ranging from 1 (no damage) to 12 (total destruction).

Degree	Explana
1	Detecte seismog
6	Felt by frighten heavy moved, fallen general small
12	Total large waves moving the objects thrown

ISOSEISMAL LINES

Intensity distribution maps can be drawn up showing the intensities of an earthquake over a region. The earthquake is most intense at the epicenter and decreases with distance.

Isoseismal lines are line joining points of equal intensity.

Richter magnitude scale

The magnitude of an earthquake is measured in terms of energy released by an earthquake. This is determined from the amplitude of the seismic wave recorded on a seismogram 100 km from the epicenter. The magnitude is equal to the logarithm of the amplitude. Therefore each successive number represents a tenfold (x10) increase in the ground motion. The Richter scale starts at 0 but has no upper limit. -However 8 represent an earthquake that causes total destruction within the region.

Magnitude	Amount of explosives (TNT) needed to release the equivalent energy, in tons
6	6,000
7	180,000
8	5.4 million

Intensity of an earthquake is a measure of its strength based on the changes it causes to the landscape.

EARTHQUAKE PREDICTIONS (WARNINGS)

Forecasting (predicting) earthquakes is very difficult, although there are a number of warning signs which occur before an earthquake happens.

(i) Change in the velocity of p – waves.

(ii) Electrical resistivity of the rocks decreases.

(iii) An increase in radon, emission (radon is an inert gas, radon is found to increase in soil and water samples).

(iv) Increase in fore shock (small tumors that occur just before an earthquake).

(v) Local variations in the magnetic field.

(vi) Animals begin to behave strange.

(vii) Water levels rise or fall in wells few days before earthquake.

(viii) Increase in temperature of the area few months before the occurrence of an earthquake

PRECAUTIONS

Some of the world's populations are living in regions where there is a high risk of an earthquake. Most of these regions lie along fault lines. However a few precautions can be taken to reduce the damage caused.

(a)Build structures that can withstand the forces of an earthquake. One method is to include shock absorbers into the buildings foundations.

(b)Scientific research has shown that pumping water out of the earth reduces the stress in the crust hence preventing an earthquake. However this technique is very expensive.

(c) Stay away from tall buildings or structures during an earthquake if you are outside on occurrence.

(d) If you are inside a house, stay in a safe place where things will not fall on you.

EARTHQUAKE HAZARDS

Earthquake give rises to a number of hazards which pose a great risk to human life, animals, property and the environment at large. The following are some hazards:

1. **Landslides and avalanches:** The shaking caused by an earthquake can cause unstable hillsides, mountain slops' and cliffs to move downwards creating landslides. Earthquakes can also trigger avalanches on snow slopes

2. **Tsunamis:** If an earthquake occurs under the sea or ocean, the shock waves disturb the water. The ocean floor can rise or fall causing the water to rise and fall too. This movement creates huge water waves called tsunamis that travel across the ocean.

3. **Collapsing building:** Buildings or structures may collapse during a strong earthquake. The collapse of the building may kill people.

4. **Fire outbreak:** Earthquakes can cause gas or oil pipes to break and or the collapse of electricity lines. This may set up fire.

5. **Backward rivers:** Tilting ground due to earthquakes can make rivers change their course.

REFLECTION SEISMOLOGY

This is the study of reflection of seismographic waves by different materials inside the earth.

Applications:

(i) Location of underground oil and water

(ii) Locate discontinuities within the earth

SEISMIC PROSPECTING

Seismic prospecting is the sending of seismic waves into the deep earth' crust in order to study the structure of the earth or detecting oils or gases in the interior of the earth by utilizing the property of reflection and refraction of the seismic waves.

THE EARTH'S MAGNETIC FIELD

The earth has a weak magnetic field, 95% of this field is created inside the Earth's core 5% is the result of atmospheric effects above the Earth's surface.

Geomagnetism is science of study of the earth magnetic field, its causes and its variations.

Generation of the Earth's magnetic field within the core

The accepted explanation for the origin of the Earth's magnetic field within the core is given by Lemoir's self exciting dynamo theory.

The Earth's Outer Core consists of molten conducting metals (Iron and Nickel) which are rich in free electrons. The Earth's rotation causes the molten metal to rotate and hence large convection currents are set up within the outer core. These currents generate a magnetic field.

Eddy currents are now generated due to a conducting material moving in a magnetic field. These Eddy currents modify the position of the Earth's magnetic field so that it does not lie along the Earth's axis of rotation. The present magnetic poles are situated 800km from the Earth's axis.

Generation of magnetic field in the Atmosphere

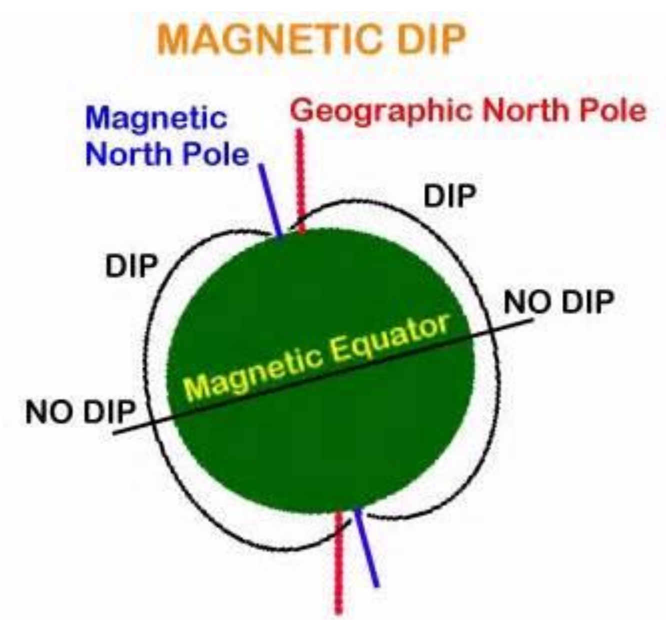
In the Earth's atmosphere there is a region know as the ionosphere which consists of free electrons and ions. The movement of these charges creates a magnetic field. This effect provides a small fraction of the Earth's total magnetic field.

TERMS ASSOCIATED WITH THE EARTH'S MAGNETIC FIELD

Magnetic meridian: A vertical plane passing through the axis of a freely suspended magnetic needle.

Geographic meridian: A vertical plane passing through the geographic axis.

Magnetic equator: Is the locus of points on earth's surface where the needle (free to rotate in a vertical plane) remains horizontal.



The Earth's magnetic field pattern is similar to that produced by a giant bar magnet or solenoid.

Note: (i) The magnetic North pole which lies in the Northern Hemisphere behaves like a south pole or a bar magnet, i.e. the field lines are directed towards it.

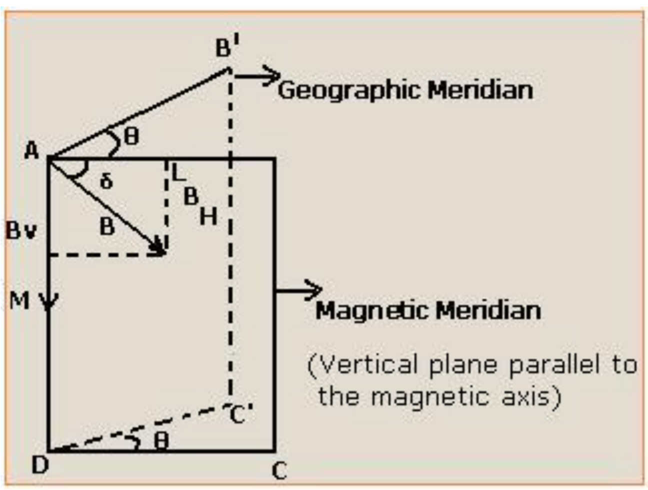
(ii) The magnetic south pole which lies in the southern hemisphere behaves like a north pole of a bar magnet, i.e. the field lines are directed away from it.

ELEMENTS OF EARTH'S MAGNETISM

Angle of variation of declination, θ at a place is the angle between the geographic meridian and the magnetic meridian at that place.

Angle of dip or declination, α at a place is the angle between the directions of intensity of the earth's total magnetic field declination B_E and the horizontal direction, B_H in the magnetic meridian at that place.

Horizontal component of Earth magnetic field B_H It is the component of the Earth's total magnetic field along the horizontal direction in the magnetic meridian.



By Pythagoras theorem

$$B_R = \sqrt{B_v^2 + B_H^2}$$

By trigonometric ratio

$$\left\{\tan\alpha = \frac{B_V}{B_H}\right\}$$

Points to note about angle of Dip

(a) At the poles, $\alpha = 90^0 = \frac{B_V}{B_H}$

$B_H = 0 \text{ and } B_V = B_E$

Therefore, only horizontal component exists at the poles

(b) At the equator $\alpha = 0^0$

$$\tan 0^0 = \frac{B_V}{B_H}$$

$$B_V = 0 \text{ and } B_H = B_E$$

At the equator only horizontal component exist.

VARIATIONS OF THE EARTH'S MAGNETIC FIELD

The Earth's magnetic field is not constant but varies continuously with time.

(i) **Short term variations (Irregular changes):** The magnetic field changes daily due to variations in the magnetic field created in the ionosphere. The charged particles in this region of the atmosphere are affected by the Sun's gravitational pull (which is stronger when the sun is directly above that area)

Also during periods of high solar activity charged particles from the **solar wind** are able to penetrate the magneto pause and arrange themselves under the influence of the magnetic field in a formation called **Van Allen Belts.**
These charged particles cause further Eddy currents within the ionosphere, altering the Earth's magnetic field strength.

Solar wind is a continuous stream of moving electrons and protons in the atmosphere which are produced from flare (eruptions) from the sun. Normally these charged particles move from west to south at 300 – 500 km/s.

Magnetic storm is a sudden worldwide disturbance of the earth's magnetic field caused by dynamic interaction of the earth's magnetic field and the sun. During magnetic storm, the earth's magnetic field is unusually active.

Effects of Magnetic Storm

(a) Large storms can cause the loss of radio communication

(b) Damage satellite electronics and affect satellite operations.

(c) Increase pipeline corrosion

(d) Induce voltage surges in electric power grids causing blackouts.

(e) Reduce the accuracy of global positioning systems.

(ii) **Long term variations (Secular changes):** The Earth's magnetic field position is constantly changing, now the magnetic North pole is moving at 8 km per year, and the magnetic South Pole at 16 km per year.

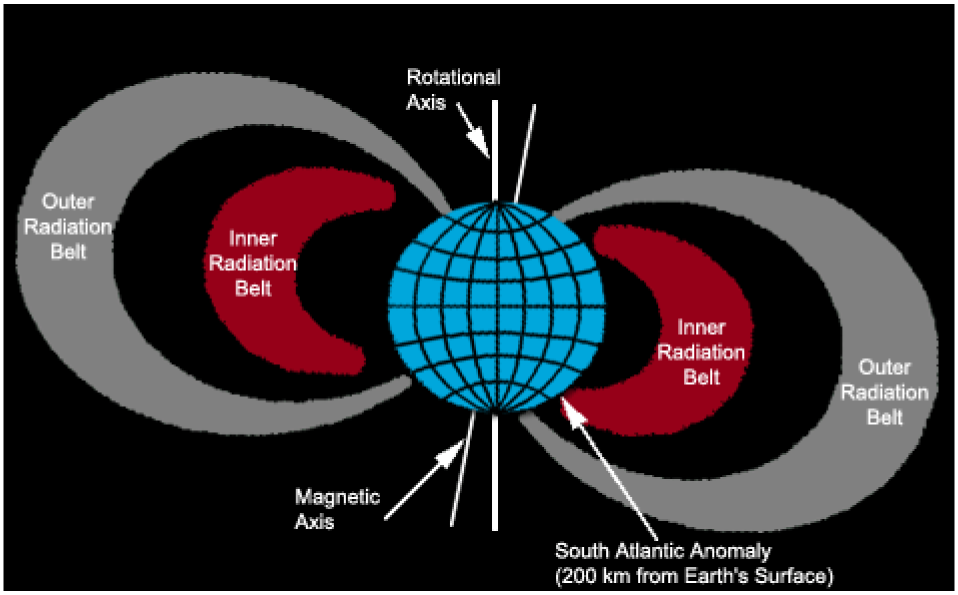
Evidence from the alignment of magnetized rocks layers in the Earth's crust show that the Earth's magnetic field has actually reversed in direction several times during the Earth's history (i.e. the direction of the fields have reversed causing a north acting pole to become a south acting pole.) The present polarity of the Earth's magnetic field has not changed for 700,000 years.

VAN ALLEN BELTS

The Van Allen belts consist of two regions of highly charged particles which are trapped within the Earth's magnetic field:

Inner Belt consists of protons and positive charged particles

Outer Belt consists of electrons and negatively charged particles.



THE ATMOSPHERE

Earth's atmosphere is divided into five main layers, the exosphere, the thermosphere, the mesosphere, the stratosphere and the troposphere. The atmosphere thins out in each higher layer until the gases dissipate in space. There is no distinct boundary between the atmosphere and space, but an imaginary line about 110 kilometers from the surface, called the **Karman line**, is usually where scientists say atmosphere meets outer space.

TROPOSPHERE

The troposphere is the layer closest to Earth's surface. It is 10 km thick and contains half of Earth's atmosphere. Air is warmer near the ground and gets colder higher up. Nearly all of the water vapor and dust in the atmosphere are in this layer and that is why clouds are found here.

Lapse rate is the rate of fall of temperature in degrees per kilometer rise. It has an average value of 6 ^{0}C per km in the troposphere.

Tropopause is the upper boundary of the troposphere.

Importance (uses) of troposphere

1. Controls the climate and ultimately determines the quality of life in the atmosphere.

2. It supports life on earth. It contains oxygen which is used to respiration by animals.

STRATOSPHERE

The stratosphere is the second layer. It starts above the troposphere and ends about 50 km above ground.

The temperature of the stratosphere slowly increases with altitude. This temperature increase is due to the presence of Ozone layer which absorbs heat from the sun in the form of ultraviolet light.

The Ozone layer occupies the middle of stratosphere between 20 and 30 km it consists of Ozone formed by oxygen molecules dissociated and reforming into 0_3.

The air here is very dry, and it is about a thousand times thinner here than it is at sea level. Because of that, this is where jet aircraft and weather balloons fly.

Stratopause is the upper boundary of the stratosphere.

Importance (uses) of stratosphere

The stratosphere prevents harmful ultraviolet radiation from reaching the earth. Ozone absorbs harmful radiation from the sun. The Ozone protects plants and shield people from skin cancer and eye cataracts.

MESOSPHERE

The **mesosphere** starts at 50 km and extends to 80 km high. The top of the mesosphere, called the **mesopause**, is the coldest part of the Earth's atmosphere with temperatures averaging about - 90^0C. The temperature of the mesosphere decreases with altitude (because there is no ozone to absorb heat).

This layer is hard to study. Jets and balloons don't go high enough, and satellites and space shuttles orbit too high. Scientists do know that meteors burn up in this layer.

Importance of mesosphere

Mesosphere, thermosphere and exosphere prevent harmful radiation such as cosmic rays from reaching the earth surface.

THERMOSPHERE

The thermosphere extends from about 80 km to between 500 and 1,000 km. Temperatures increases as it approaches nearer to the sun. The heating effects of the earth no longer exist at these higher altitudes.

The thermosphere is considered part of Earth's atmosphere (the upper atmosphere), but air density is so low that most of this layer is what is normally thought of as outer space. In fact, this is where the space shuttles flew and where the International Space Station orbits Earth.

This is also the layer where the **auroras** occur. Charged particles from space collide with atoms and molecules in the thermosphere, exciting them into higher states of energy. The atoms shed this excess energy by emitting photons of light, which we see as the colorful **Aurora Borealis and Aurora Australis**.

EXOSPHERE

The exosphere, the highest layer, is extremely thin and is where the atmosphere merges into outer space. It is composed of very widely dispersed particles of hydrogen and helium.

The upper part of the exosphere is called **Magnetosphere**. The motion of ions in this region is strongly constrained by the presence of the earth's magnetic field. This is the region where satellites orbit the earth

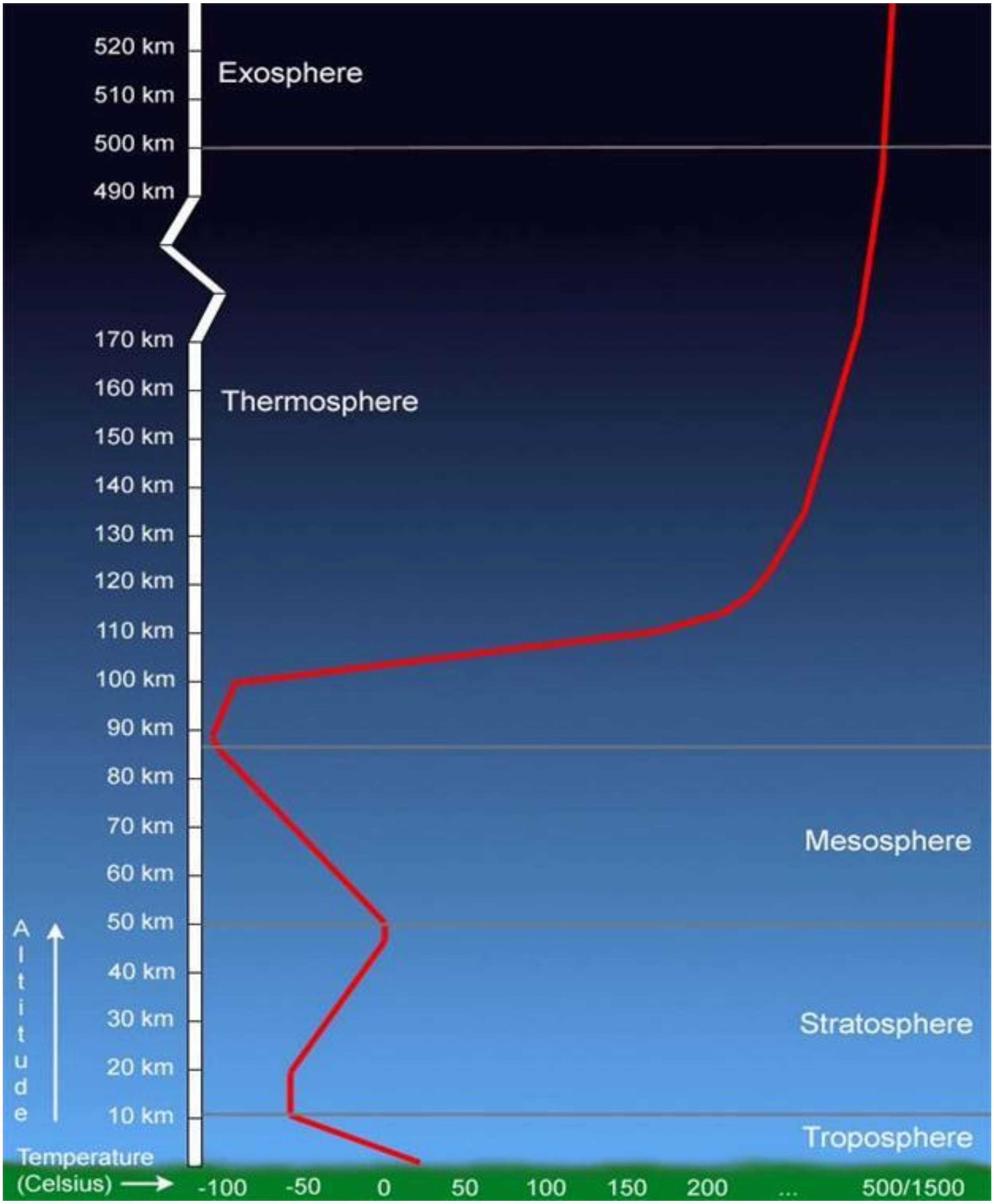
Note:

(i)The troposphere, stratosphere, and mesosphere are collectively forms the **homosphere**. These layers have the same chemical composition; 78% nitrogen, 21% oxygen, 1% argon and other gasses which sum to about 0.05%. The thermosphere is excluded due to different in chemical composition.

(ii) The upper atmosphere above 90 km is called **heterosphere**. The atmosphere is no longer a mixture of gases but separates into layers heavier ones forming the bottom layer.

VARIATION OF TEMPERATURE WITH HEIGHT

The temperature above the Earth surface varies as shown in the graph below.



The residence time, τ, is the mean lifetime of a gas molecule in the atmosphere

THE IONOSPHERE AND TRANSMISSION OF RADIO WAVES

The ionosphere is the region containing high concentrations of charged particles ions and electrons.

The ionosphere is created by atoms absorbing U.V radiation, gamma and X – rays.

The ionosphere extends from the lower thermosphere 55 km to 550 km above the earth's surface.

Ionosphere layers:

Due to difference in composition of the air in the ionosphere, the ionosphere is divided into layers.

(i) **The lower layer, called D layer**; this layer exists only in the day time at an altitude of 55 to 90 km above the earth's surface. Ionization in this region is relatively weak.

(ii) **The next layer, E – layer**: this layer is between 90 and 145 km above the earth's surface. It has a maximum density at noon but is only weakly ionized at night.

(iii) **The top layer, the F – layer:** At night exists as a single layer in a region of about 145 to 400 km above the earth's surface. During the day it splits into two layers, F_1 and F_2.

The Ionosphere and Communication

The ionosphere plays an important role in communication. Radio waves can be reflected off the ionosphere allowing radio communications over long distances. However this process is more successful during the night – time.

Why Transmission is better at Night?

During the day: the ionosphere extends into lower atmosphere (D layer). In this layer there is high concentration of particles and so recombination of electrons and ions due to collision is more likely to occur. The leads to the radio waves being absorbed rather than reflected. Hence distant communications are poor during the day.

During the night: The D layer disappears due to decrease in ionization of molecules but recombination of electrons and ions still occurs at a fast rate. The radio waves are then reflected by E and F layers in which recombination of electrons and ions is rare hence there is less absorption of the radio waves.

EXAMPLES: SET C

Example 01: Necta 1985 P_1

(a) (i) Distinguish between P and S waves, state clearly the difference between their speeds in a medium.

(ii)Draw a schematic diagram showing how one station on the Earth's surface can receive P or S waves from a distant source and state which waves can be refracted by the Earth's outer core.

(b) (i) Give a summary of the origin and composition of the ionosphere.

(ii) What is the net electric charge in the ionosphere?

(iii) Show graphically how electron density changes with altitude in the ionosphere.

Answers

(a) (i) P – waves are longitudinal compression waves which can pass through solid, gas and liquid, whereas S - waves are transverse shearing waves which cannot pass thorough a fluid (gas or liquid)

The speed of P – waves in a medium is approximately twice that of the S – waves hence P – waves are faster than S – waves.

(ii) Refer the diagram for the seismic wave paths

(b) (i) Ionosphere is the upper part of the atmosphere. The ionosphere is formed due to the ionization of gaseous atoms as they absorb ultraviolet radiation from the sun, gamma and X-rays.

(ii) The net electric charge in the ionosphere is zero.

(iii) Variations of electron density in the ionosphere Electron density increases from D to F layer

Example 02: Necta 1988/1993 P_1

(a) What are the factors that influence the velocities of P – and S – waves?

(b) Explain briefly the characteristics property of seismic waves which is used to locate discontinuities in the earth's crust.

Answer

(a) The velocities of both P and S – waves are influenced by;

(i) Density of the rock material (Media),

(ii) Moduli of elasticity.

(b) Speed is the characteristic property of seismic waves that is used to locate discontinuities

Between the crust and mantle there is abrupt change of density, which shows an abrupt change in speed of both P – and S – waves, a Mohorovicic discontinuity exists here. Both P – and S

waves travels across this discontinuity.

Between the mantle and the core there is the Gutenberg discontinuity only P – waves travel this discontinuity.

Example 03: Necta 1989 P_1

(a) State three sources of heat energy in the interior of the earth.

(b) (i) How does temperature vary with depth of the Earth?

(ii) What are the factors that influence the flow of heat from the interior of the Earth?

Answers

(a) Refer notes

(b) (i) The temperature increases with increasing depth

(ii) The rate of heat flow (conduction) is given by

$$Q = KA\frac{d\theta}{dl}$$

The heat flow from the interior of the earth depends on:

Thermal conductivity of the rock,

Temperature gradient of the rock

Example 04: Necta 1989 P2

(a) What do you understand by the terms?

(i) Solar wind,

(ii) Magnetopause

(iii) Magnetosphere?

(b) What are the various factors that contribute to the Earth's magnetic field?

(c) (i) With the aid of a suitable diagram, illustrate the components of the earth's magnetic field at a given point P in the earth's atmosphere.

(ii) An electron whose kinetic energy is 10 eV is circulating at right angles to the earth's magnetic field whose uniform induction is 1.0 x 10 Wbm^{-2}. Calculate the radius of the orbit and its frequency in that orbit.

Answers

(a) (i) Solar wind is a continuous stream of fast moving charged particles in the atmosphere which are produced from flare (eruptions) from the sun:

(ii) Magnetopause is the upper boundary of the magnetosphere.

(iii) Magnetosphere is the upper most part of the exosphere consisting mainly of charged ions. These particles move under the influence of the earth's magnetic field.

(b) **Short term variations**: Disturbances in the magnetosphere due to solar emissions, these charged ions travel and in the ionosphere they form ring currents which give rise to a magnetic field.

Long term variations: The molten inner core of the earth is partly ionized. The movement of this ionized core causes a magnetic field which contributes to the earth's magnetic field.

(c) (i) refer notes (ii) refer electromagnetism

Example 05: Necta 1990 P_1

(a) Define the term "isoseismal line".

(b) Write short notes on each of the following regions of the atmosphere.

(i) Troposphere, (ii) Stratosphere, (iii) Exosphere

Answer: Refer notes

Example 06: Necta 1990 P_2

(a) Explain clearly how P and S – waves were used to ascertain that the outer core of the earth is in liquid form.

(b) Giving reasons, discuss the temperature variation in atmosphere (above the earth's surface).

Answers

(a) P – waves are longitudinal elastic, waves capable of passing through solids and liquids and S – waves are traverse elastic waves capable of a travelling through solids only.

As both waves are projected towards the surface from interior core only the P – waves are recorded. This shows that the outer core is in liquid form.

(b) From the ground level, the atmospheric temperature decreases steadily as altitude increases steadily as altitude increases up to the troposphere. Thereafter the temperature increases with altitude up to the stratosphere. The ozone of the stratosphere absorbs the incoming sun radiation hence the temperature increases. In the mesosphere there is no ozone thus there is a decrease (cooling) with increasing altitude. The heating effect of the earth ceases in the thermosphere so, the closer to the sun, the higher graph refer notes.

Example 07: Necta 1991 P_2

(a) List down four physical changes that took place at a location just before onset of an earthquake at that particular location.

(b) Give brief accounts of the processes that give rise to:

(i) The earth's magnetic field,

(ii) Volcanic eruptions

Answers

(a) Density of rocks, stresses faults and waves

(b) (i) Explain generation of the earth's field in the atmosphere and the outer core.

(ii) The seismic or earthquakes waves result from a fracture or sudden deformation of the earth's crust. Vast stresses do occur locally in the rocks being concentrated where the rocks are sliding over one another. In regions where pressure is reduced, pockets of molten rock called magma are formed. Once the rock has melted the pressure may force it into cracks and fissures in the surrounding solid rock. This may emerge above the surface as a lava flow or volcano.

Example 08: Necta 1992 P_1

(a) What do you understand by the term ionosphere?

(b) Explain how short wave long distance transmission and reception of radio waves is more effective at night than it is during the day time.

Answer

(b) In the day time, the base of the ionosphere (D-layer) is at lower heights where the high concentration of particles allows for ionization and recombination of ions by collision. Because of this, radio waves are absorbed rather than reflected, so distance communication is poor.

During the night time, the D – layer disappear, the base of the ionosphere is higher thus the recombination of ions is rare and so less absorption of waves occurs. Obliquely transmitted waves therefore can be reflected for distant reception.

Example 09: Necta 1993 P_2

(a) What is the origin of the earth's magnetic field?

(b) The diagram below shows the structure of the Earth. Name the parts indicated by the letter A to F.

Answer

(b) A represents Mohorovicic discontinuity

B represents Gutenberg discontinuity

C represents core

D represents Mantle

E represents Epicenter

F is not clear to interpret.

Example 10: Necta 1994 P_1

(a) Define the terms: angle of inclination (dip) and angle of declination (variation) as used in specifying the earth's magnetic field at any point.

(b)The earth's total resultant flux density B_R in a certain country is found to be 5.0×10^{-5} T and the horizontal component is B_H is 2.0×10^{-5} T. Calculate ;

(i) The vertical component, Bv, and

(ii) The angle of inclination in that country

Solution

(b) (i) The vertical component is given by

$$B_V = \sqrt{B_{E^2} - B_H{}^2} = \sqrt{(5x10^{-5})^2 - (2x10^{-5})^2}$$

$$B_V = 4.58X10^{-5}T$$

(ii) Angle of inclination is given by

$$\theta = \tan^{-1}\left(\frac{B_V}{B_H}\right) = \tan^{-1}\left(\frac{4.58x10^{-5}}{2.00x10^{-5}}\right) = 66.4^0$$

Example 11: Necta 1994 P_1

(a) (i) Name the lowest layer of the atmosphere and the lowest layer of the ionosphere.

(ii) State the importance of each of these layers.

(b) What is the ozone layer?

Answers

(a)(i) The lowest layer of the atmosphere is **troposphere** and the lowest layer of the ionosphere is called the **D – layer**.

(ii) The t troposphere supports life

The D – layer is important for communication purposes as it reflects radio waves.

(b)The ozone layer is within the stratosphere. In the ozone layer molecular oxygen (O_2) is dissociated into atomic oxygen (O) which is then reformed into ozone (O_3)

The ozone so formed absorbs ultra violet radiation thus protecting plants and shielding people from skin cancer and eye cataracts.

Example 12: Necta 1994 P_2

(a) Illustrate the component of the earth's magnetic field at a given point P in the earth's atmosphere by a suitable diagram.

(b) Using a tangent galvanometer, explain how you could determine the earth's magnetic field.

Answers

Example 13: Necta 1995 P_1

(a) (i) which region of the solid earth includes the e earth's centre?

(ii) On which region of the solid earth do the continent rests directly?

(iii) Which region of the ionosphere has the highest electron density?

(b) Briefly explain how earthquake can be detected

Answers

(a) (i) inner core (ii) crust (iii) F – region

(b) Detection of earthquake is done by recording or measuring the seismic waves generated by the earthquakes. These waves are recorded by instrument called seismograph.

Example 14: Necta 1995 P_2

(a) Draw a well labeled diagram which shows the interior structure of the earth. Indicate also which part of the interior are in solid form and which are in liquid form.

(b) Name and distinguish the type of waves that are produced by an earthquake.

(c) Briefly describe the three ways in which signal form ground based transmitter can reach the receiver.

Answers

(a) There are four types of seismic waves:

Body waves – divided into P and S – waves

Surface waves – divided into love and Rayleigh

(b) A telecommunication problem.

Ground wave, sky wave and space waves

Example 15: Necta 1998 P_1

(a) State any three magnetic components of the earth's magnetic field

(b) The horizontal and vertical components of the earth's magnetic field at a certain location are; 2.73 x 10^{-5} and 2.1 x 10^{-5}T respectively. Determine the earth's magnetic field at the location and its angle of inclination θ

Solution

(a) Components of the earth magnetic field are:

Vertical component (which point vertically downward)

Horizontal component which comprise lf:

Eastly component (towards geographic north pole)

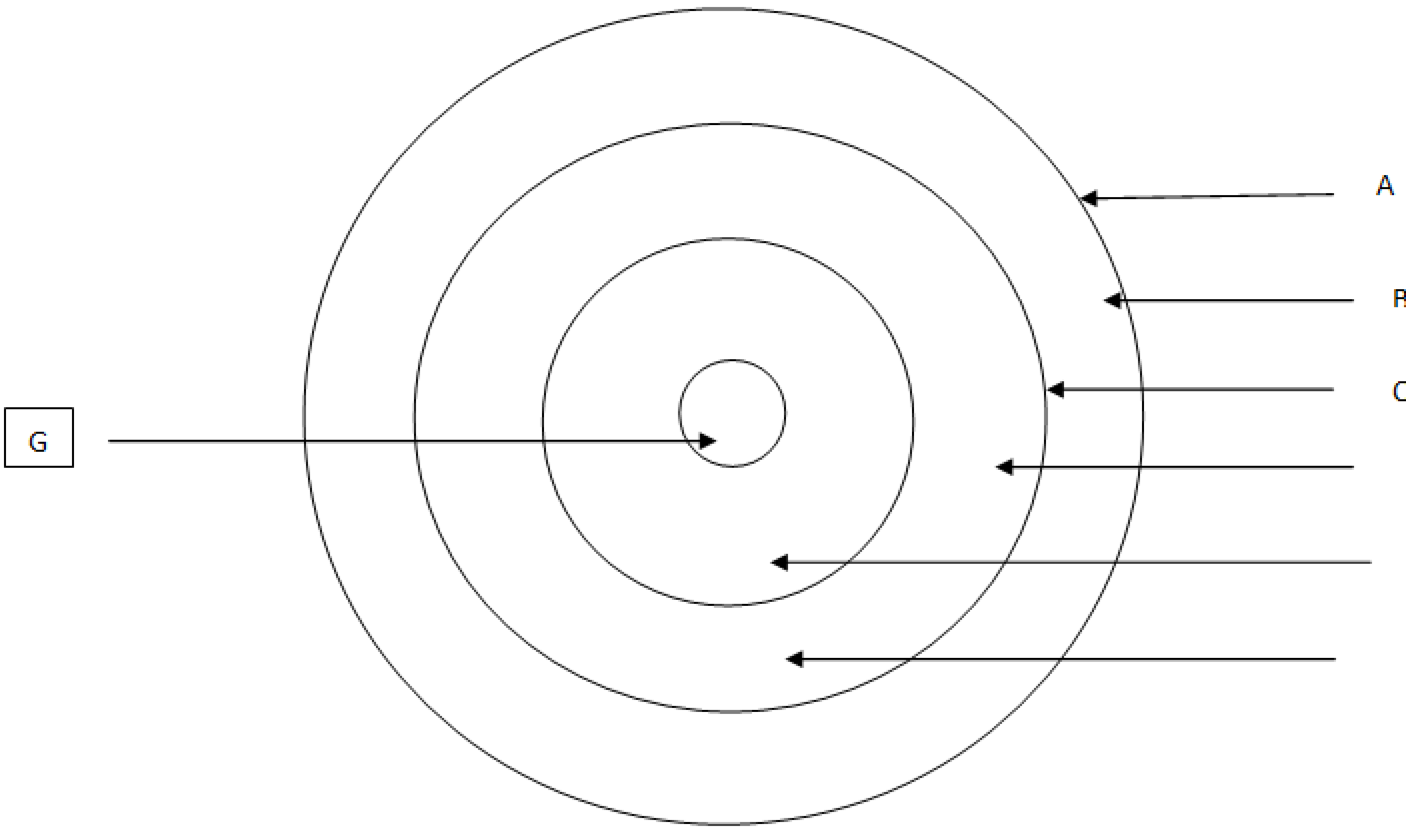
Northly component (towards magnetic north pole)

(b) $B_E = 3.44 \times 10^{-5} Tan\theta = 37.66^0$

Example 16: Necta 1998 P_1 B

(a) What is the origin of the earth's magnetic field?

(b) The following diagram shows the main layers forming the interior of the earth name the layers indicated by letters A to G.



Answers

(a) Refer notes

(b) A = Earth's surface, B = Crust, C = Moho discontinuity, D = Gutenberg discontinuity, E = outer core, F = Mantle and G = inner core.

Example 17: Necta 1998 P_2 B

(a) Explain the following terms; Earthquake, Earthquake focus, Epicenter and body waves.

(b) List down three (3) sources of earthquakes,

(c) (i) Define ionosphere

(ii) Mention the ionosphere layers that exist during the day time

(iii) Give the reason for better reception of radio waves for high frequency signal of night than during day time.

(d) Explain briefly three different types of radio waves traveling from a transmitting station to a receiving antenna.

Answers

(a) Refer notes

(b) Refer notes

(c) (i) During the day time all the layers D,E,F_1, and F_2 – layers exists.

(ii) Refer Necta 1992 (b)

(d) Ground (surface wave)

Space wave

Sky waves) (refer telecommunication notes)

Example 18: nectar 2000 P_1

(a) With reference to an earthquake on a certain point of the earth explain the terms 'focus' and 'Epicenter'

(b) What is importance of the following layer of the atmosphere?

(i) The lowest layer

(ii) The ionosphere

(c) (i) Describe two ways by which seismic waves may be produced.

(ii) Describe briefly the meaning and application of "seismic prospecting".

Answers

(a) Refer notes

(b) (i) Importance of troposphere is supports life on earth

(ii) Ionosphere enhances communication over long distances.

(c) (i) Describe any two causes of earth quake

(ii) Seismic prospecting is an artificial production of seismic waves purposely for searching underground fuels and oils or gases

Example 19: Necta 2001 P_1

(a) (i) Define the terms "angle of declination" as used in the specification of the earth's magnetic field at a point

(ii) The horizontal component of the earth's magnetic field at a location was found to be 26.0 μT while the angle of inclination was 59^0. Find the magnitude of the field and the vertical component of the field at the location

(b) (i) Define an earthquake

(ii) Distinguish between P and S waves. What factors influence their velocities?

Answers

(a) (i) Refer notes

(ii) $B_R = 50.48\ \mu T, B_V = 43.27\ \mu T$

(b) The velocities of P and S waves are influenced by;

Density, ρ of the media

Shear modulus, η of the media, and

Bulk modulus, B of the media.

Example 20: Necta 2002 P1

(a) (i) What is the importance of ionosphere to mankind?

(ii) Explain why transmission of radio waves is better at night than at day time.

(b) (i) What is an earthquake?

(ii) Explain briefly any four (4) causes of earthquake

Example 21: Necta 2003 P_2

(a) Explain the following:

(i) Earthquake (ii) Earthquake focus (iii) The epicenter.

(b) List down three sources of earthquake

(c) (i) Define the ionosphere

(ii) State the ionosphere layer that exists during day time.

(iii) Give the reason for better waves reception for light frequencies signal at night than during the day time

Example 22: Necta 2004 P_1

(a) (i) Explain the terms epicenter and focus as applied to earthquake.

(ii) State any four (4) indications that may predict the occurrence of an earthquake.

(iii) State and explain two variations of the earth magnetic field.

(iv) State one necessary precaution to be taken to people living in a region with a high risk of occurrence of earthquakes.

(b) Explain the following

(i) Solar wind (ii) Magnetopause (iii) Ionosphere.

Example 23: Necta 2005 P_1

(a) Define the following terms

(i) Epicentral distance (ii) Body wave (iii) Seismograph

(b) (i) explain the meaning of reflection seismology state its application

(ii) Show how the magnetic field within the atmosphere is generated?

(c) (i) Name the lowest layers of the atmosphere and the ionosphere

(ii) State their importance

Answers

(a) (i) Lowest layer of atmosphere is troposphere and that of the ionosphere is the D – layer.

Example 24: Necta 2006 P_1

(a) (i) State two (2) ways by which seismic wave may be produced

(ii) What is seismic prospecting?

(b) (i) Discuss briefly the importance of the lowest layer of the atmosphere and the ionosphere.

(ii) Sketch the temperature against altitude curve for the atmosphere indicating the important atmospheric layers.

(iii)The average velocity of P – waves through the earth's solid core is $8kms^{-1}$. If the average density of the earth's rock is $5.5 \times 10^3 kgm^{-3}$ find the average bulk modulus of the earth's rock.

Answer

(a) (i) Causes of an earthquake

(b) (ii) using the formula

$$B = \rho V^2 = 5.5 \times 10^3 \times (8000)^2 = 3.52 \times 10^n N/m$$

Example 25: Necta 2007 P_1

(a) (i) What are the differences between P and S waves?

(ii) Explain how the two terms of waves (P and S) can be used in studying the internal structure of the earth.

(b) Write short notes on the following terms in relation to the changes in the earth's magnetic field; long term (secular) changes, short – period (regular) changes, and short – term (irregular) changes.

(c) (i) What is geomagnetic micro pulsation?

(ii) Give a summary of location, constitution and practical uses of the stratosphere, ionosphere and mesosphere.

Answers

(c) (i) Geomagnetic micro pulsation are small rapid changes in the earth's magnetic field. They have periods between 0.2 second and 10 minutes and intensities less than 0.01% of the minimum field.

Example 26: Necta 2008 P_1

(a) Define the following terms:

(i) Earthquake (ii) atmosphere

(b) Distinguish between body waves and surface waves that are produced by an earthquake.

(c) (i) Define the terms epicenter and focus as applied to earthquake.

(ii) Draw a well labeled diagram which shows the interior structure of the earth.

Example 27: Necta 2009 P_1

(a) (i) What is meant by the shadow zone?

(ii) Why does the shadow zone occur?

(b) (i) Name the lowest layer of the atmosphere and the lowest layer of the ionosphere.

(ii) State the importance of each of these layers in b (i) above

(iii) Explain briefly the reason for better reception of radio waves for high frequency signals at night times than during day times.

(c) State the sources of heat energy in the interior of the earth.

Example 28: Necta 2010 P_1

(a) (i) Explain the terms: earthquake, earthquake focus and epicenter.

(ii) Describe clearly how P and S waves are used to ascertain that the outer core of the Earth is in liquid form.

(b) (i) Define the ionosphere and give one basic use of it.

(ii) Why is the ionosphere obstacle to radio astronomy?

Example 29: Necta 2011 P_1

(a) (i) Define the following terms: Geophysics, Atmosphere and Epicenter

(ii) Write down brief notes on the location, composition and importance of the following:

Troposphere, Stratosphere, Mesosphere and Thermosphere

(b) (i) Draw sketch diagram showing the working part of a Seismometer.

(ii) Explain how temperature varies with both altitude and depth of the Earth.

(iii) Write down two factors that governs heat flow from the interior of the Earth.

Example 30: Necta 2012 P_1

(a) (i) Name three layers of the atmosphere

(ii) Describe any two major zones of the earth.

(b) (i) What are the factors that influence the velocities of P and S waves?

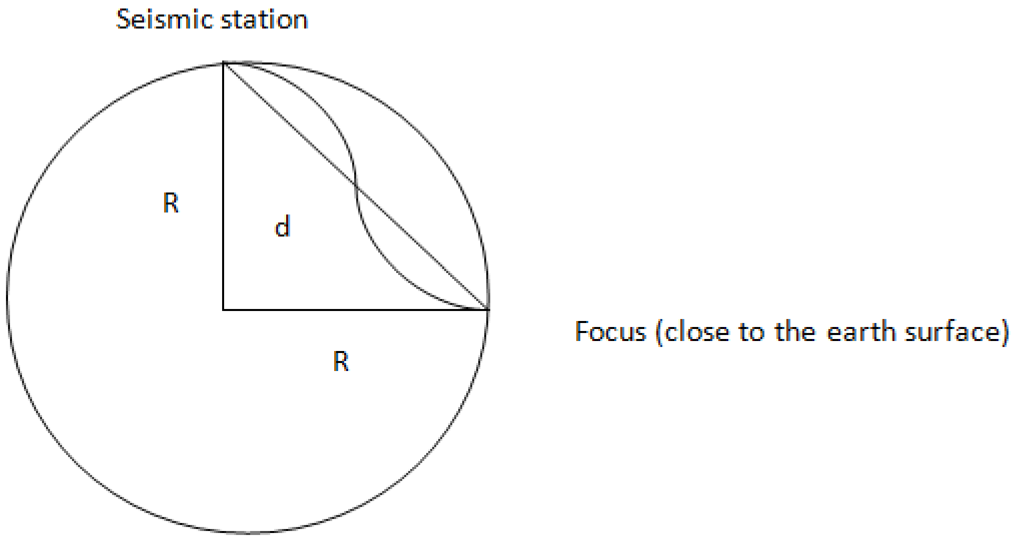
(ii) The P and S waves from an earthquake with a focus near the earth's surface travel through the earth at nearly a constant speed of 8 km/s and 6 km/s respectively. If there is no reflection and refraction of waves how long is the delay between the arrivals of successive waves at a seismic monitoring station at 90^0 in the latitude from the epicenter of the earthquake?

Solution

(a) (ii) any two of core, mantle, crust, hydrosphere, atmosphere

(b) (i) the density of rock, moduli of elasticity of rock material.

(ii) Illustration (R = earth radius)



Distance travelled by the waves (distance between focus and seismic station) is

$$\boldsymbol{d = R\sqrt{2} = 6.4\times10^{6} \times \sqrt{2} = 9.05\times10^{6} m}$$

Time taken by P – waves to arrive at the station is

$$t_1 = \frac{d}{speed} = \frac{9.05 \times 10^6}{8 \times 10^3} = 1131.25sec = 18.9min$$

Time taken by the waves to arrive at the station is

$$t_1 = \frac{d}{speed} = \frac{9.05 \times 10^6}{6 \times 10^3} = 1508.33sec = 25.1min$$

The time interval between the arrival of the two waves is $t = t_2 - t_1 = 25.1 = 18.9 = 6.2$ minutes.

Example 31: Necta 2012 P_1

(a) (i) What do you understand by the word environmental physics?

(ii) Briefly explain three effects of seismic waves.

(b) (i) Mention three types of environmental pollution

(ii) Explain on the following climatic factors which influence plant growth: Temperature, Relative humidity and wind.

Example 32: Necta 2013 P_1

(a) (i) The main interior of the earth core is believed to be in molten form. What seismic evidence supports this belief?

(ii) Explain why the small ozone layer on the top of the stratosphere is crucial for human survival

(b) Electrical properties of the atmosphere are significantly exhibited in the ionosphere.

(i) What is the layer composed of and what you think is the origin of such constituents

(ii) Mentioned two uses of the ionosphere

(c) Briefly explain why long distance radio broadcasts make use of short wave

Answers

(a) (i) When P and S seismic waves are sent from one side of earth to the other, only P waves can be detected on the other side. The fact that S waves do not travel through the core provides evidence for the existence of a liquid core.

(ii) Ozone absorbs harmful radiation from the sun. The Ozone projects plant and shield people from skin cancer and eye cataracts.

(b) (i) The layer is composed of free electrons and positive ions. The ionosphere is created by atoms absorbing UV radiation, gamma and x-rays.

(ii) Uses of the ionosphere

Ionosphere supports radio communication over long distances

Particles in the ionosphere absorbs U.V radiation gamma and X-rays, thus protecting people from harmful effects of these radiations

(c) Refer telecommunication notes.

Example 33: Necta 2013 P_1

(a) Briefly explain on the following types of environmental pollution:

(i) Thermal pollution

(ii) Water pollution

(b) Describe the soil temperature with regard to agriculture, physics which causes lower crop growth at a particular area

Answers

(b) High soil temperature causes the crop roots to rot, this leads to insufficient water supply to plant leaves and hence lower the growth of crop.

Lower soil temperature inactivates soil organisms. Decomposition of organic matter is lowered and hence the supply of nutrients to crop which in turn lead to lower crop growth.

TRY YOURSELF

(a) (i) What are auroras?

(ii) Define the homosphere

(b) (i) What are the factors which contribute toward volcanic eruptions?

(ii) What are the effects of volcanic eruptions?

(iii) What are lahars?

Lahars are rapidly flowing mixtures of rock debris and water that originate on the slopes of a volcano. They are also referred to as volcanic mudflows or debris flow. Volcanic eruptions may directly trigger one of more lahars by quickly melting snow and on a volcano or eject water from a crater lake. The form in a variety of at always including through intense rainfall on loose volcano rock deposits and as a consequence of debris of debris avalanches

ENVIRONMENTAL POLLUTION

Pollution is the addition of unwanted materials or pollutants into the environment.

Pollutant is any substance that does not belong in the natural system and disrupts the natural balance.

Type of Environmental pollution

(a) Air pollution (atmospheric pollution)

(b) Water pollution (hydrosphere pollution)

(c) Land (soil) pollution

(d) Noise pollution

(e) Thermal pollution

ATMOSPHERIC (AIR) POLLUTION

AIR POLLUTION

This is a form of environmental pollution caused by the release of gaseous materials and dust particles in the atmosphere. The main pollutants found in the air we breathe include, particulate matter, lead, ground-level ozone, heavy metals, sulphur dioxide, benzene, carbon monoxide and nitrogen dioxide

Causes of Air Pollution

Man made causes:

(i) Clearing (deforestation) and burning of vegetation. This releases carbon dioxide in the atmosphere and dust particles which may be carried by wind on bare land.

(ii) Burning of fuels: This releases green house gases in the atmosphere. Fuels are burnt in cars, power stations and industries.

(iii) Construction activities, like road, building, etc construction, can add dust particles in the atmosphere.

(iv) Automobile exhausts. Car, trains, etc burns fuels as they move his releases pollutant gases in the atmosphere.

(v) Smokes from industries also pollute the atmosphere.

(vi) Agriculture activities. The use of pesticide/insecticides pollutes the air.

(vii) Mining activities

Natural causes:

(a) Volcanic eruptions - release smoke and dust particles in the atmosphere

(b) Wind storms – carry land particles into the air

(c) Temperature inversion – the increase in temperature in the stratosphere causes high altitude particles to sink to the troposphere

WATER POLLUTION

Water Pollution is the degradation of water quality in a manner that disrupts/prevents its intended or original use.

Surface Water or Ground water may be polluted

Causes of water pollution

(i) Disposal of untreated sewage (industrial or hospital, etc) into the water bodies.

(ii) Wind may introduce dust particles into water from the land.

(iii) Agriculture activities near water bodies. Chemical used during farming may be taken to the water bodies by the rain water.

(iv) Oil spilt. The leakage of oil in under water oil pipe, leakage from boats, ships, etc pollutes the water.

(v) Fishing by using chemicals (dynamite fishing).

(vi) Volcanic activities along water bodies.

(vii) Quarrying along the coast.

LAND (SOIL) POLLUTION

Soil pollution is defined as the build – up in soils of persistent toxic compounds, chemicals, salts, radioactive materials, or disease causing agents which have adverse effects on plant growth and animal health.

A soil pollutant is any factor which deteriorates the quality, texture and mineral content of the soil or which disturbs the biological balance of the organisms in the soil.

Causes of soil pollution

(a) Chemical from industries

(b) Acid rain – this increase soil acidity

(c) Farming activities which make use of insecticides/pesticides

(d) Mining activities – increase rock sediment into the soil.

NOISE POLLUTION

Noise pollution is any disorganized loud sound.

Causes of noise pollution

(a) Noise from factories and workshops

(b) Thunderstorm explosion of bombs

(c) Low level flying aircraft

(d) Radio on large volumes

(e) Slamming of doors

THERMAL POLLUTION

Thermal Pollution is a form of environmental pollution caused by the release of waste heat into water or air

Causes of Thermal Pollution

(a) i. Hot gases released by industries and motor vehicles warm the environment.

ii. Hot wasteful liquid from industries pumped to a river, lake, or other waterway

Effects of thermal pollution

(a) Heat introduced into water can make the water so hot that no living thing can survive in it

(b) Hot gases introduced in the atmosphere leads to green house effects.

Solutions of thermal pollution

(a)One is a cooling pond into which heated waste water is released before it enters a natural waterway. The cooling pond permits evaporation of some water, carrying heat into the air and thus releasing cooler water into the waterway

(b)The cooling tower method – either wet or dry – which also transfers heat to the air. In both types, heated water is introduced into a tower through which air is blown, and some heat is passed to the air.

PARTICULATE MATTER IN THE ATMOSPHERE (AEROSOLS)

Particulate matter (aerosol) is the general term used for a mixture of fine solid particles and liquid droplets found in the air.

Haze aerosol is frequently encountered in optical studies and includes any airborne particles that affect visibility.

Classification of Particulate

Particulates matter are classified in accordance with its formation mechanisms

(i) Primary particles (ii) Secondary particles

Primary particles are directly emitted into the atmosphere from their sources while secondary particles are formed after chemical transformation of their gaseous precursors. Chemical reactions transform primary pollutants (emitted by the sources) to secondary pollutants that are formed within the atmosphere. Ozone, sulfate aerosols, nitrates, are examples of secondary pollutants.

Particulate matters in the atmosphere are categorized as:

(i) Minerals, 72 – 91%, e.g. soil particles, hematite, mica, and talc;

(ii) Combustion products, 1 – 10%, e.g. coal and oil soot, fly ash, burned paper.

(iii) Biological materials 2 – 10% e.g. pollen, spores, starch, plant tissues and diatoms

(iv) Miscellaneous matter, trace – 8% e.g. salt, rubber, iron/steel, paint pigment and humus

Dust refers to a relatively course range of solid particles (diameter, d >1pm), produced by disintegration of minerals or from re-suspension by wind when sun blasting of soil particles may often causes comminuting.

Smokes and fumes are fine particles formed from the gas phase by condensation. In the case of fume the particles are generally from 0.01 – 1 pm diameter, and are often observed as agglomerates of smaller particles. Suspended particulate matter < 15 *pm* and diameter is usually defined as smoke.

Mists and fogs are liquid droplets Mists (d > 40 pm) and fogs (d = 5 – 40 pm).

Advantages of particulate matter in the atmosphere

Aerosols acts as nuclei were water vapour collects during the formation of water droplets through condensation.

Disadvantages of particulate matter in the atmosphere

(a) Cause global warming

(b) Can block the atmosphere (impair visibility)

(c) Once deposited on leaves they block stomata and hence no photosynthesis for plant

(d) Changing the timing and location of traditional rainfall patterns

(e) Can lead to development of heart and lung diseases.

TRANSPORT MECHANISMS OF ATMOSPHERIC POLLUTANT

The transport of pollutants by the wind

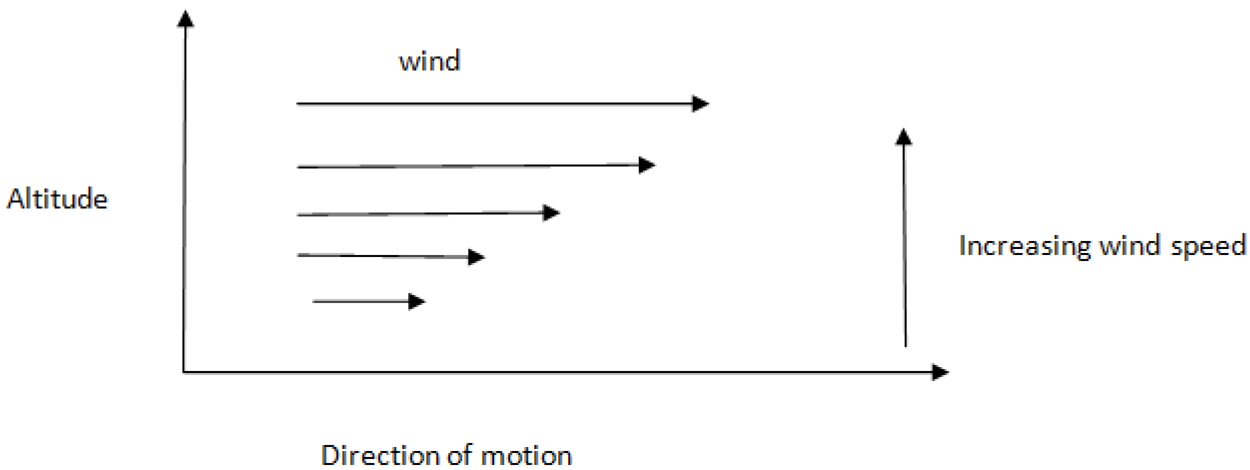
The three transport processes that influence the regional dispersion are;

(a) Wind speed (shear)

(b) Directional veer (change in direction fo wind), and

(c) Eddy motion (eddy diffusion).

Wind shear: The vertical gradient of wind speed (i.e. wind shear is responsible for lagging of low elevation pollutants behind those in the upper layers.



Directional veer: The directional veer with height causes lateral displacement of a vertically uniform puff.

The eddy motion is the vertical transport of pollutants from region of high concentration to low concentration. Eddy motions are due to random vertical and horizontal fluctuations caused by thermal and mechanical turbulence.

Both the transport speed and direction for an air parcel vary from day to day.

Stratosphere – troposphere interchange

Temperature inversion at the tropopause causes an interchange of particulate matters between Stratospheres – troposphere boundary.

EFFECTS OF POLLUTION ON VISIBILITY

Atmospheric pollution results into a reduction in visual range in the atmosphere. The reduction is visual range caused by an increase in airborne particles that affects light scattering and attenuation involves both primary and secondary aerosols, and may be experienced in rural as well as urban area.

EFFECTS OF ATMOSPHERIC POLLUTION ON THE GLOBAL ALBEDO AND CLIMATE

Increases in particulate matter in the atmosphere may:

(a) affect cloud droplet formation and precipitation,

(b) Reduce the amount of solar radiation that reaches the ground

(c) Reduce the cooling of the surface layer of the earth at night and influence the global albedo.

However, controversy still remains as to whether the presence of particulate material exerts a net warming or cooling effect to enhance or offset the global warming predicted from increases in CO_2 and chlorofluoro methanes in the atmosphere. In addition, considerable changes in global and surface albedo have been caused by deforestation, salinization, and desertification.

Global warming is the increase of the average temperatures near or on the surface of the earth as a result of greenhouse effect.

GLOBAL WARMING

Global warming is the increase of the average temperatures near or on the surface of the earth as a result of greenhouse effect.

Greenhouse effect

Greenhouse effect is the process in which the emission of radiation by the atmosphere warms the earth's surface.

Greenhouse gases include carbon dioxide, methane, chlorofluorocarbons and dinitrogen oxide.

When heat from the sun reaches the earth's surface in form of sunlight, some of it is absorbed by the earth. The rest is radiated back to the atmosphere at a long wavelength than the incoming sunlight. Some of these longer wavelengths are absorbed by the greenhouse gases in the atmosphere before they are lost out of space. The greenhouse gases reflect the heat back to the earth and warm the environment.

Sources of greenhouse gases in the atmosphere

(a) Carbon dioxide is added in the atmosphere by:

(i) Clearing and burning of vegetation

(ii) Burning of fossil fuels

(b) Methane is added in the atmosphere by:

(i) Agricultural activities;

(ii) The mining of coal and oil

(c) Dinitrogen oxide is added in the atmosphere by:

(i) Combustion of fossil fuels in vehicles and power station

(ii) Use of nitrogenous fertilizer, and

(iii) The burning of vegetation and animal waste

(d) Sources of chlorofluorocarbon include fridge, air conditioners and aerosols.

Effects of Global Warming

(a) Increase in the temperature of the oceans,

(b) Rise in sea levels,

(c) Change in world's climatic patterns,

(d) Acidification of the oceans,

(e) Extreme weather events like flood, droughts, heat waves, hurricanes and tornadoes

(f) Higher or lower agriculture yields,

(g) Melting of Arctic ice and snow caps. This causes landslides, flash floods and glacial lake overflow,

(h) Extinction of some animals and plant species,

(i) Increase in the range of disease vectors (organisms that transmit disease).

Solution to Global Warming

(a) Use of cleaner alternative sources of energy such as solar and wind,

(b) Put in place energy conservation measures to reduce the use of fossil fuel,

(c) Planting trees that would absorb carbon dioxide

(a) Use of cleaner alternative sources of energy such as solar and wind,

NUCLEAR WASTE AND METHODS OF DISPOSAL

Nuclear wastes are the chemical products (solid, liquid and or gases) of nuclear reactions in the nuclear reactor.

Categories of radioactive waste

For the purpose of disposal, radioactive waste is divided into the following categories:

(a) High – level waste (HLW): spent fuel (SF) not destined for reprocessing; vitrified fission product solutions from reprocessing of spent fuel.

(b) Alpha – toxic waste (STW): waste with a content of alpha – emitters exceeding a value of 20,000 Becquerel's per gram of conditioned waste.

(c) Low – and intermediate – level (L/ILW): all other radioactive waste.

Nuclear Waste Disposal

(a) Deep geological repository: for spent fuel and vitrified fission product solution product solutions from reprocessing. The products are buried deep into the earth.

(b) Recycling of the nuclear waste.

ATOMIC PHYSICS

1. 1. Structure of the ATOM
   -Describe the Rutherford and bohr's models of the atom.
   -Analyze atomic energy levels.
   -Discuses the hydrogen energy levels, and derives expressions for the energy levels.
   -Perform experiment to determine wavelength in the Balmer series of the hydrogen spectrum.

   2. Quantum physics
   - Describe failures of classical physics.
   - Explain Planck's quantum theory of blackbody radiation.
   - Spectral distribution of black body radiation.
   - Explain Einstein's quantum theory of light.
   - Perform experiment to determine the Planck's constant.

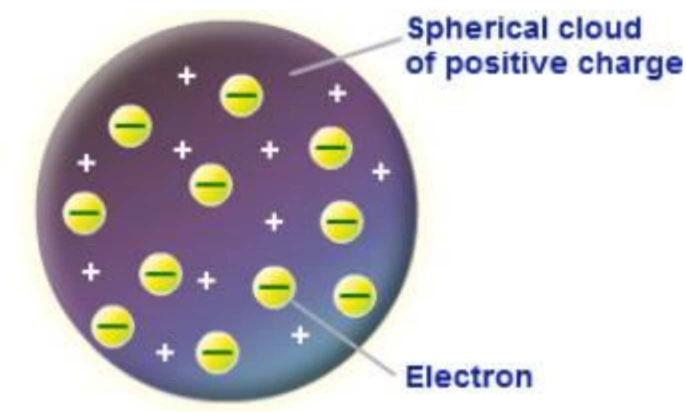
- Account for the photoelectric effect phenomenon.
- Deduce stopping potential threshold frequency and work function of a metal.
- Explain the photo electric effect.
- Deduce de Broglie wave length for electron.
- Discus the wave- particle duality of electron.
- Derive de Broglie's wavelength for the electron.
- Describe production and uses of x- rays.
- Uses in medicine, industry and in sample analysis.

3. 3. LASER
   - Describe production of laser light.
   - Explain properties of laser light.
   - Distinguish types of lasers.
   - Discuss methods of pumping in laser production.
   - Identify application of laser light.

4. 4. Nuclear Physics
   - Describe the structure of the nucleus.
     * Review Rutherford experiment.
   - Determine half life and the decay constant (λ) of a radioactive substance.
   - Explain the relation of nuclear mass and binding energy.
     * Discuss Einstein's mass energy equation.
     * Apply Einstein's mass energy relation to determine the biding energy of nuclei.
   - Identify criteria for stable and unstable nucleus.
     * Analyze the neutron and proton ratio and plot N against Z for radioactive elements.
     * Establish criteria for stable and unstable nuclei
   - Identify uses and hazards of radioisotopes
     * Application
     * Hazards
   - Distinguish between fission and fusion processes
     * Meaning of fission and fusion
     * Calculate the energy released in a nuclear fission
     * Calculate the energy absorbed in nuclear fission
     * Describe the application of nuclear fission and fusion
   - Describe operation of a nuclease reactor
     * Construction and operation of nucleus reactor for safe application

THOMSON'S MODEL OF ATOM

According to Thomson an atom is a positive charged sphere in which the entire mass and positive charge of the atom is uniform distributed with negative electrons embedded in it as shown.



The number of electrons is such that their negative charge is equal to the positive charge of the atom. This atom is electrically neutral.

This model was called Thomson's plum pudding model because the negatively charge electrons (the plums) were embedded in a sphere of uniform positive charge (the pudding).

Drawbacks of this Model

1.It could not provide stability to the atom it is because the positive and negative charges are stationary and will be drawn towards each other, thus destroying the individual negative and positive charges.

2. It could not explain the presence of discrete spectral lines emitted by hydrogen and other atoms.

RUTHER FORD'S MODEL OF ATOM

The salient features of this model are

(i)Every atom consist of a tiny central core, called the nucleus which contains all the atom's positive charge and most of its mass (99.9%).

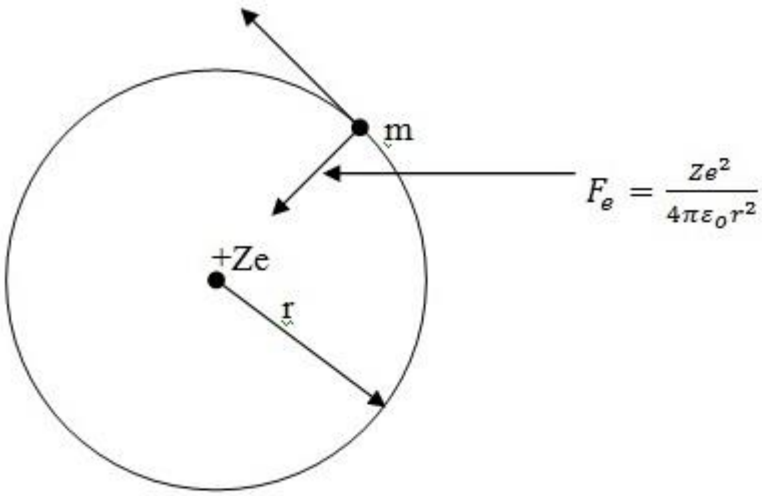
(ii)The radius of the nucleus is of the orde of 10^{-15}m and that of the atom is of the order 10^{-10}m.

Therefore nucleus occupies only an extremely small portion of the size of atom

(iii) The electrons occupy the space outside the nucleus. Since an atom is electrically neutral the positive charge on the nucleus is equal to the negative charge on electrons surrounding the nucleus.

(iv) Electrons are not stationary but revolve around the nucleus in various circular orbits as do the planets around the sun.

In this way Rutherford provided stability to the atom. It is because the centripetal force required by the electrons for revolution is provided by the electrostatic force of attraction between electrons and the nucleus.



e = charge on electron

z =total number of protons in the nucleus

m=mass of the electron

r =distance of electron from the nucleus

v= linear velocity of the electron

Force of attraction between electron and the nucleus is

$$Fe = \frac{(Ze)e}{4\pi\varepsilon_0 r^2}$$

$$Fe = \frac{Ze^2}{4\pi\varepsilon_0 r^2}$$

where Ze is a nuclear charge

The centripetal force required to keep the electron moving in circular path is

$$F_c = \frac{mv^2}{r}$$

Since the atom is stable $F_e = F_c$

$$\frac{mv^2}{r} = \frac{Ze^2}{4\pi\varepsilon_0 r^2}$$

$$mv^2 = \frac{Ze^2}{4\pi\varepsilon_0 r} \ldots\ldots\ldots\ldots\ldots\ldots\ldots. (1)$$

Kinetic energy of electron

$$KE = \frac{1}{2}mv^2$$

From equation (1)

$$mV^2 = \frac{Ze^2}{4\pi\varepsilon_0 r}$$

$$2K.E = \frac{Ze^2}{4\pi\varepsilon_0 r}$$

$$mv^2 = \frac{Ze^2}{4\pi\varepsilon_0 r}$$

$$2KE = \frac{Ze^2}{4\pi\varepsilon_0 r^2}$$

$$KE = \frac{Ze^2}{8\pi\varepsilon_0 r^2}$$

Potential energy of electron

$$PE = \frac{(Ze)(-e)}{4\pi\varepsilon_0 r}$$

$$PE = \frac{-Ze^2}{4\pi\varepsilon_0 r}$$

Total energy of electron

$$E = KE + PE$$

$$E = \frac{Ze^2}{8\pi\varepsilon_0 r} + \left(\frac{-Ze^2}{4\pi\varepsilon_0 r}\right)$$

$$E = \frac{-Ze^2}{8\pi\varepsilon_0 r}$$

The total energy of electron in the orbit is negative hence the electron is bound to the positive nucleus

For hydrogen Atom

For hydrogen atom z= 1. Therefore K. E and P.E OF electron in hydrogen atom are

$$KE = \frac{e^2}{8\pi\varepsilon_0 r}$$

$$PE = \frac{-e^2}{4\pi\varepsilon_0 r}$$

The total energy of electrons hydrogen atom is

$$E = K.E + P.E$$

$$E = \frac{e^2}{8\pi\varepsilon_0 r} - \frac{e^2}{4\pi\varepsilon_0 r}$$

$$E = \frac{-e^2}{8\pi\varepsilon_0 r}$$

Limitations of Rutherford's model of atom

1. According to Maxwell's theory of electromagnetism a charge that is accelerating radiates energy as electromagnetic waves
The electron moving around the nucleus is under constant accelerating radiates energy as electromagnetic waves.

- Due to this continuous loss of energy the electrons in Rutherford's model were bound to spiral towards the nucleus and fall into it when all of their rotational energy were radiated

- Hence Rutherford's atomic model cannot be stable while in actual practice, an atom is stable

This shows that Rutherford's model is not correct

1. During inward spiraling the electron's angular frequency continuously increases

- As result electrons will radiate electromagnetic waves of all frequency i.e. the spectrum of these waves will be continuous in nature because these are continuous loss of energy.

- But this is contrary to observation experiments shows that an atom emits line spectra and each line corresponds to a particular frequency or wavelength.

Rutherford's model failed to account for the stability of the atom. It was also unable to explain the emission of line spectra.

BOHR'S MODEL OF ATOM

According to Bohr's atomic model, the revolving electrons in the atom do not emit radiations under all conditions. They do so under certain conditions as expalined by him in his model.

BASIC POSTULATES OF BOHR'S MODEL OF ATOM

1. The electrons revolve around the nucleus of the atom in circular orbits. The centripetal force required by electrons for revolution is provided by the electrostatic force of attraction between the electrons and the nucleus.

2. An electron can revolve only in those circular orbits in which its angular momentum is an integral multiple of $h/2\pi$

$$mvr = \frac{nh}{2\pi}$$

h= Plank's constant.

Radius of orbit r

From,

$$mvr = \frac{nh}{2\pi}$$

$$r = \frac{nh}{2\pi mv}$$

Since *n* is a whole number only certain value of r is allowed.

Thus according to Bohr, an electron can revolve only in certain orbits of definite radii not in all these are called stable orbits (stationary orbit)

According to this postulate the angular momentum of the electron does not have continuous range i.e. the angular momentum of the revolving electron is quantized.

While revolving in stable or stationary orbits the electrons do not radiate energy inspite of their acceleration towards the centre of the orbit.

- For this reason these permitted orbits are called stable or stationary orbits.

e= charge on electron

m= mass of electron

r_n= radius of the n^{th} orbit

v_n= velocity of electron in the n^{th} orbit

Z= number of positive charge (protons)

Positive charge on nucleus Ze

RADIUS OF BOHR'S STATIONARY ORBITS

As the centripetal force is provided by the electrostatic force of attraction between the nucleus and electron.

$$\frac{m{v_n}^2}{r_n} = \frac{1}{4\pi\varepsilon_0}.\frac{(Ze)e}{{r_n}^2}$$

$$m{v_n}^2 = \frac{1}{4\pi\varepsilon_0}.\frac{Ze^2}{r_n} \ldots\ldots\ldots\ldots\ldots.(i)$$

According to Bohr

$$mv_n r_n = n\frac{h}{2\pi} \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots. (ii)$$

Consider equation

$$m{v_n}^2 = \frac{1}{4\pi\varepsilon_0}.\frac{Ze^2}{r_n}$$

$$mv_n r_n = \frac{1}{4\pi\varepsilon_0}.Ze^2$$

Take equation (ii) square it

$$mv_n r_n = n\frac{h}{2\pi} \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots. (ii)$$

$$m^2{v_n}^2{r_n}^2 = \frac{n^2h^2}{4\pi^2}$$

Take equation (iii) $\div$ equation (i)

$$\frac{M^2V_n^2r_n^2}{MV_n^2} = \frac{n^2h^2}{4\pi^2} . \frac{4\pi\varepsilon_o r_n}{Ze^2}$$

$$r_n = \left(\frac{\varepsilon_0 h^2}{\pi m e^2}\right)\frac{n^2}{Z}$$

It is clear that $r_n \ \alpha$ n^2, radii of the stationary orbits are in ratio 1^2: 2^2:3^2 ………..clearly the stationary orbits are not equally spaced.

For hydrogen atom

For hydrogen atom z= 1, so that equation become

$$r_n = \left(\frac{\varepsilon_0 h^2}{\pi m e^2}\right)n^2$$

Now $\frac{\varepsilon o\, h^2}{\pi m e^2}$ = 0.53 x10^{-10}m

$m = m_e$

e = electronic charge r_n = (0. 53 x 10^-10) n^2 metres

$For\ n = 1$

$r_1 = (0.53 \times 10^{-10}) \times (1^2)$

$r_1 = 0.53 \times 10^{-10}\text{m}$

$r_1 = 0.53\text{Å}$

$For\ n = 2$

$r_2 = (0.53 \times 10^{-10}) \times (2^2)$

$r_2 = 2.12 \times 10^{-10}$

$r_2 = 2.12\text{Å}$

$For\ n = 3$

$r_3 = (0.53 \times 10^{-10}) \times (3^2)$

$r_3 = 4.77 \times 10^{-10}\text{m}$

Thus the radii of the first, second and third stationary orbits of hydrogen atom are 0.53 Å, 2.12 Å and 4.77Å respectively.

2. VELOCITY OF ELECTRON IN BOHR'S STATIONARY ORBIT

From equation below, we have

$$\text{m}\text{v}_\text{n}\text{r}_\text{n} = \frac{\text{nh}}{2\pi}$$

$$v_n = \frac{nh}{2\pi m r_n}$$

Putting the value of r_n into that equation

$$v_n = \frac{nh}{2\pi m r_n} \cdot \frac{\pi m Z e^2}{2\varepsilon_0 n^2 h}$$

$$v_n = \frac{Ze^2}{2\varepsilon_0 nh}$$

It is clear that $v_n \propto \frac{1}{n}$ in other words, electrons move at a lower speed in higher orbits and vice versa.

For hydrogen atom

Z =1

Then

$$v_n = \frac{e^2}{2\varepsilon_0 nh}$$

3. FREQUENCY OF ELECTRON IN STATIONARY ORBIT

The number of revolution completed per second by the electron in a stationary orbit around the nucleus

Velocity of electron in the n^{th} orbit

$$v_n = r_n w_n$$

$$v_n = 2\pi r_n f_n$$

$$f_n = \frac{v_n}{2\pi r_n}$$

$$f_n = \frac{Ze^2}{2\varepsilon_0 nh} \cdot \frac{1}{2\pi r_n}$$

$$f_n = \frac{Ze^2}{4\pi\varepsilon_0 nhr_n}$$

For hydrogen atom

Z = 1

Then,

$$f_n = \frac{Ze^2}{4\pi\varepsilon_0 nhr_n}$$

Frequency of electron in the first orbit of hydrogen atom is n=1, r_1=0.53x10^{-10}m

$$f_n = \frac{e^2}{4\pi\varepsilon_0 nhr_n}$$

$$f_n = \frac{1}{4\pi\varepsilon_0}\cdot\frac{e^2}{1\times h\times r_1}$$

$$f_1 = 9\times 10^9 \times \frac{(1.6\times 10^{-19})^2}{1\times 6.62\times 10^{-34}\times 0.53\times 10^{-10}}$$

$$f_1 = 6.57\times 10^{15}\ r.p.s$$

Electron in first orbit of hydrogen atom will have a frequency of 6.57x 10^{15}revolutions per second.

4. TOTAL ENERGY OF ELECTRON IN STATIONARY ORBIT

The total energy E_n of the electron in the n^{th} orbit is the sum of kinetic and potential energy in the n^{th} orbit.

- The K.E of electron in the n^{th} orbit is

$$K.E_n = {}^1\!/_2\, m{v_n}^2$$

$$K.E_n = \frac{Ze^2}{8\pi\varepsilon_0 r_n}$$

The potential energy of electron in the n^{th} orbit is

$$\mathrm{PE}_n = \frac{1}{8\pi\varepsilon_0}.\frac{(Ze)(-e)}{r_n}$$

$$\mathrm{PE}_n = \frac{-Ze^2}{4\pi\varepsilon_0 r_n}$$

Total energy of electron in the nth orbit is

$$\mathrm{E_n = PE_n + K.E_n}$$

$$E_n = -\frac{Ze^2}{4\pi\varepsilon_0 r_n} + \frac{Ze^2}{8\pi\varepsilon_0 r_n}$$

$$\mathrm{E_n} = -\frac{me^4}{8\varepsilon_0 2h^2}\left(\frac{Z^2}{n^2}\right)$$

But

$$\frac{me^4}{8{\varepsilon_0}^2 h^2} = 2.17 \times 10^{-10}$$

$$E_n = 21.7 \times 10^{-19}\frac{Z^2}{n^2}$$

$$\mathrm{E_n} = -\frac{21.7 \times 10^{-19}}{1.6 \times 10^{-16}}.\frac{Z^2}{n^2}eV$$

$$E_n = -\frac{13.6}{n^2}eV$$

Thus as n increases i.e. electron moves to higher orbit, the total energy of the electron increases i.e. total energy becomes less negative.

For hydrogen atom z=1

$$E_n = \frac{-13.6}{n^2}eV$$

Thus the total energy of electron in a stationary orbit is negative which means that the electron is bound to the nucleus and it is not free to leave the atom.

We can find the total energy of electron in the various orbits of hydrogen atoms as under.

$$E_1 = -\frac{13.6}{1^2} = -13.6eV$$ First orbit n=1

$$E_2 = -\frac{13.6}{2^2} = -13.6eV$$ Second orbit n=2

$$E_3 = -\frac{13.6}{3^2} = -13.6eV$$ third orbit n=3

The total energy of electron increases i.e. becomes less negative as the electron goes to higher orbits

When $n \rightarrow \infty$ $E_n = 0$ and the electron becomes free

Ground state/ normal state

This is the state of atom when the entire electrons in it occupies their lowest energy levels as required by their n and *l* values.

The energy of an atom is least i.e. largest negative value when n=1 i.e. when electron revolves in the first orbit.

The energy of hydrogen atom in the ground state is 13.6eV.

Excited state

This is the state of an atom when electrons in an atom occupy energy levels higher than those permitted by the values of n and *l* values.

At room temperature most of the hydrogen atoms are in the ground state

If hydrogen atom absorbs energy i.e. due to rise in temperature it may be promoted to one of the higher orbits (i.e. n=2, 3, 4…..)

The atom is said to be in the excited state.

WAVE LENGTH OF EMITTED RADIATION.

When an electron jumps from a higher orbit (n_2) to the lower orbit (n_1) the energy difference between the two orbits is released because the energy of electron in the higher orbit is more than in the lower orbit. Consider two orbits having principle quantum numbers n_2 and n_1 where $n_2>n_1$

Then energy of electron in the two orbits is given by

$$E_{n_2} = -\frac{mZ^2e^4}{8\varepsilon_0{}^2n^2{}_2h^2}$$

$$E_{n_1} = -\frac{mZ^2e^4}{8\varepsilon_0{}^2n^2{}_1h^2}$$

As the electron jumps from orbit n_2 to n_1, energy is released in the form of electromagnetic radiation.

$$E_{n2} - E_{n1} = hf$$

where

f= frequency of the emitted radiation

$$\left[\frac{-mZ^2e^4}{8\varepsilon_0{}^2n^2{}_2h^2}\right] - \left[\frac{-mZ^2e^4}{8\varepsilon_0{}^2n^2{}_1h^2}\right] = hf$$

$$\frac{-mZ^2e^4}{8\varepsilon_0{}^2n_1{}^2h^2} - \frac{-mZ^2e^4}{8\varepsilon_0{}^2n_2{}^2h^2} = hf$$

$$f = \frac{mZ^2e^4}{8\varepsilon_0{}^2h^3}\left[\frac{1}{n^2{}_1} - \frac{1}{n^2{}_2}\right]$$

The wavelength of the emitted radiation is given by

c=λf

$\frac{1}{\lambda} = \frac{f}{c}$

$$\frac{1}{\lambda} = \frac{mZ^2e^4}{8\varepsilon_0{}^2h^3c}\left[\frac{1}{n^2{}_1} - \frac{1}{n^2{}_2}\right]$$

This equation gives the wavelength of emitted radiation.

Now,

$\frac{1}{\lambda} = V$ = wave number

$$V = \frac{mZ^2e^4}{8\varepsilon_0{}^2h^3c}\left[\frac{1}{n^2{}_1} - \frac{1}{n^2{}_2}\right]$$

Wave number

These are the number of waves in a unit length.

For hydrogen atom

For hydrogen atom z = 1

$$\frac{1}{\lambda} = \frac{mZ^2e^4}{8\varepsilon_0{}^2h^3c}\left[\frac{1}{n^2{}_1} - \frac{1}{n^2{}_2}\right]$$

This gives the mathematical formula for the wavelength of radiation emitted by hydrogen atom when electron jumps from outer orbit to inner orbit.

$$\frac{1}{\lambda} = R_H\left[\frac{1}{n^2{}_1} - \frac{1}{n^2{}_2}\right]$$

where

R_H is Rydberg constant. The value of R_H can be calculated as the value of e, m, h and c are known

$$R_H = \frac{mZ^2e^4}{8{\varepsilon_0}^2ch^3}$$

$$R_H = \frac{(9.1 \times 10^{-31}) \times (1.6 \times 10^{-19})^4}{8 \times (8.854 \times 10^{-12})^2 \times (3 \times 10^{-12}) \times (6.62 \times 10^{-24})^3}$$

$$R_H = 1.097m^{-1}$$

HOW TO CALCULATE THE RYDBERG CONSTANT USING CALCULATOR

From

$$R_H = \frac{mZ^2e^4}{8{\varepsilon_0}^2ch^3}$$

$$m = 9.1 \times 10^{-31}\ kg$$

$$m = 1.6 \times 10^{-19}\ c$$

$$\varepsilon_0 = 8.854 \times 10^{-12} Fm^{-1}$$

$$c = 3 \times 10^8 ms^{-1}$$

$$h = 6.62 \times 10^{-34} Js$$

$$R_H = \frac{(9.1 \times 10^{-31}) \times (1.6 \times 10^{-19})^4}{8 \times (8.854 \times 10^{-12})^2 \times (3 \times 10^8) \times (6.62 \times 10^{-24})^3}$$

$$R_H = \frac{9.1 \times (1.6)^4}{8 \times (8.854)^2 \times 3 \times (6.627)^3} \cdot \frac{10^{-31} \times (10^{-19})^4}{(10^{-12})^2 \times 10^8 \times (10^{-34})^3}$$

$$R_H = 1.093 \times 10^{-4} \times \frac{10^{-31} \times 10^{-76}}{10^{-34} \times 10^8 \times 10^{-102}}$$

$$R_H = 1.093 \times 10^{-4} \times \frac{10^{-31} \times 10^{-76}}{10^{-118}}$$

$$R_H = 1.093 \times 10^{-4} \times \frac{10^{-107}}{10^{-118}}$$

$$R_H = 1.093 \times 10^{-4} \times 10^{11}$$

$$R_H = 1.093 \times 10^7 m^{-1}$$

Clearly, wavelength/frequency of radiation emitted from the excited atom is not continuous. They have definite value depending upon the values of n_1, and n_2

SPECTRAL SERIES OF HYDROGEN ATOM

Bohr gave a mathematical explanation for the spectrum of hydrogen atom.

The whole hydrogen spectrum can be divided into district groups of lines each group of lines is called spectral series.

The wavelength of the lines in each group can be calculated from Bohr's formula

$$\frac{1}{\lambda} = R_H\left[\frac{1}{n_1^2} - \frac{1}{n_2^2}\right]$$

The following are spectral series of hydrogen atom

i) Lyman series

ii) Balmer series

iii) Paschen series

iv) Bracket series

v) Pfund series

i) Lyman series

The Lyman series is obtained when electron jump to first orbit n_1=1 from outer orbits (n_2=2, 3, 4…)

Therefore the formula for calculating the wavelength of the lines in this series is,

$$\frac{1}{\lambda} = R_H\left[\frac{1}{n^2{}_1} - \frac{1}{n^2{}_2}\right]$$

where

$$n_1 = 2, 3, 4 \ldots \ldots \ldots$$

This series lies in the ultraviolet region which is the invisible region.

ii) Balmer series

The Balmer series is obtained when electrons jump to second orbit $n_1 = 2$ from

outer orbit $(n_1 = 3, 4 \ldots.)$

Therefore the formula for calculating the wavelength of the lines in this series is
outer orbit $(n_1 = 3, 4 \ldots.)$

$$\frac{1}{\lambda} = R_H\left[\frac{1}{n^2{}_1} - \frac{1}{n^2{}_2}\right]$$

$$\frac{1}{\lambda} = R_H\left[\frac{1}{2^2} - \frac{1}{n^2{}_2}\right]$$

where

$n_2 = 3, 4, 5 \ldots \ldots$

This series lies in the visible spectrum and was found first of all in the hydrogen series

iii) Paschen series

The paschen series is obtained when electrons jump to third orbit $n_1 = 3$ from outer orbit $(n_2 = 4, 5, 6 \ldots \ldots)$

Therefore the formula for calculating the wavelength of the lines in this series is

$$\frac{1}{\lambda} = R_H\left[\frac{1}{3^2} - \frac{1}{n^2{}_2}\right]$$

where

$n_2 = 4, 5, 6 \ldots \ldots$

This series lies in the infrared region.

iv) **Brackett series**

The blackett series is obtained when electrons jump to fourth orbit $n_1 = 4$ from outer orbit $(n_2 = 5, 6, 7 \ldots \ldots)$

Therefore the formula for calculated the wavelength of the lines in this series is

$$\frac{1}{\lambda} = R_H\left[\frac{1}{4^2} - \frac{1}{n^2{}_2}\right]$$

This series lies in the infrared region.

v) **Pfund series**

The Pfunds series is obtained when electrons jump to fifth orbit $n_1 = 5$ from outer orbits ($n_2 = 6, 7, 8\ldots\ldots$)

Therefore the formula for calculating the wavelength of the lines in this series is

$$\frac{1}{\lambda} = R_H\left[\frac{1}{5^2} - \frac{1}{n^2{}_2}\right]$$

where

($n_2 = 6, 7, 8\ldots\ldots$)

This series also lies in the infrared region

ENERGY LEVEL DIAGRAM

Energy level diagram is a diagram in which the total energies of electron in different stationary orbit of an atom represented by parallel horizontal lines drawn according to some suitable energy scale

In order to draw energy level diagram of an atom we must know the total energy of electron in different stationary orbits.

The total energy of an electron in the n^{th} orbit of hydrogen atom is given by

$$E_n = \frac{-13.6}{n^2}\,eV$$

By putting value of n=1, 2, 3….. we can find the total energy of electron in various stationary orbits of hydrogen atom as

$$E_1 = -\frac{13.6}{1^2} = -13.6eV$$

$$E_2 = -\frac{13.6}{2^2} = -3.4eV$$

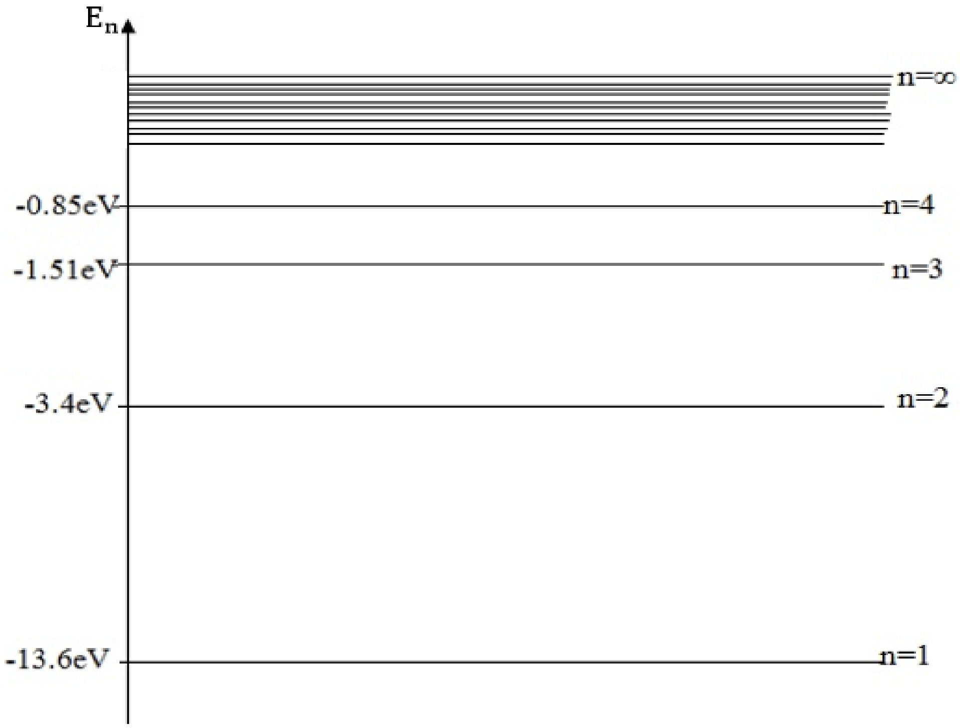
$$E_3 = -\frac{13.6}{3^2} = -1.51eV$$

$$E_4 = -\frac{13.6}{4^2} = -0.85eV$$

Similarly we can find the total energy of electron in the higher orbits

The table below gives the total energy of electron of hydrogen atom in different stationary orbits.

Orbits(n)	Total electron of electron
$n = 1$	$-13.6eV$
$n = 2$	$-3.4eV$
$n = 3$	$-1.51eV$
$n = 4$	$-0.85eV$
$n = 5$	$-0.54eV$
$n = 6$	$-0.38eV$
$-$	$-$
$-$	$-$
$n = \infty$	$0eV$



T

he energy level diagram of hydrogen atom is shown below

Total energy of electron in a stationary orbit is represented by a horizontal line drawn to some suitable energy scale.

(i) The hydrogen atom has only one electron and this normally occupies the lowest level and has energy of -13.6eV

When the electron is in this level the atom is said to be in the ground state.At room temperature nearly all the atoms of hydrogen are in ground.

(ii) If hydrogen atom absorbs energy (due to rise in temperature)the electron may be promoted into one of the higher energy levels

The atom is now said to be in an excited state.Thus when the electron occupies other than the lowest energy level the atom is said to be in the excited state.

(iii) Once in an excited state the atom is unstable after a short time interval the electron falls back into the lowest state so that the atom is again in the ground state.

The energy that was originally impacted is emitted as electromagnetic waves.

(iv) The total energy of electron for (n=∞) it becomes free of atom.

The minimum energy required to free the electron from the ground state of an atom is called ionization energy

For hydrogen atom ionization energy is +13. 6eV

(v) The difference between the adjacent energy goes on decreasing as the value of n increases.

So much so that when n>10 the energy difference is almost zero this is show by closeness of energy level lines at higher levels.

(vi) Note that region is labeled continuous at energy above zero n=∞ level, the electron is free from the atom and is at rest

Higher energy represents the translation kinetic energy of the free electron

This energy is not quantized and so all energies above n =∞ are allowed

IMPORTANT TERMS

It is desirable to discuss some important terms much used in the study of structure of atom.

(i) EXCITATION ENERGY

Excitation energy is the minimum energy required to excite an atom in the ground state to one of the higher stationary state.

Hydrogen atoms are usually in their lowest energy state where n=1

In this state (ground state) they are said to be unexcited.

However if you bombard the atoms with particles such as electron or proto collision can excite them

In other words a collision may give an atom enough energy to change it from ground state to some higher stationary state.Consider the case of hydrogen atom we know that E_1= -13.6eV (ground state E_2= -3.4eV (first excited state) E=1.51eV (second excited state) and E_∞=0

In order to lift an electron from ground state n =1 to the first excited state n=2 energy required is E

E = E_2- E_1

E= -3.4 - (- 13.6)

E = 10.2eV

Therefore the bombarding particle must provide an energy of 10.2eV to excite the atom from n =1 state to n=2 state

Similarly to excite the atom from n=1 state to n=3 state energy required is

$$E = E_2 - E_1$$

E = - 1.51 – (-13.6)

E = 12.1eV

We say that first and second excitation energies of hydrogen are 10.2eV and 12.1eV respectively

(ii) EXCITATION POTENTIAL

Excitation potential is the minimum accelerating potential which provide an electron energy sufficient to jump from the ground state n=1 to one of the outer orbits

In this case of hydrogen atom $E_2 = -3.4eV, E_3 = -1.15eV,\ E_\infty = 0$

Energy required to lift a electron from ground state n=1 to n=2 state is

$$\mathrm{E} = \mathrm{E}_2 - \mathrm{E}_1$$

$$\mathrm{E} = -3.4 - (-13.6)$$

$$E = 10.2eV$$

$$\text{Excitation potential} = \frac{10.2eV}{e} = 10.2V$$

Hence excitation potential for the first excited state of hydrogen is 10. 2V

Similarly energy required to lift an electron from ground state n=1 to n=2

$$E = E_2 - E_1$$

$$E = -1.51 - (-13.6)$$

$$E = 12.1eV$$

$$\text{Excitation potential} = \frac{12.1eV}{e} = 12.1V$$

The value of excitation potential depend upon the state to which the atom is excited to which the atom is excited from the ground state

(iii) IONIZATION ENERGY

Ionization energy is the minimum energy needed to ionized an atom

Consider the case of hydrogen atom it has only one electron and this normally occupies the ground state.

The energy of the electron for n=∞ state is zero and if the electron is lifted to this level (n=∞) it becomes free of hydrogen atom i.e. hydrogen atom is ionized

$$\text{Ionization energy of hydrogen atom} = E_\infty - E_1$$

$$= 0 - (-13.6)$$

$$= 0 + 13.6$$

$$= 13.6eV$$

(iv) IONIZATION POTENTIAL

Ionization potential is the minimum accelerating potential which would provide electron energy sufficient to just remove the electron from the atom.

$$\text{Ionization potential} = \frac{E_\infty - E_1}{e}$$

$$= \frac{0 - (-13.6)}{e}$$

$$= \frac{13.6eV}{e}$$

$$\text{Hence ionization potential of h atom} = 13.6V$$

The ionization potential of one electron atom or ion is given by

$$\text{Ionization potential} = \frac{13.6Z^2}{n^2}$$

(v) QUANTIZATION OF ENERGY

Quantization of energy is the existence of energy radiated by atoms in a specific amount which is are integral multiples of a constant (hf).

SUCCESS OF BOHR'S THEORY

The success of bohr's theory is not to be attributed so much to the mechanical picture of atom he proposed but rather to the development of mathematical explanation that agrees exactly with experimental observations. Bohr's theory achieved the following successes.

i) MADE ATOM STABLE

Bohr's theory made the atom stable according to this theory an electron moving in the formatted (quantum) orbits cannot lose energy even though under constant acceleration. This provided stability to the atom.

ii) INTRODUCED QUANTUM MECHANICS

Bohr's theory introduced quantum mechanics in the realm of atom for the first time

Bohr's explained that sub- atomic particles e.g. electrons are governed by the laws of quantum mechanics and not by classical laws of electron hydrogen as assumed by Rutherford

This completely changed our thinking and was the major step towards the discovery of the rudiment laws of the atomic world

iii) GAVE MATHEMATICAL EXPLANATION OF HYDROGEN SERIES

The hydrogen series found by various scientists were based on empirical relation but had no mathematical explanation

However these relations were easy derived by applying Bohr Theory

Further the size of hydrogen atom as calculated from this theory agreed very closely with the experimental value.

LIMITATIONS OF BOHR'S THEORY

Bohr's simple theory of circular orbits inspire of its many successes was found inadequate to explain many phenomena observed experimentally.

This theory suffered from the following drawbacks.

(i) It could not explain the difference in the intensities of emitted radiations.

(ii) It is silent about the wave properties of electron

(iii) It could not explain experimentally observed phenomena such as Zeeman Effect, Stack effect etc.

(iv) Bohr's model does not explain why the orbit are circular while elliptical path is also possible

(v) It could only partially explain hydrogen atom. For example this theory does not explain the fine structure of spectral lines in the hydrogen atom

WORKED EXAMPLES

1. 1. Find the radius of the first orbit of hydrogen atom. What will be the velocity of electron in the first orbit? Hence find the size of hydrogen atom

Solution

The radius of nth orbit of it atom is given by

$$r_n = \frac{\varepsilon_0 n^2 h^2}{\pi m e^2}$$

Radius of first orbit of it atom n=1

$$r_n = \frac{\varepsilon_0 n^2 h^2}{\pi m e^2}$$

$$n = 1$$

$$r_1 = \frac{\varepsilon_0 1^2 h^2}{\pi m e^2}$$

$$r_1 = (\frac{8.854 \times 10^{-12} \times (6.62 \times 10^{-34})^2}{3.14 \times (9 \times 10^{-31}) \times (1.6 \times 10^{-19})})$$

$$\frac{8.854 \times 10^{-12} \times (6.62 \times 10^{-34})^2}{3.14 \times (9 \times 10^{-31}) \times (1.6 \times 10^{-19})})$$

$$r_1 = 0.53 \times 10^{-10} \text{ m}$$

$$r_1 = 0.53\text{Å}$$

Velocity of electron in the n^{th} orbit of hydrogen atom is given by

$$v_n = \frac{e^2}{2\varepsilon_o n h}$$

Velocity of electron in the first orbit of hydrogen atom is given by

$$v_1 = \frac{e^2}{2\varepsilon_0 h}$$

$$v_1 = \frac{(1.6 \times 10^{-19})}{2 \times (8.854 \times 10^{-12} \times (6.62 \times 10^{-34}))} = 2.18 \times 10^6 ms^{-1}$$

Since there is one electron in hydrogen atom the size hydrogen atom is equal to double the radius of the first orbit

Size of the atom

$= 2r_1$

$= 2 \text{ x } 0.53\text{Å}$

Size of an atom =1.06Å

2 2. (a) The hydrogen atom is stable in the ground, state why?

(b) The ionization energy of hydrogen is 13. 6eV what does it mean?

(c) Calculate the wavelength of second line of Lyman series

Solution

If the hydrogen atom is in the ground state (n=1) there is no state of lower energy to which a down ward transition can occur thus a hydrogen atom in the ground state is stable

a) It means that energy required to remove the single electron from the lowest energy state of hydrogen atom to becomes free electron is 13.6eV

b) Second line of Lyman series is obtained when electron jumps from third orbit n_2 =3 to the first orbit n=1

According to Bohr's theory the wavelength of emitted radiation is given by

$$\frac{1}{\lambda} = R_H \left[\frac{1}{n_1^2} - \frac{1}{n_2^2}\right]$$

$$\frac{1}{\lambda} = R_H \left[\frac{1}{1} - \frac{1}{3^2}\right]$$

$$\frac{1}{\lambda} = R_H \times \frac{8}{9}$$

$$\lambda = \frac{9}{8R_H} = \frac{9}{8 \times 1.097 \times 10^7} = 1.0255 \times 10^{-7}\text{m}$$

3 3.(a) What is the meaning of negative energy of orbiting electron?

(b) What would happen if the electron in atom were stationary?

(c)Show that he paschen series of spectral lines lies entirely in the infrared part of the spectrum given $R_H = 1.097 \times 10^{-7}\text{m}^{-1}$

Solution

a) The negative total energy means that it is bound to the nucleus. If it acquires enough energy from some external source (a collision for example) to make its total energy zero the electron is no longer bound it is free.

b) If the electrons were stationary they would fall into the nucleus due to electrostatic force of attraction so atom would be unstable i.e. it would not exist

c) For Paschen series we have longest wavelength line n_2=4

This is a wavelength in the infrared part. Other lines in this series have shorter wavelength bad approach series limit of wavelength to given by This wavelength is also in the hydrogen part. This the range or centre series (820.4nm to 1875nm) is the infrared

4. a) If an electron jumps from first orbit to third orbit will it absorb energy?
b) Name the series of hydrogen spectrum lying in the infrared region
c) Calculate the shortest wavelength of the Balmer series
d) What is the energy possessed by an electron for n=?

Solution

a) Yes it is because the energy level of third orbit is more than that of the first orbit

b) * Paschen series

* Bracket series

* P fund series

Solution

In Balmer series the radiation of shortest wavelength (i.e. of highest of highest energy) is emitted when electron jumps from infinity orbit $n_2=\infty$ to the second orbit n_1=2 of hydrogen atom.

5 5. a) The ionization potential of hydrogen is 13.6V what does it mean?

b) Find the longest wavelength in Lyman series

c) How much is the ionization potential of hydrogen atom?

d) The energy of the hydrogen atom in the ground state is 13.6eV. Determine the energies of those energy levels whose quantum numbers are 2 and 3.

Solution

a) The ionization energy of hydrogen is 13.6eV. Therefore, if an electron which has been accelerated from rest through a p.d of 13.6V collides with a hydrogen atom it has exactly the right amount of energy to produce ionization.

This is a common method of producing ionization and therefore the term ionization potential is often used.

b) **Solution**

In Lyman series the radiation of longest wavelength (i.e. lowest energy) is emitted when electron jumps from second orbit n_2=2 to first orbit n=1 of hydrogen atom

c) The energy of hydrogen atom in the ground state is – 13.6eV. therefore its ionization energy is 13.6eV and ionization potential =13.6V

d) **Solution**

The energy of an electron in the n^{th} orbit of hydrogen atom is given by

$$E_n = -\frac{13.6eV}{n^2}$$

$$For\ n = 2$$

$$E_2 = -\frac{13.6}{2^2}eV$$

$$E_2 = -3.4eV$$

$$For\ n = 3$$

$$E_3 = -\frac{13.6}{3^2}eV$$

$$E_3 = -1.15eV$$

6

6. a) Name the series of hydrogen spectrum lying in the

i) Visible region

ii) Utraviolet region of electromagnetic spectrum

b) Write the empirical relation for Paschen series lines of hydrogen spectrum

c) What are the values of first and second excitation potential of hydrogen atom?

d) Calculate the radii and the energy of three lowest energy allowed orbits for the electron in Lithium ion. What is the energy of a photon that when absorbed causes an electron in Lithium ion to be excited from n=1 to n=3 state?

Solution

a) i) Balmer series

ii) Lyman series

b) The wavelength of the spectral lines in paschen series are given by

c) Excitation energy for first excited state $= -3.4 - (-13.6)$

$= 10.2\text{eV}$

For second excited state

= - (1. 51 - (-13.6)

= 12.1eV

Solution

d) (I) for a single electron atom or ion the radius of the n^{th} orbit is given

$$r_n = (\frac{\varepsilon_0 h^2}{\pi m e^2})\frac{n^2}{2}$$

$$But \frac{\varepsilon_0 h^2}{\pi m e^2} = 0.53 \times 10^{-10}$$

$$Then\, r_n = (0.53 \times 10^{-10})\frac{n^2}{z}$$

Thus the radiin of the $n = 1, 2$ and 3 orbits are

$$n = 1$$

$$r_1 = \frac{(0.53 \times 10^{-10})}{1} \times \frac{1^2}{3}$$

$$r_1 = 0.18 \times 10^{-10}\text{m}$$

$$r_1 = 0.18\ \text{Å}$$

$$n = 2$$

$$r_2 = \frac{(0.53 \times 10^{-10})}{1} \times \frac{2^2}{3}$$

$$r_2 = 0.71 \times 10^{-10}$$

$$r_2 = 0.71\text{Å}$$

$$n = 3$$

$$r_3 = \frac{(0.53 \times 10^{-10})}{1} \times \frac{3^2}{3}$$

$$r_3 = 1.59 \times 10^{-10}$$

$$r_3 = 1.59\text{Å}$$

For a single electron atom or ion the energy of electron in the nth orbit is given by

$n = 1$

$$E_1 = -\frac{13.6 \times 3^2}{1^2}$$

$E_1 = -122.4eV$

$n = 2$

$$E_2 = \frac{-13.6 \times 3^2}{2^2}$$

$E_2 = -30.6eV$

$n = 3$

$$E_3 = \frac{-13.6 \times 3^2}{3^2}$$

$E_3 = -13.6eV$

Thus the energy of n=1, 2 and 3 orbits. The photo energy must be equal to the energy needed to excite the electron

$E_3 – E_1$ =hf

(-13.6) – (-122.4) =photon's energy

Photon's energy = 108.8eV

7. The ionization energy of hydrogen like atom is 4rydbergs

(a) What is the wavelength of radiation emitted when electron jumps from first excited state to the ground state?

(b)What is the radius of the first orbit for this atom?

Given Bohr's radius of hydrogen atom is $5 \times 10m$ and 1rydberg $= 2.2 \times 10^{18} j$

(c)Which state of triply ionized beryllium (Be^{3+}) has the same orbital radius as that of the ground state

(d) According to Bohr's theory what is the angular momentum of a electron in the third orbit

Solution

The energy electron in the n^{th} orbit of hydrogen like atom is

$$E_n = -\frac{mZ^2e^4}{8\varepsilon_0(2n^2h^2)}$$

$$But\ R_H = \frac{me^4}{8\varepsilon_0 ch^2}$$

$$Then\ E_n = \frac{me^4}{8{\varepsilon_0}^2h^2}\cdot\frac{Z^2}{n^2}$$

$$E_n = \frac{-Z^2R_Hhc}{n^2}$$

$$\text{Ionization energy} = E_\infty - E_1 = Z^2R_Hhc$$

$$\text{Then }\ Z^2R_Hhc = 4rydbergs$$

$$Z^2 = 4rydbergs/R_Hhc$$

$$Z^2 = \frac{4rybergs}{1rybergs}$$

$$Z^2 = 4$$

$$Z = 2$$

The energy required to excite the electron from n=1 level to n=2

$$E = E_2 - E_1$$

$$E = \frac{-Z^2R_Hhc}{2^2} - \left(\frac{-Z^2R_Hhc}{1^2}\right)$$

$$E = Z^2R_Hhc\left(1 - \frac{1}{4}\right)$$

$$E = \frac{3}{4}Z^2R_Hhc$$

$$E = \frac{3}{4} \times 4rydbergs$$

$$E = 3rydbergs$$

If is the wavelength of the emitted radiations then, Radius of first orbit for this atom

$$r_1 = \frac{5 \times 10^{-10}}{Z}$$

$$r_1 = 2.5 \times 10^{-11}\ \mathrm{m}$$

Solution

(b) Radius of nth orbit

$$r_n \propto \frac{n^2}{Z}$$

Let

$$r_n \mathrm{Be}^{3+} = \mathrm{r}, (\mathrm{H})$$

$$\left[\frac{n^2}{\mathrm{F}}\right]\mathrm{Be}^{3+} = \left[\frac{n^2}{Z}\right]\mathrm{H}$$

$$\frac{n^2}{4} = \frac{1^2}{1}$$

$$n = 2$$

Therefore second orbit of Be^{3+} ions have the same radius as that of the ground state of hydrogen atom.

(c) **Solution**

Angular momentum L of an electron in nth orbit is

L = n $\frac{h}{2\pi}$

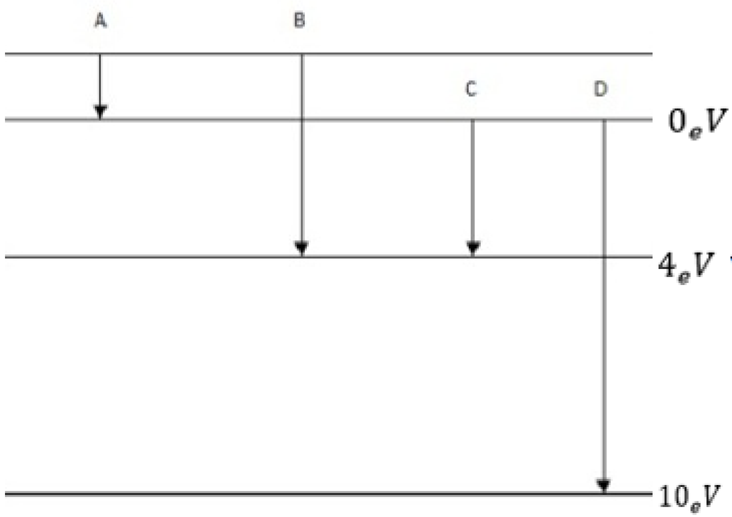
Here n=3

Then

L= 3 $\frac{h}{2\pi}$

L= $\frac{3h}{2\pi}$

7 8. The energy levels of an atom are shown in figure below.



(a)Which one of these transitions will result in the emission of photon of wavelength 275nm?

(b) An electron orbiting in hydrogen atom has energy level of 3.4eV what will be its angular momentum
(c) The total energy of an electron in the first excited state of hydrogen atom is about
– 3. 4eV what is the wavelength?
solution
(a) Energy of emitted photon E

$$E = \frac{12400}{275} eV$$

$$E_n = 4.5eV$$

Therefore photon of wavelength 275nm will be emitted for transition B

Solution

$$(b) E_n = \frac{-13.6}{n^2} eV$$

$$-3.4 = \frac{-13.6}{n^2}$$

$$n^2 = \frac{-13.6}{3.4} = 4$$

$$n = 2$$

(c) Angular momentum

$$L = n\frac{h}{2\pi} = 2 \times \frac{6.6 \times 10^{-34}}{2\pi}$$

$$L = 2.1 \times 10^{-34} Js$$

(d) K.E of electron = -(total energy of electron)

K.E of electron =3.4eV

ii) P.E of electron = 2xtotal energy

P.E of electron = -6.8eV

10. (a) How many lines can be drawn the energy level diagram of hydrogen atom?

(b) Use Bohr's model to determine the ionization energy of the He ion also calculate the minimum wavelength a photo must have to cause ionization

(c)Determine the speed of electron in the $n = 3$

orbit of He^+ ion , is non relativistics approximation valid?

(d) i) In neon atom the energies of the 3s and 3p states are respectively 16.70eV and 18.70eV. What wavelength corresponds to 3p -3s transitions in neon atom?

ii) The wavelength of the first member of the Balmer series in hydrogen spectrum is 6563Å. Calculate the wavelength of first member of Lyman series in the same spectrum

PLANCK'S QUANTUM THEORY OF BLACK BODY RADIATION

The findings in the black body radiation led Max plank in 1901 to postulate that radiant energy is quantized i.e. it is radiated in form of energy packets.

BASIC POSTULATES OF THE PLANK'S THEORY

1. Any radiation is associated with energy.
2. Radiant energy is emitted or absorbed in small packets known as quanta.
3. The energy associated with a quantum is proportional to the frequency f of the radiation

$$E \propto f$$

$$E = hf$$

$$h = Planck's constant = 6.626 \times 10^{-34} Js$$

4. The energy is absorbed or emitted only in whole number of quanta

Black Body Radiation

$$E = nhf$$

$$where\ n = 1, 2, 3 \ldots \ldots \ldots \ldots$$

A blackbody is a substance that absorbs all light fall on it and does not reflect any light.

It is not easy to get a black body however a sealed metal box with a very small hole on it is very close to a black body.

From law of physics it follows that a good absorber of radiation also is a good radiator. A black body is supposed to be the best radiator.

When a black body is heated it emits light. The colour of light emitted changes from red to yellow then to white as the temperature is increased.

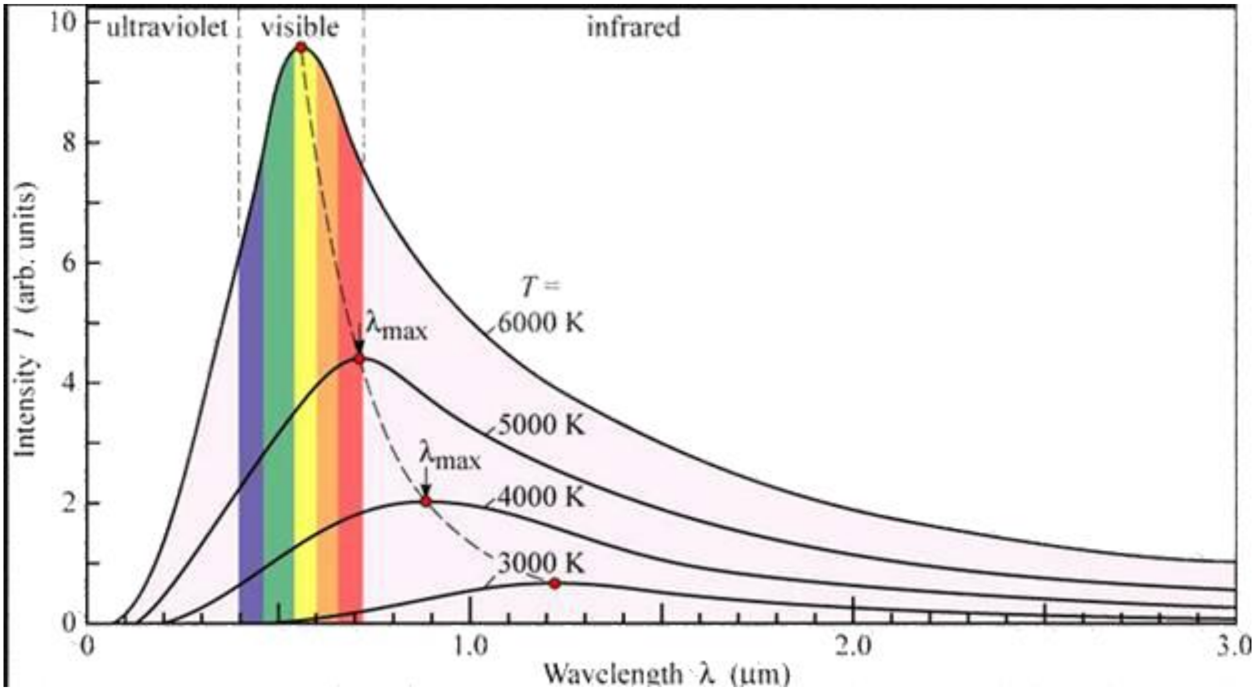
The change in colour with temperature shows that the frequency changes with temperature.

This in contradiction with the classical wave theory since in the classical wave theory energy is uniformly distributed over the wave form when heating the black body the colour of radiation should stay the same

Only the intensity is supposed to increase with temperature.

DISTRIBUTION OF ENERGY IN THE SPECTRUM OF A BLACK BODY

Lamer and Pringshein investigated the distribution of energy amongst the different wavelength of a thermal spectrum of a back body radiation



The results obtained by Lamer and Pringshein are shown in figure below

Results:

1. At a given temperature the energy is not uniformly distributed in the radiation spectrum of a hot body

2. At a given temperature the intensity of radiations increases with increased in wavelength and at a particular wavelength λ its value is maximum with further increase in wavelength the intensity of heat radiations decreased

3. With increase in temperature wave length increases, wavelength emission of energy takes place.

   The points on the dotted line represent wavelength at various temperatures

4. For all wavelength an increase in temperature causes an increase in the energy emission The area under each curve represents the total energy emitted for the complete spectrum at a particular temperature

5. This area increases in temperature of the body. It is found that the area is directly proportional to fourth power of the temperature of the body

   $E \propto T^4$

This represents Stefan's Boltzmann's law.

Plank's constant

Plank's constant is a fundamental constant equal to the ratio of the quantum energy to the frequency of the radiation.

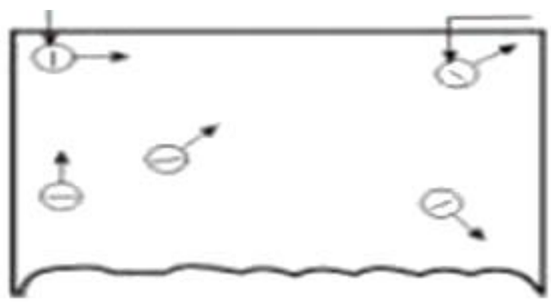
$$H = \frac{E}{f}$$

ELECTRON EMISSION

This is the liberation of electron from the surface of a substance.

For electron emission metals are used because they have many free electrons.

If a piece of metal is investigated at room temperature the random motion of free electrons is as shown in figure below



However these electrons are free only to the extent that they may transfer from one atom to another within the metal but they cannot leave the metal surface to provide electron emission.

It is because the free electron that start at the surface of metal find behind them positive nuclei pulling them back and none pulling forward.

This at the surface of a metal a free electron encounters force that prevents it to leave the metal

In other words the metallic surface offers a barrier to free electrons and is known as surface barriers

However if sufficient external energy is given to the free electron its kinetic energy is increased and thus electron will cross over the surface barrier to leave the metal.

WORK FUNCTION OF THE METAL

This is the additional energy required by an electron to overcome the surface barrier of the metal.

Or

This is the minimum energy required by an electron to just escape from the metal surface.

It is denoted by W_0 Or $\emptyset_0$

$$W_0 = hf_0$$

The work function depends on

i) Nature of metal

ii) Conditions of the metal surface

It is measured by a smaller unit of energy called electron volt. (eV) because this is the conventional unit of energy i.e. joule is very large for computations in atomic and nuclear physics.

Electron volt.

One electron volt is the amount of energy acquired by an electron when it is accelerated through a potential difference of IV

Since potential difference V

$$V = \frac{Workdone(W)}{Charge(Q)}$$

Work done = QV

For an electron

$$Workdone\ eV$$

$$1eV = 1.6 \times 10^{-19}J$$

$$1eV = 1.6 \times 10^{-19}C \times 1V$$

The electron volt is the kinetic energy gained by an electron being accelerated by a potential difference of one volt.

The work function of pure metal varies roughly from $2\ e$V to $6\ e$V as shown in table below

Metal	Work Function W_o (eV)
Cs	2.14
K	2.30
Na	2.75
Ca	3.20
Mo	4.17
Pb	4.25
Al	4.28
Hg	4.49
Cu	4.65
Ag	4.70
Ni	5.15
Pt	5.65

It is clear from the table above that the work function of platinum is the highest while it is lowest for Cesium.

It is desirable that metal used for electron emission should have low work function so that a small amount of energy is required to cause emission of electrons

PHOTOELECTRIC EFFECT

Photoelctric effect is the phenomenon of emission of electron from a metallic surface when radiation of suitable frequency falls on it or is the phenomenon where electromagnetic radiation of certain frequency when incident on certain material liberates electron from the surface of the material.

Photo emission

Photo emission is the emission of electron by electromagnetic radiation.

Photo electrons

These are emitted or rejected electrons from the surface of the cathode.

Photo electric effect is a general phenomenon exhibited by all substances but is most easily observed with metals.

When radiation of suitable frequency the threshold frequency is incident on a metallic surface electrons are emitted from the metal surface.

The threshold frequency is different for metals.

Certain alkali, metals e.g. sodium potassium, calcium show photo electric effect when visible light falls on them.

However, metal like zinc, calcium, magnesium etc show photo electric effect to ultra violet light.

Threshold Frequency

The threshold frequency is the minimum frequency of the incident radiation which is just sufficient to eject photo electron from surface of a metal Or is the minimum frequency of radiation below which no photo electron emission occurs.

It is denoted by f_0

Illuminating a metal surface with light of frequency less than f_o will not cause ejection of photo electrons, no matter now great is the intensity of radiation.

But illumination with a frequency greater than f_o causes emission of photo electrons even if the radiation intensity is very small

Threshold wavelength

The threshold wavelength is the maximum wavelength of the incident radiation at which photo electric emission occurs.

It is denoted by λ_0

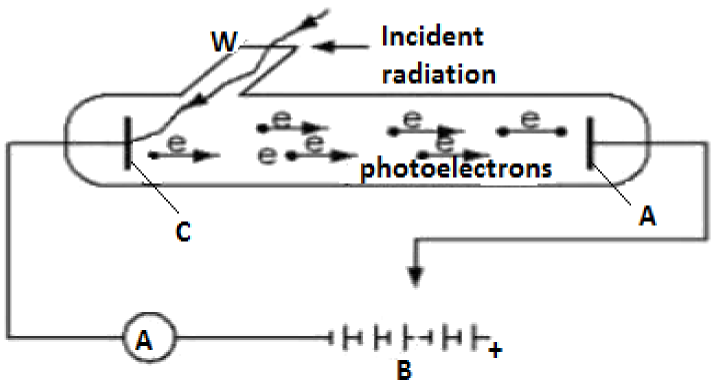
It is the wavelength corresponding to threshold frequency

If the wavelength of the incident radiation is greater than threshold wavelength then there will be no photo electric emission.

Photoelectric current is the photo-electrons emitted per second

EXPERIMENTAL STUDY OF PHOTO ELECTRIC EFFECT

Figure below shows the experimental set up for studying the photoelectric effect.



The arrangement consists of an evacuated glass or quartz tube inclusive a photosensitive cathode C and metallic A.

A transparent window W is sealed onto the glass tube which can be covered with different filters to obtain the desired frequency.

The anode and cathode are connected to a battery through a potential divided by which potential difference between anode and cathode can be changed.

The reversing switch RS tends to make anode positive or negative with respect to cathode.

The P.D between anode and cathode is measured by the voltmeter V while photoelectric current is indicated by the micro ammeter.

1.

1.Effect of intensity of light on photo electric current

The anode A is maintained at positive potential with respect to cathode C and a radiation of suitable frequency (above threshold frequency) is incident on cathode C.

As a result photo electric current is set up.

Keeping the frequency of incident radiation and accelerating potential fixed the intensity of the incident radiation is changed in steps.

For each value of intensity of radiations the corresponding value of photo electron current is noted.

If we plot a graph between intensity of radiation and photoelectric current it is found to be a straight line passing through the origin O as shown in figure below

This shows that photoelectric current is directly proportional to the intensity of incident radiation

The intensity of radiation can be changed by changing the distance between cathode C and the source of radiation.

Effect of potential of anode with respect to cathode on photoelectric current

We keep the anode at some positive accelerating potential with respect to cathode C and illuminate the cathode with radiation of fixed frequency f above threshold frequency and fixed intensity I

If we increase the positive potential on anode gradually, it is found that photo electric current also increases a stage comes when the photo electric current becomes maximum.

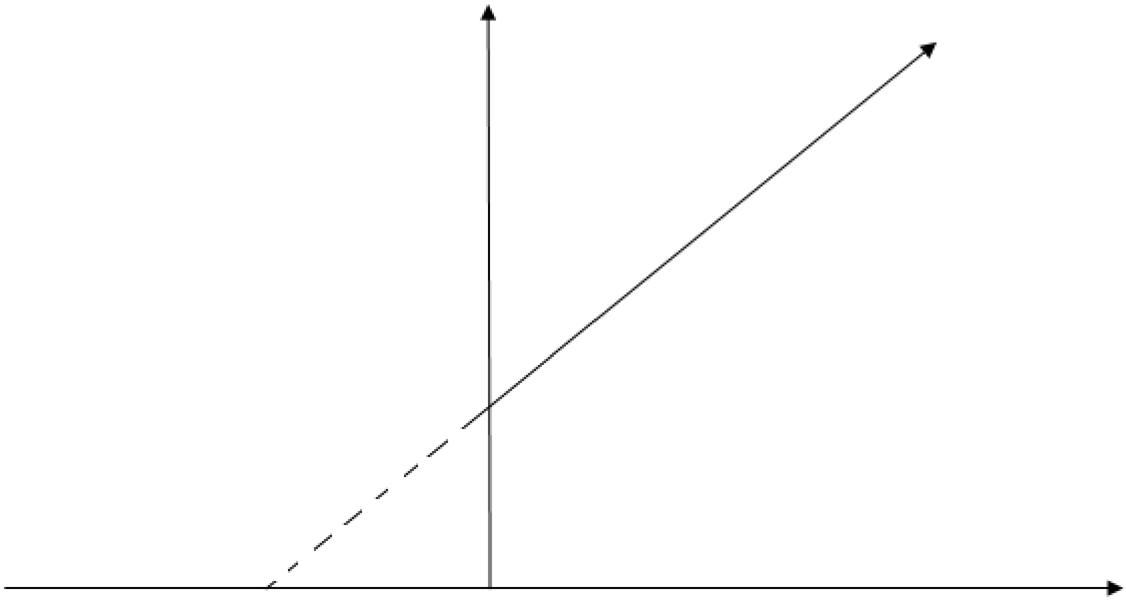
If we increase the positive potential on anode further the photo electric current does not increase.

This maximum value of photo electric current is called saturation current and corresponds to the photo electrons emitted by the cathode reach the anode A

We can now repeat this experiment with incident radiation of

the same frequency but of higher intensity I_2 *and* $I_3 I_3 > I_2 > I_1$

Now saturation current is of higher value as shown in figure below



This is expected because the greater the intensity of incident radiation the greater is the photo electric current

Stopping potential

Stopping potential is the minimum retarding potential at which photoelectric current becomes zero
Or is the potential difference when no electrons are able to reach the anode.

It is also known as stopping voltage or cut-off potential.

It is denoted by V_0 or V_s

Stopping potential is a measure of the maximum kinetic energy of the photo electrons.

Since potential difference V

$$V = \frac{Working\ done(W)}{Charge(Q)}$$

For stopping Voltage V_o

$W = QVo$

For an electron

$W = QV_o \quad \therefore Q = e$

$W = eV_o$

$\frac{1}{2}mv^2 = \mathrm{eV_o}$

At V_0, even the photo electrons having maximum kinetic energy $K.E_{max}$ (i.e. fastest photo electrons) cannot reach the anode A.

Therefore, the stopping potential V_0 is a measure of the maximum kinetic energy $K.E_{max}$ of the photo electrons.

eV_0 is the work done by the retarding force to stop the photo electron with maximum kinetic energy and is therefore equal to $K.E_{max}$.

At V_0, it is found that the photoelectric current cannot be obtained even if we increase the intensity of radiation. It is same for different intensities I_1, I_2 and I_3 of incident radiation.

3. Effect of frequency of incident radiation on stopping potential.

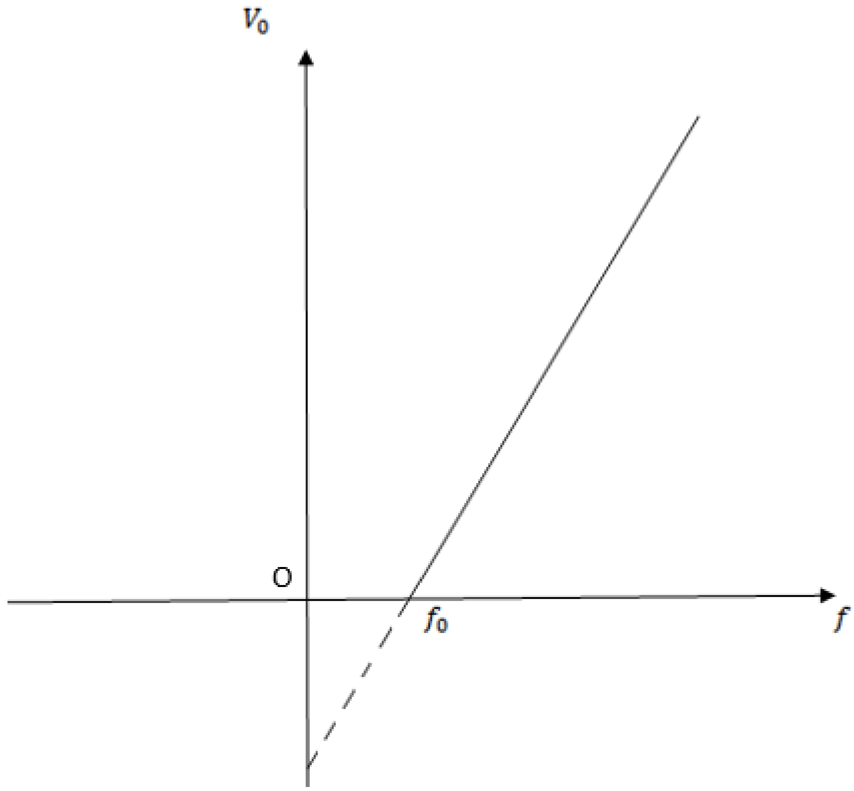
We now study the relation between the frequency *f* of the incident radiation and the stopping potential V_0.

For this purpose, we take the radiations of different frequencies but of the same intensity.

For one frequency say f_1, of the incident radiation, we plot the graph between photoelectric current and potential of anode A with respect to cathode C at a constant intensity of incident radiation

Keeping the intensity of incident radiation the same, we repeat the experiment for frequency f_2 of the incident radiation.

The following is the resulting graph



Observations from the graphs

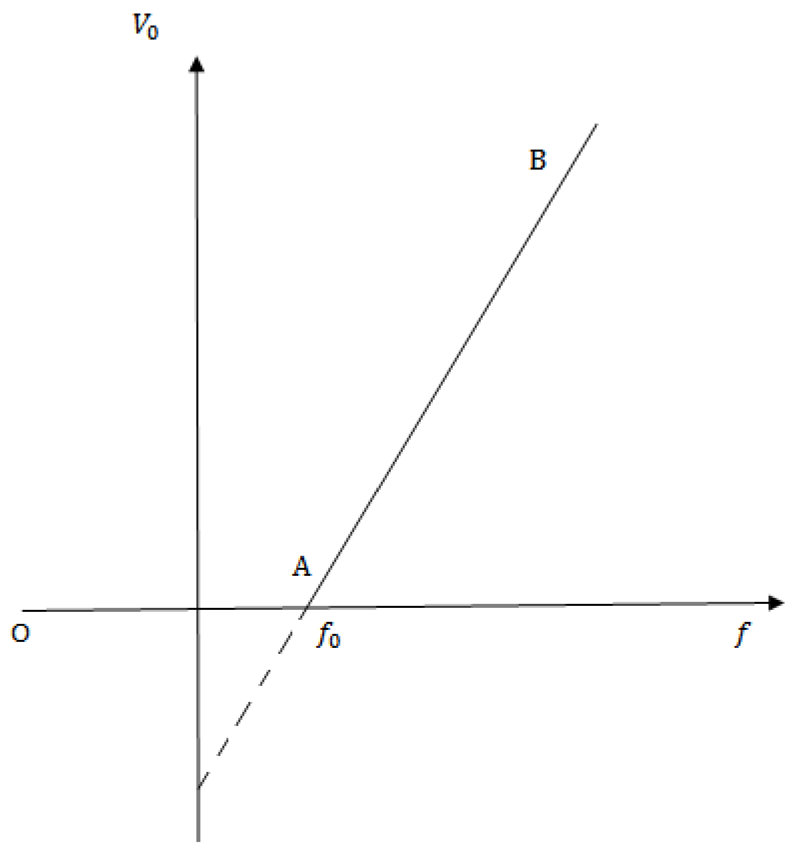
1. The value of stopping potential is different from radiation of different frequencies

2. The value of stopping potential is move in low higher frequency. This implies that the value of maximum kinetic energy depend on the frequency of incident radiation.

   The greater the frequency of incident radiation, the greater is the kinetic energy of emitted photo electrons.

3. The value of saturation current depends on the intensity of incident radiation but is independent of the frequency of incident radiation.

If we draw a graph between the frequency of incident radiation (f) and the stopping potential (V_0) at constant intensity of radiation, it will be a straight line AB as shown in figure below



From the graph

At f_o, stopping potential $V_0 = 0$. It means that at f_o, the photo electric current is just zero (i.e. photo electrons and emitted with zero velocity) and there is no retarding potential.

$V_0 = 0$

This limiting frequency f_o is called threshold frequency for the cathode material.

It is a minimum frequency of the incident radiation which is just sufficient to eject photo electrons (i.e. with zero velocity) from the surface of a metal.

Stopping potential is directly proportional to the frequency of incident radiation.

$$V_0 \propto f$$

The greater the frequency of incident radiation, the higher is the stopping potential and vice versa.

Experiments show that photo electric emission is an instantaneous process.

As soon as light of suitable frequency (equal to or greater than f_o) is incident on the surface of the metal, photo electrons are emitted from the metal surface.

The time delay is less than 10^{-9} second

LAWS OF PHOTOELECTRIC EMISSION

The above experimental study of photoelectric effect leads to the following laws of photoelectric emission.

i. For a given metal, there exists a certain minimum frequency of incident radiation below which no emission of photo-electrons takes place. This cut off frequency is called threshold frequency f_o.

ii. For a given metal and frequency of incident radiation ($>f_o$) the photo electric current is directly proportional to the intensity of incident radiation.

iii. Above the f_o, the maximum kinetic energy of the emitted photo-electron is independent of the intensity of the incident radiation but depends only upon the frequency of the incident radiation.

iv. The photoelectric emission is an instantaneous process.

The above laws of photoelectric emission cannot be explained on the basis of light or radiation. This gave death blow to the wave theory of light or radiation.

FAILURE OF WAVE THEORY/CLASSICAL PHYSICS TO EXPLAIN PHOTOELECTRIC EFFECT

The wave theory of radiation failed to explain photoelectric effect. This will become clear from the following discussion.

I. According to wave theory of radiation the greater the intensity of the wave the greater the energy of the wave.

So wave theory does explain why the number of emitted photoelectrons increase as the intensity of radiation is increased.

But it fails to explain the experimentally observed fact that the velocity or kinetic energy of the emitted photoelectron is independent of the intensity of incident radiation.

According to the wave theory, an increasing in the intensity of radiation should increase the kinetic energy of the emitted photoelectrons but it is contrary to the experimentally observed fact.

II. According to wave theory, intensity of radiation is independent of it is frequency it depends upon the amplitude of electric field vector.

Therefore, an increase in the frequency of radiation should not affect the velocity or kinetic energy of the emitted electrons.

But it is observed experimentally that if the frequency of the incident radiation is increased, the kinetic energy of the emitted electrons also increases.

III. According to the wave theory, electrons should always be emitted from a metal by radiation of any frequency if the incident been is strong enough.

However experiments show that no matter how great is the intensity of the incident radiation; no electrons are emitted from the metallic surface if the frequency of radiation is less than a particular value i.e. threshold frequency.

IV. According to the wave theory the energy of radiation is spread continuously over the wave fronts of the radiation.

Therefore, a single electron in the metal will intercept only a small fraction of the wave's energy.

Consequently considerable time should be needed for an electron to absorb enough energy from the wave to escape the metal surface.

But experiment show that electron are emitted as soon as radiation of suitable frequency falls on the metallic surface.

In other words photoelectric emission is instantaneous there is no delay.

The above discussion is a convincing proof of the inability of the wave theory to explain the photoelectric effect.

EINSTEIN QUANTUM THEORY OF LIGHT

Einstein explained photoelectric effect on the basis of Planck's quantum theory.

According to Einstein light radiation consist of tiny packets of energy called quanta.

Photon

Photon is the single quantum of light radiation which travels with the speed of light.

The energy of a photon is given by E.

$$E=hf$$

where

f –frequency of light radiation

h – Plank's constant.

Further, Einstein assumed that one photon of suitable frequency (=f_o or >f_o) can eject only one photoelectron from the metal surface.

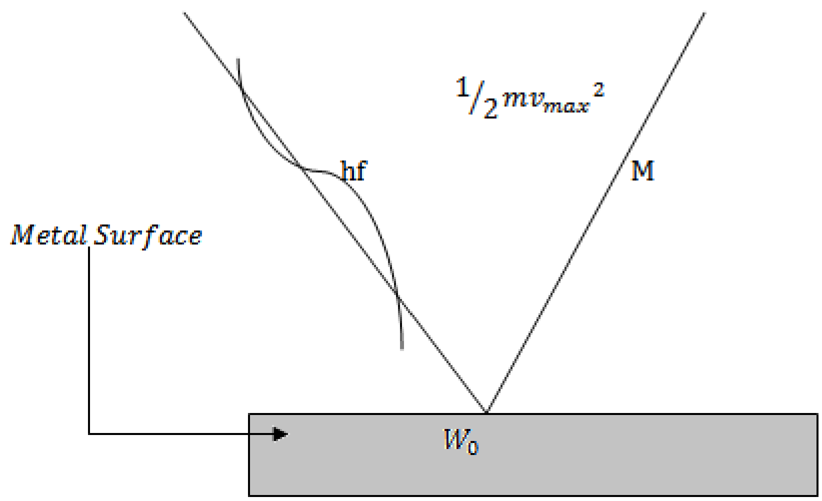
He suggested that the energy of a single photon cannot be shared among the free electron in the metal.

Only one electron can absorb the energy of a single photon.

EINSTEIN'S PHOTOELECTRIC EQUATION

According to Einstein, one photoelectron is emitted from a metal surface if one photon of suitable frequency is incident on the metal.

Suppose a photon of suitable frequency f (< than the f_o for the metal) is incident on the metal.



The energy hf of the photon is spent in two ways

I. A part of photon energy is used in liberating the least tight bound electron from the metal surface which is equal to the work function W_0 of the metal.

II. The rest of the energy of photon appears as maximum kinetic energy of the emitted electron.

Einstein summarized this idea in what is called the Einstein photoelectric equation.

Photon = work function + maximum kinetic energy

$$hf = W_0 + KE_{max}$$

$$hf = hf_0 + {}^1/_2 m v_{max}{}^2$$

The above equation is known as **Einstein's photoelectric equation**

$$hf = hf_0 + {}^1/_2 m(0)^2$$

If the frequency of the incident radiation is fo then the emitted photoelectron will have zero velocity.

$$hf = hf_0$$

Maximum kinetic Energy of emitted photoelectrons is

$$K.E_{max} = hf - W_0$$

$$K.E_{max} = hf - hf_0$$

$$W_0 = hf_0$$

If $f < f_o$, then from above equation $K.E_{max}$ is negative which is impossible therefore, photoelectrons emission cannot occur if the frequency of incident radiation is less than f_o.

If $f > f_o$, then equation above $K.E_{max} \alpha f$. This means that max kinetic energy of photoelectrons depends only on the frequency (f) of the incident radiation.

NUCLEAR PHYSICS

RADIOACTIVITY

This is the emission of radiations from heavily elements such as uranium whose nuclei are unstable.

Radiations emitted are called alpha (α), Beta (β) particles and gamma (γ) rays.

RATE OF DISINTEGRATION

The number of atoms of radioactive elements disintegrating per second dN/dt is directly proportional to the number of atom present at that instant.

$$dN/dt \ \alpha \ -N$$

$$dN/dt = -\lambda N \qquad \text{----(i)}$$

Where λ = decay constant

The negative sign (-) indicates that N decreases as time (t) increases.

If N_0 is the number of atoms at time t= 0 and N is the number of atoms at time t. then:-

$$\int_{No}^{N} \frac{dN}{N} = -\lambda \int_0^t dt$$

$$[\ln N]_{N_o}^{N} = -\lambda t$$

$$\ln \frac{N}{No} = -\lambda t$$

$$e^{-\lambda t} = \frac{N}{No}$$

$$\boxed{\therefore N = N_o e^{-\lambda t}} \qquad \text{.. (2)}$$

HALF LIFE

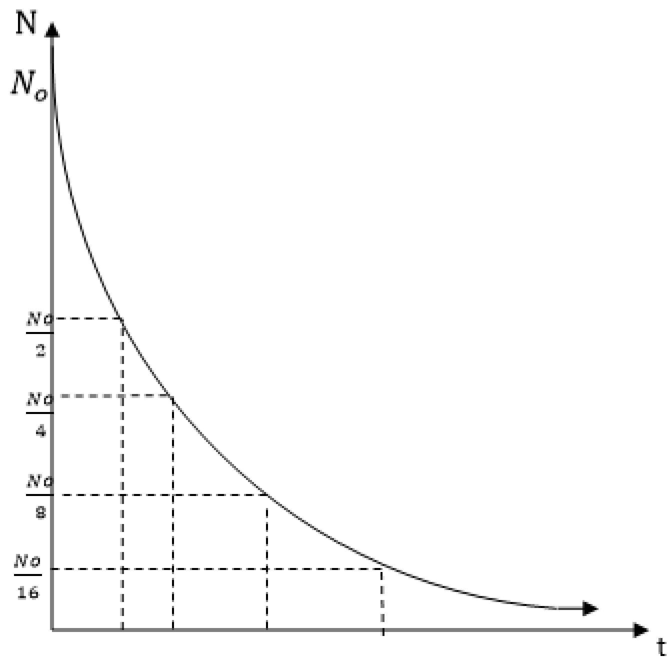
The half life time $(T_{\frac{1}{2}})$ of a radioactive element is the times taken for the atoms disintegrate to half their initial number.

NECTA 1994/1/19

Draw a graph of $N = N_0 \ e^{-\lambda t}$

SOLUTION:

This is the graph of radioactive decay in time.



If N_o is the initial number of atoms

From

$$N = No\, e^{-\lambda t}$$

$$\frac{No}{2} = No e^{-\lambda t_{\frac{1}{2}}}$$

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

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

$$\ln 2^{-1} = -\lambda t_{\frac{1}{2}}$$

$$\boxed{\therefore \frac{T_1}{2} = \frac{0.693}{\lambda}}$$

X-RAYS

X-rays are produced by bombarding a target of heavy metal with high energy electron.

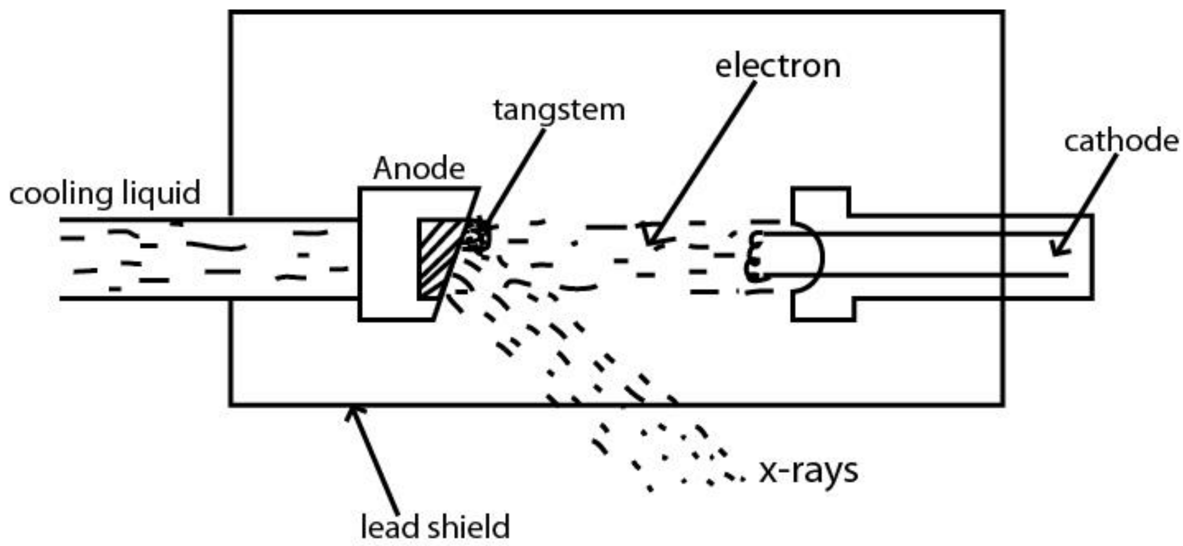
NECTA 1984/2/8

The emission of X-rays may be regarded as the inverse of photo electric effect.Explain

SOLUTION

X-Rays which are waves are produced by bombarding a hard metal with electrons(particles)where as in photoelectric effect electron(particles) are liberated from ametal surface by incident radiation(waves)

MODERN X-RAY TUBE



Electron are obtained from the filament by thermionic emission and are accelerated to the anode having small target of high melting point such as tangstem.

X-RAY QUANTITY

Refers to the intensity of X-rays which increases with the number of electrons limiting the target.This depends on the cathode temperature controlled by the heating current

X-RAY QUALITY

Refer to the penetrating power of X-ray and determined by velocity whith which electrons strikes the target.

SOFT X-RAY:
Are those which can penetrate soft objects such as flesh

HARD X-RAY
Are those which can vibrate much more solid material

PROPERTIES OF X-RAY

1.They travel in a straight line.

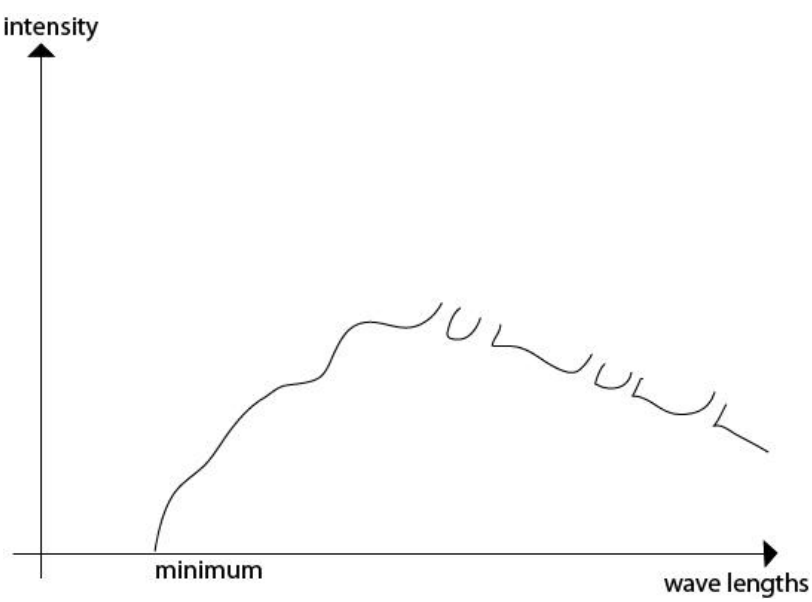
2.They readily penetrate matter.

3.They affect photographic plates

4.They are not deflected by electric or magnetic field,because they have no charge

They are wave of wavelength 1^0A

X-RAY SPECTRA



CONTINUOUS SPECTRUM

Explained by electromagnetic theory.

It is due to electrical interaction between the bombarding electrons and the nuclei of the target atom

An electron approaching a positively charged nucleus is accelerated and according to electromagnetic theory accelerated motion is accompanied by emission of radiations in this case x-ray

DISCRETE SPECTRA

These are explained by quantum theory.They are produced when the incident electron interact with electron close to nucleus of the target atom.The bombarding electrons knock off electron in their orbits and the created gaps are filled by electrons in the high energy levels,when they moves in these gaps they produce x-ray photon.

The cut off wavelength or short wave limit(λ min)

This is the minimum wavelength of x-radiations which corresponds to the maximum energy of the X-rays produced by electrons which have given out all their K.Æ•   on a single encounter with the target nucleus

ENERGY OF X – RAY:

Energy of an electron striking the atom of the target is eV where e = electronic charge.

V = p.d across the X-ray tube.

If a direct collision is made with a target atom and the energy is absorbed then on quantum theory X-rays produced hence a maximum energy hV

Therefore

$$eV = 1/2mV2 = hV$$

$$eV = hc/ \lambda \text{ minimum}$$

$\lambda min = hc/eV$

$\therefore \boxed{\lambda_{minimum} = hc/eV}$

NECTA 1989/1/18

Calculate the wavelength of most energetic X-rays produced by a tube operating at 1.0×10^5V.

USES OF X – RAYS

1.in investigation of suspected bone fractures

2.Detection of lung tuber culosis (feased tissue is denser than a heatthy one).

3.Treatment of cancer (cancer cells are killed by X-rays)

4.Casting and welding joints can be inspected for internal imperfection.

5.It is used in the study of crystal structure refer to Bragg's law.

NUCLEAR FISSION FUSSION

Nuclear fission is a splitting disintegrating of heavy nucleus such as uranium into two other lighter nuclei (with several neutrons)

Example:

$^{235}_{92}U + ^{1}_{0}n \longrightarrow ^{148}_{57}La + ^{85}_{35}Br + 3^{1}_{0}n$

Mass of $^{235}_{92}U + ^{1}_{0}n$

$235.1 + 1.009 = \underline{236.1\mu}$

Mass of $^{148}_{57}La + ^{85}_{35}Br + 3^{1}_{0}n$

$=148.0 + 84.9 + 3(1.009) = \underline{235.9\mu}$

Mass difference

$= 236.1 - 235.9 = 0.2u$

Energy released

$= 0.2 \times 931 =$ mev

NOTE:

$\boxed{1\mu \longrightarrow 931\text{mev}}$

Nuclear fusion

This is the combination of light nuclei to form a heavier nucleus.

Example:

In the sun the fusion of two Deuterons results to helium nucleus

$^{2}_{1}H + ^{2}_{1}H \longrightarrow ^{3}_{2}He + ^{1}_{0}n$

Mass Deterons $= 2\times2.015 = 4.030\mu$

Mass of He + neutron = 3.017 1.009 = 4.026μ

Mass difference = 0.004μ

Energy released

$= \underline{0.004 \times 931 = 3.7 \text{ mev}}$

NECTA 1990/2/8

(1)Both fission and fusion are source of energy

Explain.

SOLUTION:

Fission-During disintegration of heavy nucleus such as uranium, energy is released which is

transformed into K.E of fragments.

Fussion-When light nuclei are fused together a large amount of energy is released in term of heat.(eg

in the sun)

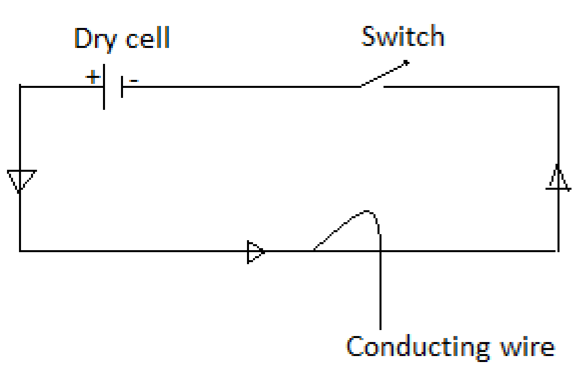
2*T1*

ELECTROMAGNETISM

This is the production of a magnetic field by current flowing in a conductor.

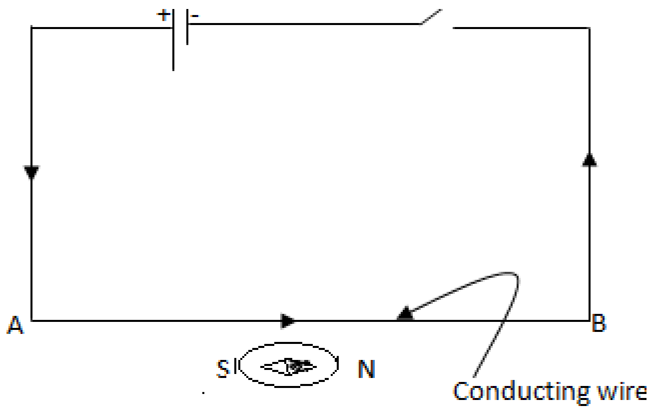
The magnetic effect of current was discovered by Ousted in 1820. The verified magnetic effect of current by the following simple experiment.

Figure below shows a conducting wire AB Above a magnetic needle parallel to it.



So long as there is no current in the wire, the magnetic needle remains parallel to the wire i.e. there is no deflection in the magnetic needle.

As soon as the current flows through the wire AB, the needle is deflected.



Magnetic needle

When the current in wire AB is Reversed the needle is deflected in the opposite direction

This Deflection is a convincing proof of the existence of a magnetic field around a current carrying conductor.

On increasing the current in the wire AB the deflection of the needle is increased and vice versa.

This shows that magnetic field strength increases with the increase in current and vice versa

It is clear from Worsted's experiment that current carrying conductor produces a magnetic field around it.

The larger the value of current in the conductor the stronger is the magnetic field and vice versa.

Magnetic field
Is the region around a magnet where magnet effect can be experienced.
OR
Is the space around a current carrying conducting (magnet) where magnetic effects can be experienced.

The Direction of a field at a point is taken to be the direction in which a north magnetic pole would move more under the influence of field if it were placed at that point.
The magnetic field is represented by magnetic lines of force which form closed loops.

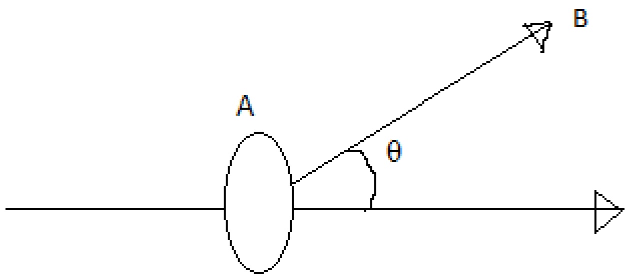
The magnetic field disappears as soon as the current is switched off or charges stop morning.

Magnetic flux Φ

is a measure of the number of magnetic field lines passing through the region.

The unit of magnetic flux is the Weber (Wb)

The flux through an area A on figure below the normal to which lies at angle ð• œf to a field of flux density B



$\emptyset = AB\cos\theta$

Magnetic flux density $(\overrightarrow{\mathrm{B}})$

Is a quantity which measures the strength of the magnetic field

It is sometimes called magnetic

It is a vector quantity

The SI unit of Magnetic flux density is Tesla (T) or Wb/m^2

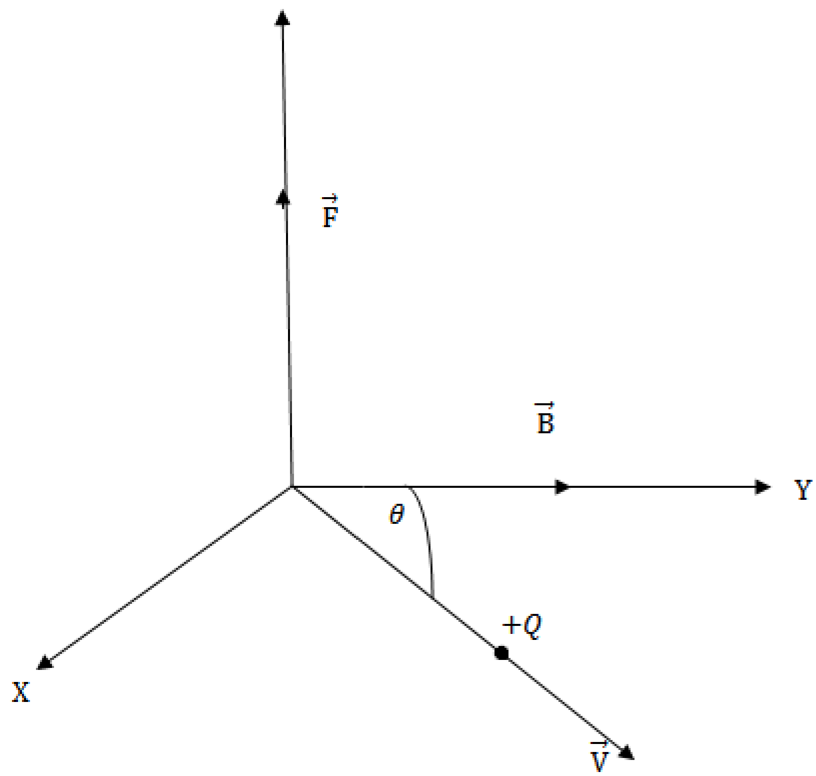
Magnetic flux density is simply called magnetic field B

B = θ/A

FORCE ON A MOVING CHARGE IN A MAGNETIC FIELD

Consider a positive charge +Q moving in a uniform magnetic field $\vec{B}$ with a velocity $\vec{V}$

Let the Angle between $\vec{V}$ and $\vec{B}$ be θ as shown



It has been found experimentally the magnetic field exerts a force F on the charge. The magnitude F of this force depends on the following factors

(i) F α θ

(ii)F α B

(iii) $F \alpha V \sin\theta$

Combining the factors we get

$$F \alpha BQV \sin\theta$$

$$F = kBQV \sin\theta$$

Where K is a constant of proportionality

The unit of B is so defined that K = 1

$$F = BQV \sin\theta \dots\dots\dots (a)$$

Equation (a) can be written in a vector form as:-

$$F = Q(\vec{V} \times \vec{B})$$

F = the force of the particle (N)
B = the magnitude of the magnetic flue density of the field T
Q = the charge on the particle
V= the magnitude of the velocity of the particle

Definition of $\vec{B}$
From
F = BQVsinâ• ¡θ
If V = 1, Q = 1, θ= 90 then

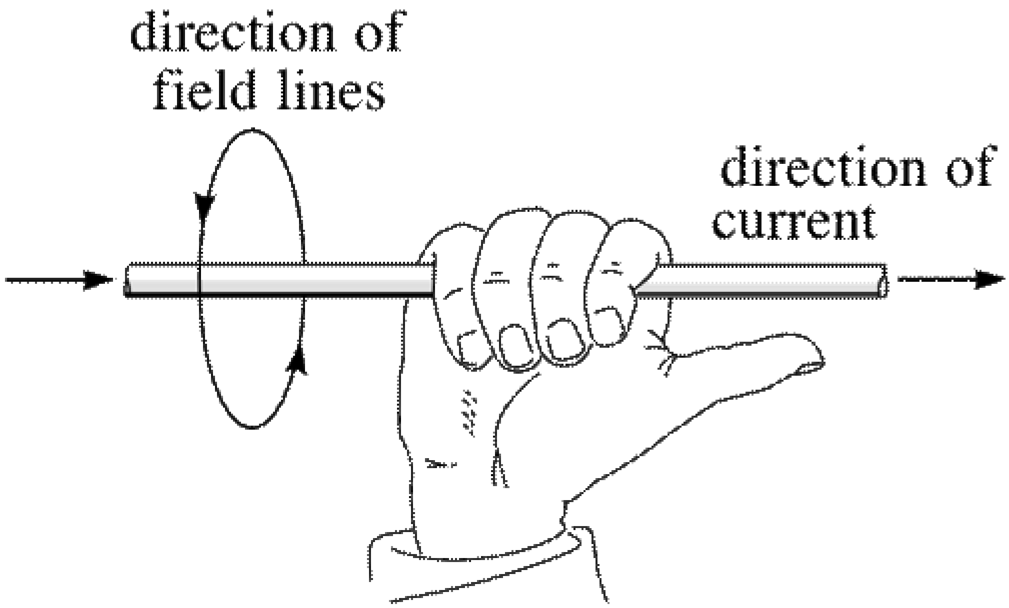
F = Sin90
F=B

Magnetic field ($\vec{B}$) at a point in space is equal to the force experienced by a unit charge moving with a unit velocity perpendicular to the direction of magnetic field at that point

Right Hand Grip Rule

Grip the wire using the right hand with the thumb pointing in the direction of the current the other fingers unit point in the direction of the field.

- For an electron (negatively charged) entering the magnetic field as shown below



The Direction of positive charge will be exactly opposite. Applying Right hand Grip Rule it is clear that Direction of force on the electron will be vertically upward

For a positively charged particle, it will be vertically downward

Direction of magnetic field means from N –pole to S-pole.

SOME CASES OF MAGNETIC FORCE F

Consider an electric charge Q moving with a velocity V through a magnetic field B. then the magnetic force F on the charge is given by

$$F = BQV\sin\theta$$

(i) When $\theta = 0^o$ or 180^0

$$F = BQV\sin\theta$$

$$F = BQV\sin 0^0 \text{ or } F = BQV\sin 180$$

$$F = 0$$

Hence a charged particle moving parallel(or Anti parallel) to the direction of magnetic field experiences no force

(ii) When $\theta = 90^0$

$F = BQV\sin\theta$

$F = BQV\sin 90$

$\sin 90 = 1$

$F = BQV$

Hence a force experienced by charged particle is maximum when it is moving perpendicular to the direction of magnetic field.

(iii) When V=O, the charge particle is at rest.

$F = BQV\sin\theta$

$F = BQ(0)\sin\theta$

$\therefore F = O$

If a charged particle is at rest in a magnetic field it experiences no force.

(iv) When Q = O

$F = BQV\sin\theta$

$F = 0$

Hence electrically neutral particle (eg neutron) moving in a magnetic field experiences no force.

The magnetic force F acts perpendicular to velocity V (as well as B)

This means that a uniform magnetic field can neither speed up nor slow down a moving charged particle; it can charge only the Direction of V and not magnitude of V

Since the magnitude of V does not charge the magnetic force does not change the kinetic energy of the charged particle.

UNITS AND DIMENSIONS OF $\vec{B}$

The SI unit of magnetic field B is **Tesla**

Now

F = BQV$^{\sin\theta}$

$$B = \frac{F}{QV\sin\theta}$$

If Q = 1C, V =1m/s, Q= 90^0 F= 1N

B = 1T

Hence the strength of magnetic field at a point is 1T if a charge of 1C when moving with a velocity of 1m/s at right angles to the magnetic field, experiences a force of 1N at that points.

Magnetic field of earth at surface is about 10^{-4}T. On the other hand, strong electromagnets can produce magnetic fields of the order of 2T.

Dimensions of $\vec{B}$

$$B = \frac{F}{QV\sin\theta}$$

$$B = \frac{MLT^{-2}}{[AT][LT^{-1}]}$$

$$[B] = [MA^{-1}T^{-2}]$$

$$\therefore \text{The dimension of } \vec{B} = \frac{MLT^{-2}}{[AT][LT^{-1}]}$$

Worked Examples

1. A proton is moving northwards with a velocity of $5\ x\ 10^6$ m/s in a magnetic field of 0.1Tdirected eastwards. Find the force on the proton. Charge on proton = 1.6 x 10^{-19}C.

Solution

F = BQV$^{\sin\theta}$

B= 0.1T

V= $5x10^6$ m/s

F=0.1 X 1.6 X 10^{-19} X 5 X 10^6X Sin 90

Q = 1.6 X 10^{-19}C

θ= 90^0

$$F = 8 \times 10^{-14}N$$

2. An electron experiences the greatest force as it travel at 3.9 x10^5 m/s in a magnetic field when it is moving westward. The force is upward and is of magnitude $8.2x10^{-13}$ N what is the magnitude and direction of the magnetic field.

Solution

The conditions of the problem suggest that the electron is moving at right angles

$\theta = 90^\circ$ To the direction of the magnetic field

F = BQV$\sin\theta$, F = 8.7 x 10 $^{-13}$N

Q= 1.6 X10 $^{-19}$C

V=3.9X10^5m/s

$$B = \frac{F}{QV\sin\theta}$$

$$B = \frac{8.2 \times 10^{-13}}{1.6 \times 10^{-19} \times 3.9 \times 10^5 \sin 90}$$

B = 13.14T

By right hand rule per cross product, the direction of the magnetic field is towards northward.

3. An α - particle of mass 6.65 x 10^{-27} kg is travelling at right angles to a magnetic field with a speed of 6x10^5m/s. The strength of the magnetic field is 0.2T.calculate the force on the α - particle and its acceleration.

Solution

Force on α – particle F = BQV$\sin\theta$

M = 6.65 X10^{-27}Kg

V = 6 x 10^5m/s

B = 0.2T

θ = 90^0

F = BQV$\sin\theta$

= (0.2 x 2x 1.6x10^{-19}) x $(6\ x\ 10^5)$x Sin90Ëš

$$F = 3.84 \times 10^{-14}N$$

Acceleration of α – particle

F= mÉ‘

É‘= $\frac{F}{m}$ = $\frac{3.84\,x10^{-14}}{6.65x\,10^{-27}}$

$$a = 5.77 \times 10^{12}\ m/s^2$$

4. A copper wire has 1.0 x 10^{29} free electrons per cubic meter, a cross sectional area of 2mm^2 and carries a current of 5A . The wire is placed at right angle to a uniform magnetic field of strength 0.15T. Calculate the force the acting on each electron.

Solution

I = neAV_d

Drift velocity = $\frac{I}{neA}$

n= 1x$10^{29}m^{-3}$ e = 1.6x10^{-19}c A= 2mm^2 = 2x$10^{-6}m^2$

I = 5A

$$V_d = \frac{1}{neA}$$

$$V_d = \frac{1}{1 \times 10^{29} \times 1.6 \times 10^{-19} \times 2 \times 10^{-6} m^2}$$

$$V_d = 1.56 \times 10^{-4}\, m/s$$

Force on each electron F= BQV_d Sinθ

Q= 1.6 x 10^{-19}c

B= 0.15T

Q=90^0

$$F = 0.15 \times 1.6 \times 10^{-19} C \times 1.56 \times 10^{-4} \sin 90$$

$$F = 3.75 \times 10^{-24} N$$

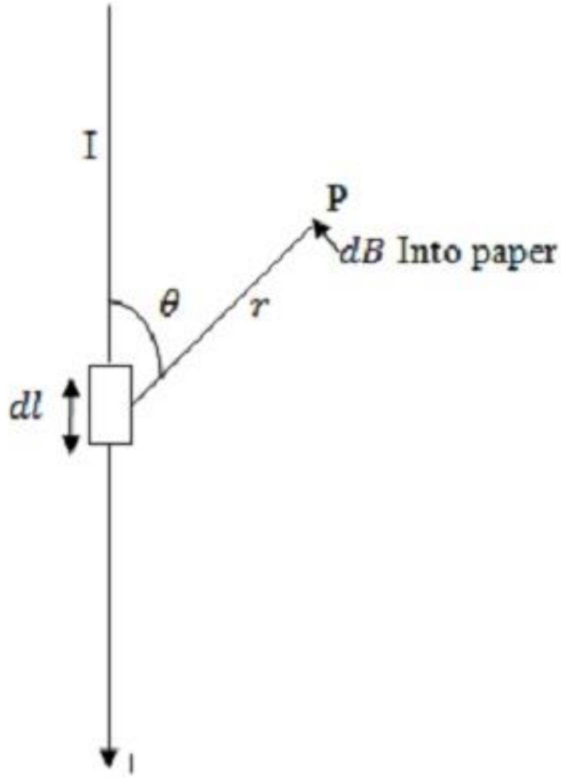
BIOT –SAVART LAW

The Biot – Savart law states that the magnitude of magnetic flux density dB at a point P which is at a distance r from a very short length dl of a conductor carrying a current I is given by.

$$dB \propto \frac{Idl \sin\theta}{r^2}$$

$$dB = \frac{KIdl \sin\theta}{r^2}$$

where θ is the Angle between the short length dl and the line joining it to point P



K is a constant of proportionality its value depends on the medium in which the conductor is situated and the system of units adopted.

For free space vacuum or air

$$K = \frac{\mu_0}{4\pi}$$

where μ_0 = Absolute permeability of free space $= 4\pi \times 10^{-7} Tm/A$

$$\overrightarrow{dB} = \frac{\mu_0}{4\pi}\frac{Idl\sin\theta}{r^2}$$

This equation is known as Biot –Savart Law and gives the magnitude of the magnetic field at a point due to small current element

Current element

Is the product of current (I) and length of very small segment ($\overrightarrow{dl}$) of the current carrying conductor.

Current element = $I\vec{dl}$

Current element produces magnetic field just as a stationary charge produces an electric field the current element is a vector.

Its Direction is Tangent to the element and acts in the direction of current flow in the conductor

Biot -Savart law holds strictly per steady currents

Direction of B

$$\overrightarrow{dB} = \frac{\mu_0}{4\pi}.\frac{I(\vec{dl} \times \vec{r})}{r^2}$$

The direction of $\overrightarrow{dB}$ is perpendicular to the plane containing $\vec{dl}$ and $\vec{r}$ by right hand rule for the cross product the field is directed inward.

Special cases

$$dB = \frac{\mu_0}{4\pi}.\frac{Idl\sin\theta}{r^2}$$

(i) When $\theta = 0^0$ or 180^0
i.e Point P lies on the axis of the conductor

$$dB = \frac{\mu_0}{4\pi}\frac{Idl\sin 0^\circ}{r^2}$$

$$dB = 0$$

Hence there is no magnetic field at any point on the thin current carrying conductor minimum value.

(ii) When θ = 90^0

When point P lies at a perpendicular position *w .r. t* current element

$$dB = \frac{\mu_0}{4\pi}\frac{Idl\sin 90°}{r^2}$$

$$dB = \frac{\mu_0}{4\pi}\frac{Idl}{r^2}$$

Hence magnetic field due to a current element is maximum in a plane passing through the element and perpendicular to its axis.

Important point about Biot – Savant law

(i) Biot – Savant law is valid per symmetrical current distributions.

(ii) Biot – Savant law cannot be proved experimentally because it is not possible to have a current carrying conductor of length dl

(iii) Like coulomb's law in electrostatics, Biot- Savant law also obeys inverse square law

(iv) The Direction of dB is perpendicular to the plane containing $I\vec{dl}$ and $\vec{r}$

(v) This law is also called Laplace's law and inverse square law'

BIOT – SAVART LAW VERSUS COULOMB'S LAW IN ELECTROSTATICS

According to coulomb's law in electrostatics, the eclectic field due to a charge element dQ at a distance r is given by

$$dE = \frac{1}{4\pi\varepsilon_0}\cdot\frac{dQ}{r^2}$$

According to Biot – Savart law the magnetic field due to a current element $I\vec{dl}$ at a distance r is given by

$$dB = \frac{\mu_0}{4\pi} \cdot \frac{Idl \sin\theta}{r^2}$$

From the above two equations we note the following points of Similarities and Dissimilarities.

Similarities

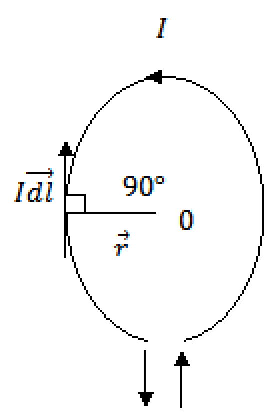
(i) Both laws obey inverse square law
(ii) Both the fields(magnetic field and Electro static field) obey superposition principles
(iii)Both the fields are long range fields.

Dissimilarities

(i) The Electric field is produced by a scalar source i.e. Electric charge dQ. However the magnetic field is product by a vector source i.e. current $I\vec{dl}$
(ii) The Direction the Electric field is along the displacement vector i.e. The line joining the source and field point. However the direction of magnetic field is perpendicular to the plane containing current element $I\vec{dl}$ and displacement vector $\vec{r}$
(iii) In Biot –Savant law the magnitude of magnetic field dB α Sin θ Where θ is the Angle between current element $I\vec{dl}$ and displacement vector $\vec{r}$ However there is no angle dependence in coulomb's law for electrostatics

MAGNETIC FIELD AT THE CENTER OF CURRENT CARRYING CIRCULAR COIL

Consider a circular coil of radius r and carrying current I in the Direction shown in figure



Suppose the loop lies in the plane of paper it is desired to find the magnetic field at the centre O of the coil

Suppose the entire circular coil is divided into a large number of current elements each of length dl

According to Biot – Savant law, the magnetic field $\overrightarrow{dB}$ at the centre O of the coil due to current element Idl is given by

$$dB = \frac{\mu_0}{4\pi} \cdot \frac{Idl \sin\theta}{r^2} \text{..............} (i)$$

The direction of dB is perpendicular to the plane of the coil and is Directed inwards

Since each current element contributes to the magnetic field in the same direction, the total magnetic field B at the centre O can be found by integrating equation…………(i)

$$B = \int dB$$

$$B = \int \frac{\mu_0}{4\pi} \cdot \frac{Idl \sin\theta}{r^2}, \theta = 90°$$

$$B = \int \frac{\mu_0}{4\pi} \cdot \frac{Idl}{r^2}$$

$$B = \frac{\mu_0 I}{4\pi r^2} \int dl$$

$$B = \frac{\mu_0 I}{4\pi r^2} \times L$$

L- Total length of the coil = 2 πr

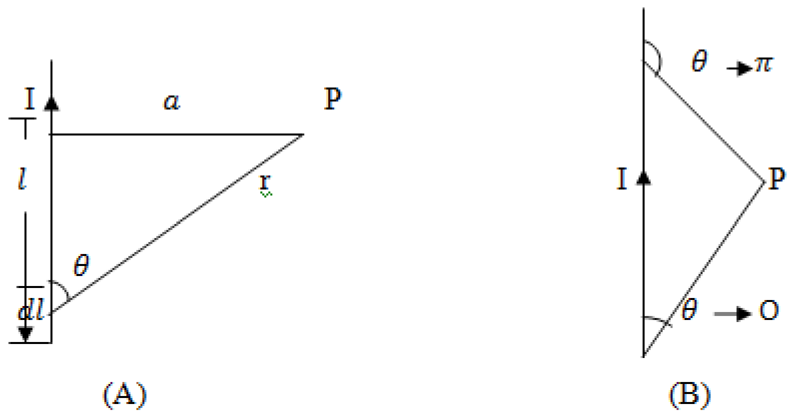
$$\text{Then, } B = \frac{\mu_0 I}{4\pi r^2} \times 2\pi r$$

$$B = \frac{\mu_0 I}{2r}$$

If the coil has N turns each carrying current in the same direction then contribution of all turn are added up.

B= $\frac{\mu_0 NI}{2r}$

MAGNETIC FIELD DUE TO INFINITELY LONG CONDUCTOR



The flux density dB at P due to the start length dl given by equation as

$dB = \frac{\mu_0}{4\pi} \cdot \frac{Idl \sin\theta}{r^2}$

From the figure (A)

$\sin\theta = \frac{a}{r}$, $\tan\theta = \frac{a}{l}$

$r = \frac{a}{\sin\theta}$ $l = \frac{a}{tan\theta}$

r = $\frac{a}{sin\theta}$

l = a cot θ

dl = -a$cosec^2\theta d\theta$

Substituting for r and dl gives

$$dB = \frac{\mu_0}{4\pi}.\frac{I \times (-Cosec^2\theta d\theta)Sin\theta}{{}^{a}/_{Sin^2\theta}}$$

$$dB = \frac{-\mu_0 I}{4\pi a}.Sin\theta d\theta$$

The total flux density B at P is the sum of the flux densities of all the short lengths and can be found by letting dθ→O and integrating over the whole length of the conductor.

$$\int_{\pi}^{0} dB = \int_{\pi}^{0} \frac{-\mu_0 I}{4\pi a}.Sin\theta d\theta$$

$$B = -\frac{\mu_0 I}{4\pi a}\int_{\pi}^{0} Sin\,\theta d\theta$$

The limits of the integration are π and 0 because these are values of ð• œf at the ends of the conductor

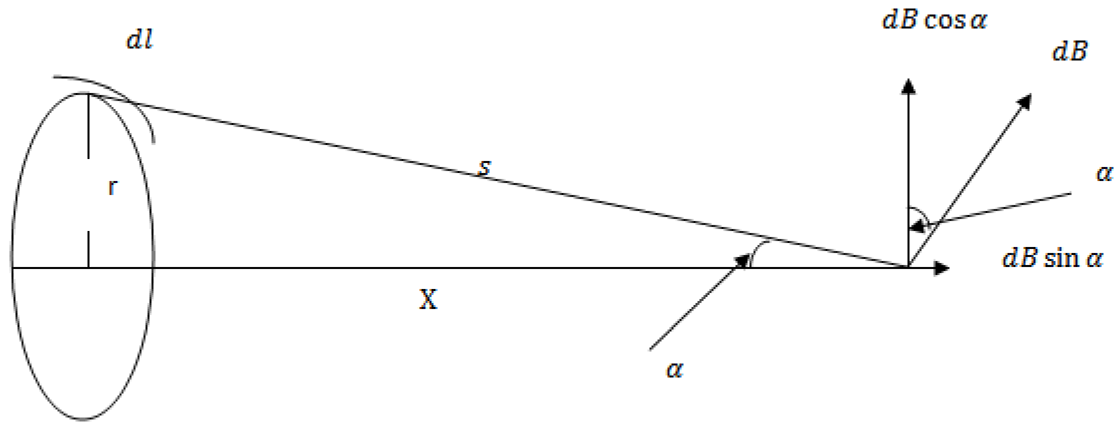
$$B = -\frac{\mu_0 I}{4\pi a}[-cos\theta - -cos\pi]$$

$$B = -\frac{\mu_0}{4\pi a}[-1 + (-1)]$$

$$B = -\frac{\mu_0 I}{4\pi a} \times (-2)$$

$$B = \frac{\mu_0 I}{2\pi a}$$

FLUX DENSITY AT ANY POINT ON THE AXIS OF A PLANE CIRCULAR



Circular coil with its plane perpendicular to that of the paper

The flux density dB at p due to the short length dl of the coil at X, where X is in the plane of the paper, is given by equation as

$$dB = \frac{\mu_0}{4}.\frac{Idlsin\theta}{s^2}$$

By symmetry, when all the short lengths dl are taken into account the components of magnitude $dB\cos\alpha$ sum to zero.

Each short length produces a component of magnitude dB Sin α parallel to the axis and all those components are in the direction shown

The total flux density is therefore in the direction of dB Sin α and its magnitude B is given by

$$B = \sum dB\ Sin \propto$$

$$B = \sum \frac{\mu_0}{4\pi}.\frac{Idlsin\theta}{s^2}.Sin \propto$$

The radius vector XP of each small length is perpendicular to it, so that $\theta=90^0$ and there pore Sinθ = 1

$$B = \frac{\mu_0}{4\pi} \cdot \frac{I\sin \propto}{s^2} \sum dl$$

Since,

$\sum dl$ = 2 πr (the circumference of the coil)

$$B = \frac{\mu_0}{4\pi S^2} \cdot I \sin \propto \times 2\pi r$$

$$B = \frac{\mu_0 I r S i n\alpha}{2\, S^2}$$, But $\sin \propto = \frac{r}{s}$

$$B = \frac{\mu_0 I r}{2S^2} \times \frac{r}{s}$$

$$B = \frac{\mu_0 I r^2}{2s^3}$$

For a coil of N Turns

$$B = \frac{\mu_0 N I r^2}{2s^3}$$

When S= r

$$B = \frac{\mu_0 N I}{2r}$$

Also from the figure

$$S^2 = r^2 + x^2$$

$$B = \frac{\mu_0 N I r^2}{2(r^2 + x^2)^{3/2}}$$

AMPERE'S CIRCUITAL LAW

States that the line integral of magnetic field $\vec{B}$ around any closed path in vacuum/air is equal to μ_0 times the total current (I) enclosed by that path

$$\oint \vec{B}.\vec{dl} = \mu_0 I$$

$\oint \vec{B}.\vec{dl}$ = line integral of $\vec{B}$ around closed path

I = current enclosed by that path.

Ampere's law is an alternative to Biot - Savart law but it is useful for calculating magnetic field only in situations with considerable symmetry.

This law is true for steady currents only.

In order to use law it is necessary to choose a path for which it is possible to determine the value of the line integral

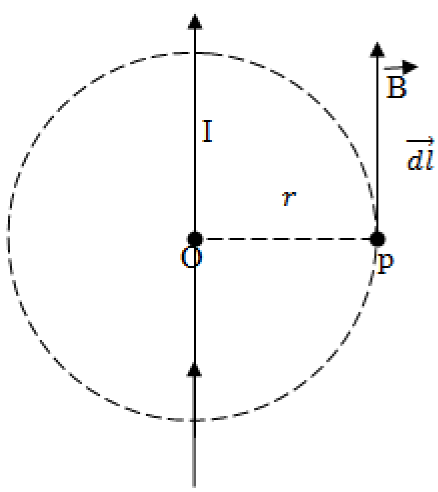
It is because there are many situations where there is no such path that the law is of limited use.

Hence the application of ampere law
(i) Magnetic field due to constraining conductor carrying current
(ii)Magnetic field due to solenoid carrying current
(iii)Magnetic field due toroid

MAGNETIC FIELD DUE TO STRAIGHT CONDUCTOR CARRYING CURRENT

Consider a long straight conductor carrying current I in the direction as shown in the figure below

It is desired to find the magnetic field at a point p at a perpendicular distance r for the conductors



Applying Ampere's circuital law to this closed path

$$\oint \vec{B}.\overrightarrow{dl} = \mu_0 I$$

$$B\oint dl = \mu_0 I$$

$$B \times 2\pi r = \mu_0 I$$

$$B = \frac{\mu_0 I}{2\pi r}$$

SOLENOID

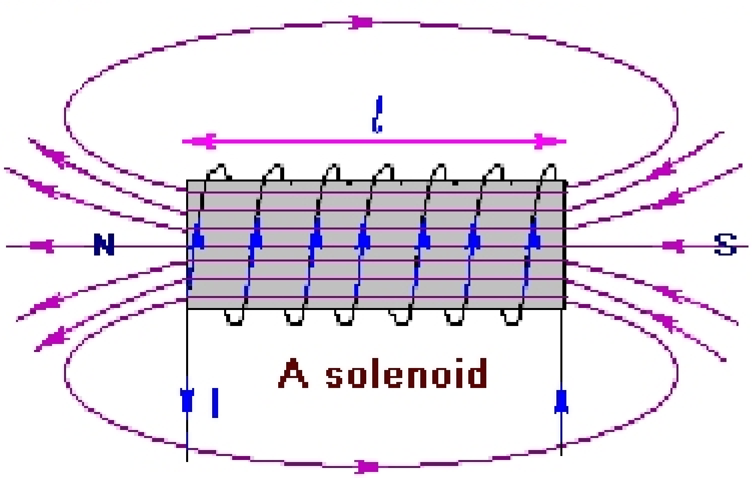
Is a long coil of wire consisting of closely packed loops

Or

Is a cylindrical coil having many numbers of turns

By long solenoid we mean that the length of the Solenoid is very large as compared to its Diameter.

Figure below shows the magnetic field lines due to an air cored solenoid carrying current



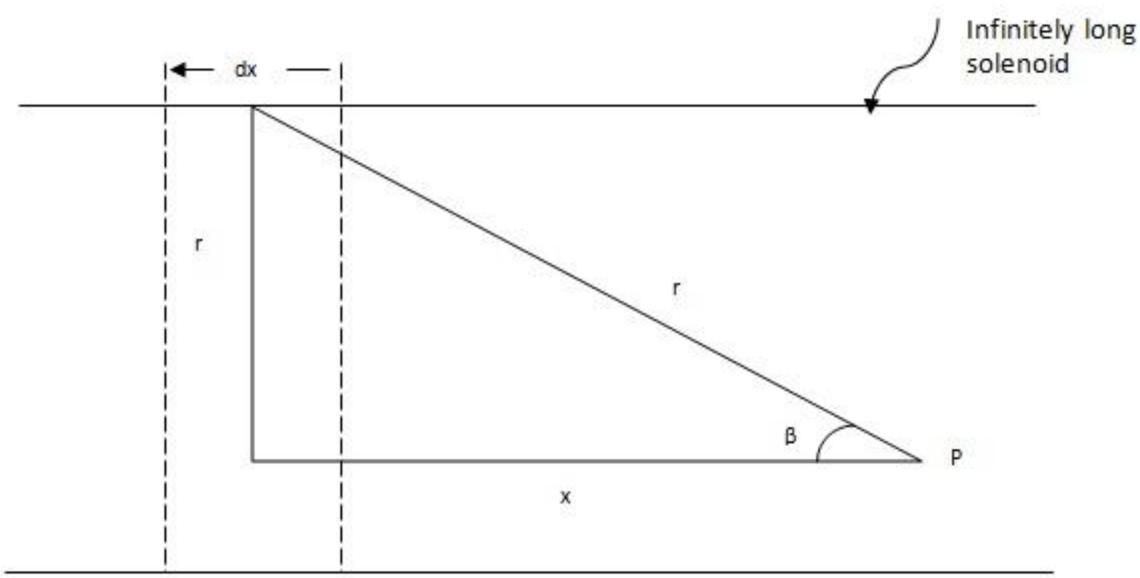
Inside the solenoid the magnetic field is uniform and parallel to the solenoid axis.

Outside solenoid the magnetic field is very small as compared to the field inside and may be assumed zero.

It is because the same no of field line that are concentrated inside the solenoid spread out into very faster space outside

Magnetic flux density due to an Axis of an in finely long Solenoid

Consider the magnetic flux density dB at P due to a section of the solenoid of length dx



n = number of turns per unit length.

N= number of turns the section can be treated as a plane circular coil of N turns in which case dB is given by

$$n = \frac{N}{dx}$$

Since *dx* is small, the section can be treated as a plane circular coil or N turns in which case dB is given by

$$dB = \frac{\mu_0 N I r^2}{S^3}$$

$$dB = \frac{\mu_0 I r^2}{S^3} \; .nd \ldots\ldots\ldots (1)$$

From the figure

$$\sin\beta = \frac{r}{s}$$

$$s = \frac{r}{\sin\beta} = rcosec\,\beta$$

$$S^3 = r^3 cosec^3\beta$$

Also

$$\tan\beta = \frac{r}{x}$$

$$x = \frac{r}{\tan\beta} = rcot\beta$$

$$x = rcot\beta$$

$$dx = -rcosec^3\beta d\beta$$

Substituting for s^3 and dx gives,

$$dB = \frac{\mu_0 I r^2}{s^3} \; .ndx$$

$$dB = \frac{\mu_0 I r^2}{r^3 cosec^3\beta} \; .n(-rcosec^2\beta d\beta)$$

$$d\beta = -\frac{\mu_0 nI}{2} Sin\beta d\beta$$

$$dB = -\frac{\mu_0 nI}{2} Sin\,\beta d\beta$$

The flux densities at P due to every section of the Solenoid are all in the same direction and therefore the total flux density B can be found by letting dB→o and integrate over the whole length of the solenoid.

$$B = \int_{\pi}^{0} dB$$

$$B = -\frac{\mu_0 nI}{2}\int_{\pi}^{0} \sin\beta d\beta$$

The limits of integration are π and 0 because these values of β at the end of the solenoid.

$$B = -\frac{\mu_0 nI}{2}[-\cos\beta]_{\pi}^{0}$$

$$B = \frac{\mu_0 nI}{2}\,[\![-Cos\,0 - \, -cos\pi]\!]$$

$$B = -\frac{\mu_0 nI}{2} \times -2$$

$$B = \mu_0 NI$$

$B = \mu_0 nI \longrightarrow$ Inside the Solenoid

If the Solenoid is Iron-cored of relatively permeability μ_r magnitude of magnetic field inside the Solenoid is

From

$$\mu_r = \frac{\mu_m}{\mu_0}$$

$$\mu_m = \mu_r \mu_0$$

$$B = \mu_m nI$$

At points near the ends of an air cored Solenoid, the magnitude of magnetic field is

$$B = -\frac{\mu_0 nI}{2}\int_{\pi}^{\pi/2} Sin\, \beta d\beta$$

$$B = \frac{-\mu_0 nI}{2}[\![-Cos\beta]\!]_{\pi}^{\pi/2}$$

$$B = \frac{-\mu_0 nI}{2}\left[\!\!\left[-Cos\frac{\pi}{2} - cos\pi\right]\!\!\right]$$

$$B = \frac{-\mu_0 nI}{2}[0-1]$$

$$B = \frac{\mu_0 nl}{2}$$

The magnetic field outside a solenoid is zero

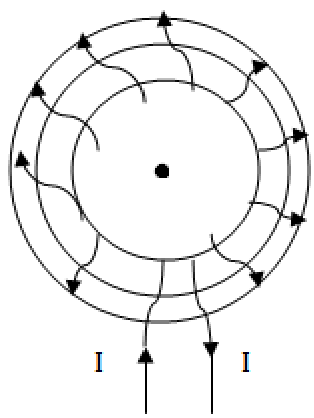
Also in a current carrying long solenoid the magnetic field produced does not depend upon radius of the Solenoid.

TOROID

Toroid is a solenoid that bent into the form of the closed ring.

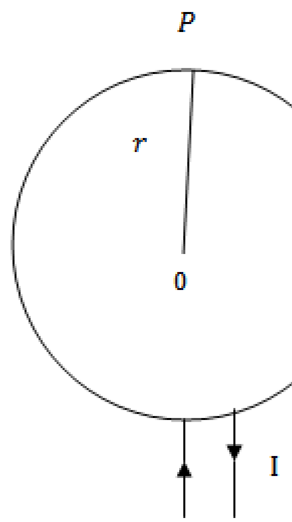
The magnitude field B has a constant magnitude every where inside the toroid while it is zero in the open space interior and exterior to the toroid.

If any closed path is inside the inner edge of the toroid then ther is no current enclosed. Therefore, by Ampere's circular law B=0.

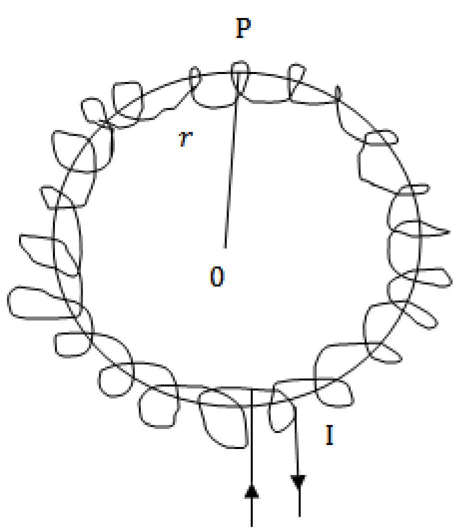


Magnetic field $\vec{B}$ due to toroid

Consider the diagram below



Let r = mean radius of toroid
I = Current through toroid
n = number of turns permit length
B = magnitude of magnetic field inside the toroid



$$\oint \vec{B}.\overrightarrow{dl} = \oint Bdl\cos\theta = B(2\pi r)$$

$$= B(2\pi r) \ldots \ldots \ldots \ldots . (\mathrm{i})$$

According to Amperes circular law

$$\oint \bar{B}.\,\bar{d}\bar{l} = \mu_0 \times \mathrm{I}(\text{enclosed path})$$

The total current enclosed

$$= \mathrm{n} \times \text{length of path} \times \mathrm{I}$$

$$= \mathrm{n}(2\pi \mathrm{r})\mathrm{I}$$

$$\therefore \oint \vec{B}.\overrightarrow{dl} = \mu_0 \times \mathrm{n}(2\pi \mathrm{r})\mathrm{I} \ldots \ldots \ldots . (\mathrm{ii})$$

from eq(i)and (ii) we get

$$\mathrm{B}(2\pi \mathrm{r}) = \mu_0 \times \mathrm{n}(2\pi \mathrm{r})\mathrm{I}$$

$$\therefore \mathrm{B} = \mu_0 \mathrm{nI}$$

This expression is the same as for air cored Solenoid

Then

FORCE ON A CURRENT CARRYING CONDUCTOR PLACED IN A MAGNETIC FIELD

We know that a moving charge in a magnetic field experiences a force

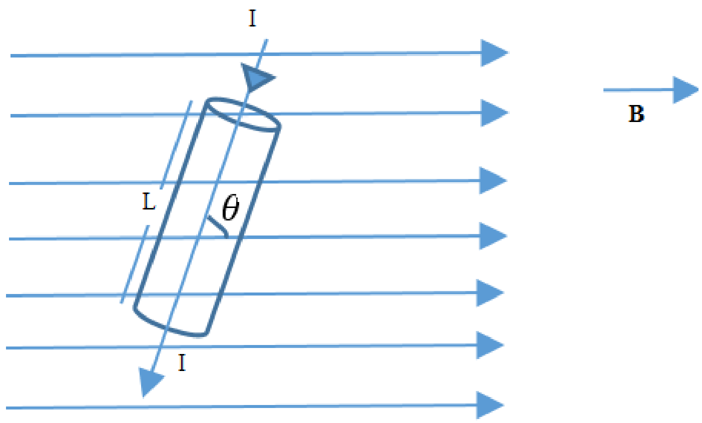
Now electric current in a conductor is due to the drifting of the force electrons in a definite direction in the conductor

When such a current carrying conductor is placed in a uniform magnetic field, each free electron experiences a force.

Since the free electrons are constrained in the conductor, the conductor itself experiences a force.

Hence a current carrying conductor placed in magnetic field experiences a force F.

Consider a conductor of the length L and area of cross- section a placed at an angle ð• œf to the direction of uniform of magnetic field B.



θ- is the angle between the plane of the conductor. The magnetic force experienced by the moving charge in a conductor is F = BQV Sinθ

For steady current $I = \frac{Q}{t}$

Q =I t

$F = BIt\frac{L}{t}\sin\theta$

The velocity for direct current is constant

$V = \frac{L}{t}$

$F = B\,I\,t\frac{L}{t}\sin\theta$

$F = BIL\sin\theta$

F= Force on the conductor (N)

B= Magnitude of the magnetic flux density of the field (T)

I = Current in the Conductor (A)

L= length of the conductor (M)

The current in the conductor I

$I = neAV_d$

Special cases

$(i)\ \boldsymbol{\theta = 0°}$

$F = BIL \sin 0$

$F = 0$

Thus if current carrying conductor is placed parallel to the direction of the magnetic field of the conductor will experience no force.

ii.) $\boldsymbol{\theta = 90°}$

F = BIL$\sin 90°$

$\therefore F = BIL$

Hence current carrying conductor will experience maximum force when it placed at right angles to the direction of the field.

One Tesla

Is the magnetic flux density of a field in which a force of IN acts on a 1M length of a conductor which is carrying a current of IA and is perpendicular to the field.

$$B = \frac{F}{IL}$$

B = Tesla

$$Tesla = \frac{1N}{1A.1M}$$

The Direction of the force

Experiment shows that the force is always perpendicular to the plane which contains both the current and the external field at the site of the conductor

The direction of the force can be found by using Fleming's left hand rule

Fleming's left hand rule

States that if the first and the second fingers and the thumb of the left hand are placed comfortably at right angles to each other, with the first finger pointing in the direction of the current then thumb points in the direction of the force i.e. Direction in which Motion takes place If the conductor is free to move.

Maxwell's Corkscrew rule

States that if a right handed corkscrew is turned so that its point travels along the direction, the direction of rotation of corkscrew gives the direction of the magnetic field.

FORCE BETWEEN TWO PARALLEL CONDUCTORS CARRYING CURRENTS

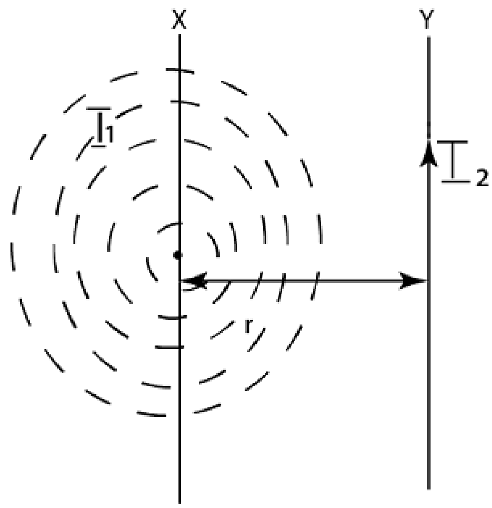
When two parallel current carrying conductors are close together, they exert force on each other.

It is because one current carrying conductor is placed in the magnetic field of the other

If currents are in the same direction the conductor attract each other and If currents are in the opposite directions conductors repel each other

Thus like currents attract, unlike currents repel.

Consider two infinitely long straight parallel conductors X and Y carrying currents I_1 and I_2 respectively in the same direction.



Suppose the conductors are separated by a distance ***r***in the plane of the paper.

As each conductor is in the magnetic field produced by the other, therefore each conductor experiences a force

The current carrying conductor Y is placed in the magnetic field produced by conductor X

Therefore force act on the conductor Y. The magnitude of the magnetic field at any point P on the conductor Y due to current I, in the conductor X is

$$B_X = \frac{\mu_0 I_1}{2\pi r}$$

By right hand grip rule ; the direction of B is perpendicular to the place of the paper and is directed inwards.

Now conductor Y carrying current I_2 is placed in the magnetic field B_X produced by conductor X

Therefore force per unit length of conductor Y will experience a force F_Y given by

$$\frac{F_Y}{L} = B_X I_2$$

$$\frac{F_Y}{L} = \frac{\mu_0}{2\pi} \cdot \frac{I_1 I_2}{r}$$

According to FLHR, force F_Y on conductor Y acts in the place of the paper perpendicular to Y and is directed towards to the conductor X.

Similarly, the Force on conductor X per unit length is $F_X = B_y I_1 L$

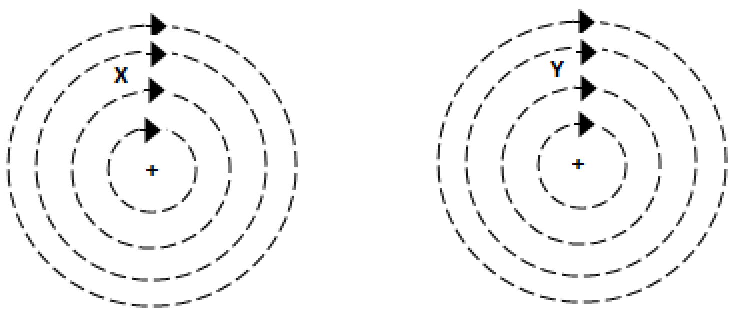
But

$$B_Y = \frac{\mu_0 I_2}{2\pi r}$$

$$\frac{F_x}{L} = \frac{\mu_0 I_1 I_2}{2\pi r}$$

Hence when two long parallel conductors carry currents in the same direction they attract each other. The force of attraction per unit length is

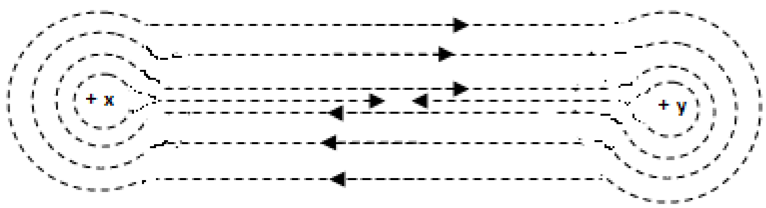
$$\frac{F}{L} = \frac{\mu_0}{2\pi}\frac{I_1 I_2}{r}$$



This shows that the attraction between two parallel straight conductors carrying currents in the same direction in terms of magnetic field lines of conductors

It is clear that in the space between X and Y the two fields are in opposition and hence they tend to cancel each other

However in the space outside X and Y the two fields assist each other. Hence resultant field distribution will be



If two straight current carrying conductors of unequal length are held parallel to each other then force on the long conductor is due to the magnetic field of the short conductor

I_1 = Current through short conductor

l = Length of short conductor

I_2 = Current through long conductor

L = Length of long conductor

If r is the separation distance between these parallel conductors

Force on Long conductor = force on short conductor

$$= \frac{\mu_0}{2\pi} \frac{I_1 I_2}{r} l$$

Force on each conductor is the same in magnitude but opposite in direction (Newton's third law)

DEFINITION OF AMPERE

Force between two current currying conductors per unit length

$$F = \frac{\mu_0}{2\pi} \cdot \frac{I_1 I_2}{r}$$

If $I_1 = I_2 = 1A$ And r =1m then

$$F = \frac{4\pi \times 10^{-7}}{2\pi}$$

$$F = 2 \times 10^{-7} N$$

Ampere

Is that steady current which when it is flowing in each of two infinitely long, straight parallel conductors which have negligible areas of cross – section and are 1m apart in a vacuum, causes each conducts to exert a force of $2x10^{-7}$ N on each mete of the other.

WORKED EXAMPLES

1. The plane of a circular coil is horizontal it has 20 turns each of 8cm radius A current of 1A flows through it which appears to be clockwise from a point vertically above it. Find the Magnitude of the magnetic field at the centre of the coil.

Solution

The magnitude of the magnetic field at the centre of the coil carrying current is given by,

$$B = \frac{\mu_0 NI}{2r}$$

$$\mu_0 = 4\pi \times 10^{-7} TA^{-1}$$

$$N = 20$$

$$I = 1A$$

$$r = 8cm = 8 \times 10^{-2} m$$

$$B = 1.57 \times 10^{-4} T$$

As the currents appears to be clockwise from appoint vertically above the coil the direction of the field will be vertically downward (By R.H.G.R)

2. A wire placed along the South-North direction carries currents of 5A from South to North. Find the magnetic field due to a 1cm piece of wire at a point 200cm North-East from the place.

Solution

$$I = 5A$$

$$dl = 1cm = 1 \times 10^{-2}m$$

$$r = 200cm = 200 \times 10^{-2}m$$

$$dB = \frac{\mu_0}{4\pi} \cdot \frac{Idl \sin\theta}{r^2}$$

$$dB = \frac{10^{-7} \times 5 \times 0.01 \times \sin 45°}{2^2}$$

$$dB = 8.8 \times 10^{-10}\mathrm{T}$$

By RHGR, The field is vertically vertical downwards

3. A coil of radius 10cm and having 20 turns carries a current of 12A in a clockwise direction when seen from east. The coil is in North – South plane. Find the magnetic field at the centre of the coil.

Solution

The magnitude of the magnetic field at the centre of the coil

$$B = \frac{\mu_0 NI}{2r}$$

$$\mu_0 = 4\pi \times 10^{-7}, r = 10 \times 10^{-2}\mathrm{m}$$

$$N = 20 \text{ turns}$$

$$B = \frac{4\pi \times 20 \times 12}{2 \times 10 \times 10^{-2}}$$

The electron of hydrogen atom moves along a circular path of radius 0.5×10^{-10} with the uniform speed of 4×10^{6} m/s. Calculate the magnetic field produced by electron at the centre (e= 1.6×10^{-9}c)

Number the revolution made by the electron in 1 second is

$$f = \frac{V}{2\pi r}$$

$$f = \frac{4 \times 10^{6}}{2\pi \times 0.5 \times 10^{-10}}$$

$$f = 1.27 \times 10^{16} rev/s$$

$$\text{Current} = \frac{Fe}{t}$$

$$I = \frac{1.27 \times 10^{16} \times 1.6 \times 10^{-19}}{1s}$$

$$I = 2.04 \times 10^{-3} A$$

Magnetic field produced by the electron at the centre is

$$B = \frac{\mu_0 NI}{2r}$$

$$B = \frac{(4\pi \times 10^{-7}) \times 2.04 \times 10^{-3}}{2 \times (0.5 \times 10^{-10})}$$

$$B = 24.6\text{T}$$

5. A circular coil of 100 turns has a radius of 10cm and carries a current of 5A Determine the magnetic field

(i) At the centre of the coil
(ii) At a point on the axis of the coil at a distance of 5cm from the centre of the coil.

Solution

(i) Magnetic field at the centre of the coil is

$$B = \frac{\mu_0 I}{2r}$$

μ_0 = 4^{π} x 10^{-7} TA^{-1}

N = 100 turns

I = 5A

r = $10x10^{-2}$m

B = $\frac{4^{\pi} \text{ x } 10^{-7} \text{x } 100 \text{ x S}}{2 \text{ X } 0.1}$

B= 3.14 $X10^{-3}$ T

∴ The magnetic field of the centre of the coil = 3.14 $X10^{-3}$ T

(ii) Magnetic field on the axis of the coil at a distance X from the centre is

$$B = \frac{\mu_0 N I r^2}{2(r^2 + x^2)^{3/2}}$$

μ_0 = 4^{π} x 10^{-7} TA^{-1}
N = 100 turns
I = 5A
r = 10 x 10^{-2}
x = 0.05m

$$B = \frac{4\pi \times 10^{-7} \times 100 \times (10 \times 10^{-2})^2 \times 5}{2((10 \times 10^2)^2 + (0.05)^2)^{3/2}}$$

$$\mathrm{B} = 5 \times 10^{-8}\mathrm{T}$$

6. An electric current I is flowing in a circular wire of radius at what dose from the centre on the axis of circular wire will the magnetic field be 1/8th of its value at the centre?

Solution

Magnetic field B at the centre of the circular coil is

$$B = \frac{\mu_0 I}{2a}$$

Suppose at a distance X from the centre on the axis of the circular coil the magnetic field is $\frac{B}{8}$

$$\frac{B}{8} = \frac{\mu_0 I a^2}{2(a^2 + x^2)^{3/2}}$$

$$\frac{\mu_0 I}{16a} = \frac{\mu_0 I a^2}{2(a^2 + x^2)^{3/2}}$$

$$\frac{1}{8a} = \frac{a^2}{(a^2 + x^2)^{3/2}}$$

$$(a^2 + x^2)^{3/2} = 8a^3$$

$$a^2 + x^2 = 4a^2$$

$$x^2 = 3a^2$$

$$x = \sqrt{3a^2}$$

$$x = \sqrt{3} \cdot a$$

7. In Bohr's model of hydrogen atom the electron circulates around nucleus on a path of radius 0.51Å at a frequency of 6.8x10^{15} is rev/second calculate the magnetic field induction at the centre of the orbit.

Solution

The circulating electron is equivalent to circular current loop carrying current I given by

$$I = \frac{dQ}{dt} = \frac{e}{1/f}$$

$$I = ef$$

I = 1.6 $\times 10^{-19} \times 6.8 \times 10^{15}$

I = 1.1$\times 10^{-3}$ A

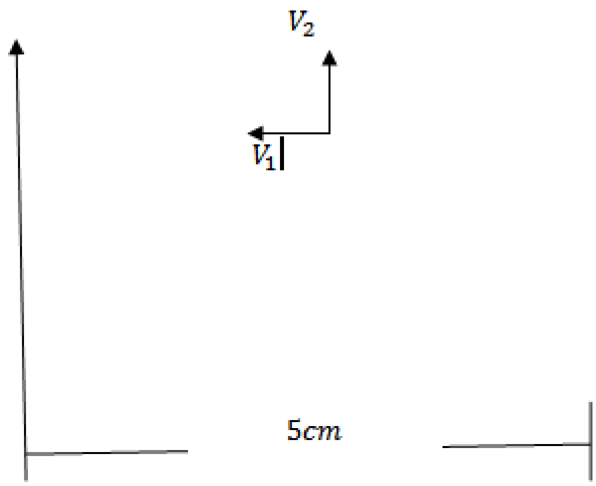
Magnetic field at the centre due to this current is

$$B_{centre} = \frac{\mu_0 I}{2r}$$

$$= \frac{(4\pi \times 10^{-7}) \cdot (1.1 \times 10^{-3})}{2 \times (0.51 \times 10^{-10})}$$

$\boldsymbol{B_{centre}}$ **= 14T**

8. A long straight wire carries a current of 50A. An electron moving at 10^7ms is 5cm from the wire



Find the Magnetic field acting on the electron velocity is directed

(i) Towards the wire

(ii) Parallel to the wire

(iii) Perpendicular to the directions defined by I and ii

Solution

The magnetic field produced by current carrying long wire at a distance r

$$B = \frac{\mu_0 I}{2\pi r}$$

$$B = \frac{4\pi \times 10^{-7}}{2\pi} \cdot \frac{50}{5 \times 10^{-2}}$$

$$\boldsymbol{B = 2 \times 10^{-4} T}$$

The field is directed downward perpendicular to the plane of the paper

(i) The velocity V_1 is towards the wire. The angle between V_I and B is 90^0 force on electron

F= $BQV^{\sin\theta}$

$F = 2x\ 10^{-4}\ x\ 1.6x10^{-19}x10^{7}x\ Sin\ 90^{0}$
$F = 3.2\ x\ 10^{-16}$ N

(ii) When the electron is moving is moving parallel to the wire ,angle between V2 and B is again 90Ëš Therefore, force is again **$3.2x10^{-16}$N**

(iii) When the electron is moving perpendicular to the directions defined by (i) and (ii) the angle between V and B is O

F = O

9. A solenoid has a length of 1 .23 m and inner diameter 4cm it has five layers of windings of 850 turns each and carries a current of 5.57A. what is the magnitude of the magnetic field at the centre of the solenoid

Solution

The magnitude of the magnetic field at the centre of a solenoid is given by
$\boldsymbol{B = \mu_0 nI}$

But

$$n = \frac{N}{L} = \frac{5 \times 850}{1.23} = 3455.3$$

$$\mathrm{B} = 4\pi \times 10^{-7} \times 3455.3 \times 5.557$$

$$\mathrm{B} = 24.2 \times 10^{-3}\mathrm{T}$$

10. A to void has a core (non - ferromagnetic) of inner radius 20cm and over radius 25cm around which 1500 turns of a wire are wound. If current in the wire is 2A

Calculate the magnetic field

(i) Inside the to void

(ii) Outside the to void

Solution

$$Mean\ radius = \frac{20+25}{2}$$

$$r = 22.5cm = 22.5 \times 10^{-2}m$$

$$l = 2\pi r$$

$$mean\ length\ l = 2\pi r$$

$$= 2\pi \times 22.5 \times 10^{-2}$$

$$= 1.413\text{m}$$

(i) The magnitude of the magnetic field inside the toroid is given by

$$\boldsymbol{B = \mu_0 nI}$$

$$\boldsymbol{B = 4\pi \times 10^{-7} \times \frac{1500}{1{\cdot}413}} \times 2$$

B = 0.003T

(ii)The magnetic field outside the toroid is Zero. It is all inside the toroid.

11. A solenoid 1.5m long and 4cm in diameter possess 10 turns/cm. A current of 5A is flowing through it. Calculate the magnetic induction

(i) Inside and

(ii) At one end on the axis of the solenoid

Solution

n = $\frac{N}{l}$ = 10turns/cm = $10^3\ \text{turns/cm}$

(i) Inside the solenoid , the magnetic induction is given by

$$B = \mu_0 nI$$

$$B = 4\pi \times 10^{-7} \times 10^3 \times 5$$

$$B = \pi \times 10^{-3} T$$

(ii) At the end of the solenoid the magnetic induction is given by

$$B = \frac{\mu_0 nI}{2} = \frac{2\pi \times 10^{-3}}{2}$$

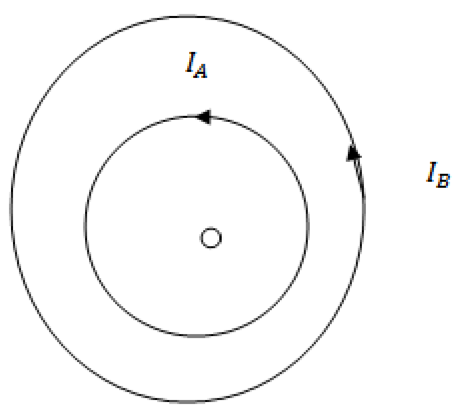
$$\boldsymbol{B = \pi \times 10^{-3} T}$$

12. (a) How will the magnetic field intensity at the centre of a circular loop carrying current change, if the current through the coil is doubled and the radius of the coil is halved?

(b) A long wire first bent in to a circular coil of one turn and then into a circular

coil of smaller radius having n turns, if the same current passes in both the cases, find the ratio of magnetic fields produced at the centers in the two cases.

(c) A and B are two concentric coils of centre O and carry currents I_A and I_B as shown in figure



If the ratio of their radii is 1:2 and ratio of flux densities at O due to A and B is 1:3, find the value of $\frac{I_B}{I_A}$

Solution

(a) Magnetic field at the centre of circular coil

$$B = \frac{\mu_0 nI}{2r}$$

$$B \propto \frac{I}{r}$$

$$B = \frac{KI}{r}$$

$$\frac{Br}{I} = k$$

$$\frac{B_1 r_1}{I_1} = \frac{B_2 r_2}{I_2}$$

$$\frac{B_2}{B_1} = \frac{I_2}{I_1} \cdot \frac{r_1}{r_2}$$

$$I_1 = I$$

$$r_1 = r$$

$$I_2 = 2I$$

$$r_2 = \frac{r}{2}$$

$$\frac{B_2}{B_1} = 4$$

$$B_2 = 4B_1$$

(b)Suppose r is the radius of one turn coil and the r^1 is the radius of n-turn coil. Then

$$N \times 2\pi r^1 = 2\pi r$$

$$r^1 = \frac{r}{N}$$

First case

$B_1 = \frac{\mu_0 I}{2r}$

Second case

$B_2 = \frac{\mu_0 NI}{2r^1}$

$$\frac{B_2}{B_1} = \frac{\mu_0 NI}{2(r/_N)} \cdot \frac{2r}{\mu_0 I}$$

$$\frac{B_2}{B_1} = N^2$$

Solution

C. Magnetic field at the centre of circular coil

$$B = \frac{\mu_0 I}{2r}$$

$$B \propto \frac{I}{r}$$

$$\frac{Br}{I} = k$$

$$\frac{B_A r_A}{I_A} = \frac{B_B r_B}{I_B}$$

$$\frac{I_A}{I_B} = \frac{B_A r_A}{B_B r_B}$$

$$\frac{I_A}{I_B} = \frac{1}{3} \times \frac{1}{2}$$

$$\frac{I_A}{I_B} = \frac{1}{6}$$

13. A helium nucleus makes a full rotation in a circle of radius 0.8m in two seconds. Find the value of magnetic field at the centre of the circle.

Solution

The charge on helium nucleus

Q= $+2$e

Q= $+2\ x$ 1.6 X10^{-19}c

Current produced I = $\frac{Q}{t}$

I = $\frac{2 \times 1.6 \times 10^{-19}}{2}$

I =1.6 x10^{-19}A

Magnetic field at the centre of the circle orbit of the helium is,

$$B = \frac{\mu_0 I}{2r}$$

$$B= \frac{2\pi\ x10^{-7}x1.6x\ 10-19}{2x0.8}$$

B= 1.256 x 10^{-25}T

14. A soft Iron ring has a mean diameter of 0.20m and an area of cross section 5x$10^{-4}$$m^2$ it is uniformly wound with 2000turns carrying a current of 2A and the magnetic flux in the iron is 8x 10^{-3}Wb. What is the relative permeability of iron?

Solution

Length of ring l

l = 2πr

l = 2πr x 0.10m

Number of turns per unit length n

n = $\frac{N}{l}$ = $\frac{2000}{2\pi \times 0.1}$

If M is the absolute permeability of iron, then magnetic flux density of iron ring is

B = $\mu_0 nI$

B = $\mu_0 \times \frac{2000}{2\pi \times 0.10} \times 2$

$$B = \frac{\mu_0}{\pi} \times 2 \times 10^4 \, {}^{Wb}/_{m^2}$$

Magnetic flux $\Phi = BA$

$$\frac{\mu}{\pi} \times 2 \times 10^4 \times 5 \times 10^{-4}$$

$$\Phi = \frac{10\mu}{\pi}$$

$$8 \times 10^{-3} = \frac{10\mu}{\pi}$$

Magnetic flux Φ = BA

$$\mu = 8\pi \times 10^{-4} H/M$$

Relative permeability of Iron μ_r

$$\mu_r = \frac{\mu}{\mu_0} = \frac{8\pi \times 10^{-4}}{4\pi \times 10^{-7}}$$

$$\mu_r = 2000$$

15. Two flat circular coils are made of two identical wires each of length 20cm one coil has number of turns 4 and the other 2. If the some current flows though the wire in which will magnetic field at the centre will be greater?

Solution

For the first coil

$$B_1 = \frac{\mu_0 n_1 I}{2r_1}$$

For second coil

$$B_2 = \frac{\mu_0 NI}{2r_2}$$

$$\frac{B_1}{B_2} = \frac{n_1}{n_2} \cdot \frac{r_2}{r_1}$$

$$Length\ of\ the\ wire\ l = n_1; 2\pi r_1 = n_2 2\pi r_2$$

$$n_1 r_1 = n_2 r_2$$

$$\frac{n_1}{n_2} = \frac{r_2}{r_1}$$

$$\frac{B_1}{B_2} = \frac{n_1}{n_2} \cdot \frac{n_1}{n_2}$$

$$\frac{B_1}{B_2} = \left(\frac{n_1}{n_2}\right)^2$$

$$\frac{B_1}{B_2} = \left(\frac{4}{2}\right)^2$$

$$\frac{B_1}{B_2} = 4$$

$\therefore$ Therefore, magnetic field will be greater in coil with 4 turns

16. A plat circular coil of 120 turns has a radius of 18cm and carries currents of 3A. What is the magnitude of magnetic field at a point on the axis of the coil at a distance from the centre equal to the radius of the coil?

Solution

Number of turns n = 120

Radius of the coil r = 0.18 m

Axial distance x = 0.18m

Current in coil I = 3A

$$B = \frac{\mu_0 N I r^2}{2(r^2 + x^2)^{3/2}}$$

$$B = \frac{(4\pi \times 10^{-7}) \times 120 \times 3 \times 0.18^2}{2(0.18^2 + 0.18^2)^{3/2}}$$

B= 4.4 x 10-4T

17. A current of 5A is flowing upward in a long vertical wire. This wire is placed in a uniform northward magnetic field of 0.02T. How much force and in which direction will this field exert on 0.06 length of the wire?

Solution

$F = BIL\sin\theta$

B = 0.02T

I = 5A

L = 0.06

$\theta = 90^0$

F= 0.02 X 5 X 0.06Sin900

F = 0.006N

By Fleming's Left hand rules the force is directed towards West

18. A straight wire of mass 200g and length 1.5m carries a current of 2A. It is suspend in mind air by a uniform horizontal magnetic field B. What is the magnitude of the magnetic field?

solution

M = 200 X 10^{-3} kg

I = 2A

l = 1.5m

B =?

F=BIL

Mg = BIL

$B = \frac{Mg}{IL} = \frac{200 \times 10^{-3} \times 9.8}{1.5 \times 2}$

B = 0.65T

19. Two long horizontal wires are kept parallel at a distance of 0.2cm apart in a vertical plane . both the wires have equal currents in the same direction the lower wire has a mass of 0.05kg/m if the lower wire appears weightless what is the current in each wire ?

Solution

Let I amperes be the current in each wire the lower wire is acted upon by two forces.

$$\text{Upward force}/\text{meter length} = \frac{\mu_0}{2\pi}.\frac{I^2}{r}$$

$$\frac{F}{L} = \frac{4\pi \times 10^{-7}}{2\pi}.\frac{I^2}{0.2 \times 10^{-2}}$$

$$\frac{F}{L} = \frac{2 \times 10^{-7} I^2}{0.2 \times 10^{-2}}$$

$$\frac{F}{L} = 10^{-4} I^2$$

$$\text{Downwars force}/_{\text{meter length}} = mg$$

$$= 0.05 \times 9.8$$

$$= 0.49\text{N}$$

Since the lower wire appears weightless the two forces were equal over 1m length of the wire

$10^{-4}I^2 = 0.49$

$$I^2 = \frac{0.49}{10^{-4}}$$

$$I = \sqrt{0.49 \times 10^4}$$

$$I = 70A$$

20. The horizontal component of the earth magnetic field at a certain place is 3 x 10^{-5} and the direction of the field is from the geographic south to the geographic North A very long straight conductor is carrying a steady current of 1A. what is the force per unit length on it when it is placed on a horizontal table and the direction of the current is
   (a) East to West
   (b) South to North

Solution

(a) When current is flowing from east to west $\theta = 90^0$

Force on the conductor per unit

$$\frac{F}{L} = \frac{BIL\sin\theta}{L}$$

$$\frac{F}{L} = BIL\sin\theta$$

$$\frac{F}{L} = 3 \times 10^{-5} \times 1 \times Sin\,90$$

$$\frac{F}{L} = 3 \times 10^{-5}\, {}^{N}/_{m}\, (Downwards)$$

(b) When current is flowing from south to north $\theta = 0^0$

Force on the conductor per unit length

$$\frac{F}{L} = \frac{BIL\sin\theta}{L}$$

$$\frac{F}{L} = BI\sin 0$$

$$\frac{F}{L} = 0$$

21. A horizontal straight wire 5cm long of mass $1.2gm^{-1}$ placed perpendicular to a uniform magnetic field of 0.6T if resistance of the wire is $3.85cm^{-1}$ calculate the P.d that has to be applied between the ends of the wire to make it just self supporting

Solution

The current (i) in the wire is to be in such a direction that magnetic force acts on it vertically upward. To make the wire self supporting its weight should be equal to the upward magnetic force.

$$F = BIL \sin 90$$

$$F = BIL$$

$$Mg = BIL$$

$$I = \frac{Mg}{BL}$$

$$I = \frac{1.2 \times 10^{-3} \times l \times 9.8}{0.6 \times l}$$

$$I = 19.6 \times 10^{-3} A$$

Resistance of the wire

R = 0.05 x 3.8

= 0.19Ω

Required P. (I) V = IR

V = $19.6X10^{-3}$ X 0.19

V = 3.7 $X\ 10^{-3}$V

22. A conductor of length 2m carrying current of 2A is held parallel to an infinitely long conductor carrying current of 10A at a distance of 100mm.
find the force on small conductor

Solution

$I_I = 2A$

$I_2 = 10A$

$r = 100 \times 10^{-3} m$

$l = 2m$

Force on unit length of short conductor by the long conductor is give by

$$f = \frac{\mu_0}{2\pi} \cdot \frac{I_1 I_2}{r}$$

Force on length l = 2m of short conductor by the long conductor is

$$F = f \times l$$

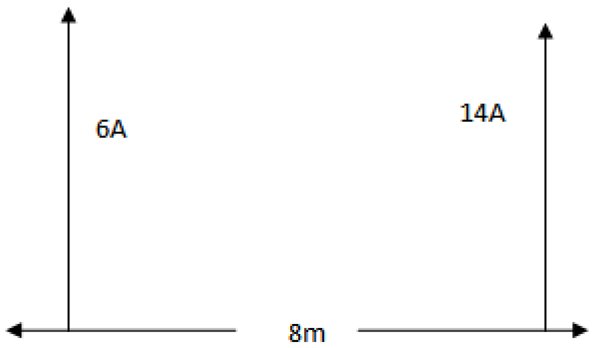
$$F = \frac{\mu_0}{2\pi} \cdot \frac{I_1 I_2 \cdot l}{r}$$

$$F = \frac{4\pi \times 10^{-7}}{2\pi} \cdot \frac{2 \times 10}{0.1} \times 2$$

$$F = 8 \times 10^{-5} N$$

The force will be attractive if the direction of current is the same in two conduction and it will be repulsive if the conductors carry current in the opposite directions.

23. In the figure below, determine the position between two wire which experience zero resultant force due to charge Q placed at that point



Solution

The force unit length acting in each wire of the parallel wire is given by

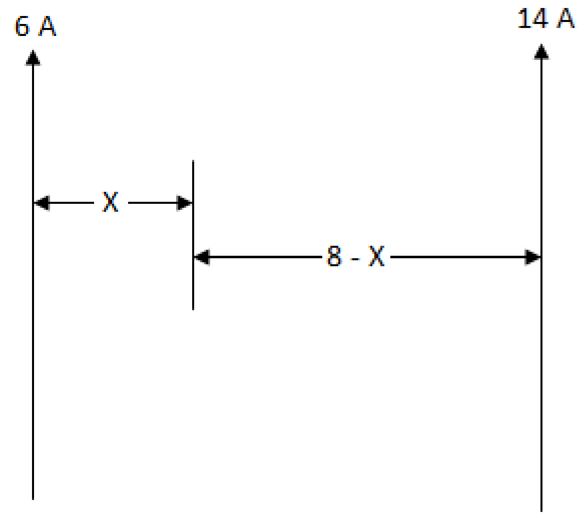
$$\frac{F}{L} = \frac{\mu_0}{2\pi} \cdot \frac{I_1 I_2}{r}$$

Let $\frac{F_1}{L_1}$ be the force per unit length in the wire carrying a current of 14A

Since F_1 and F_2 have the same magnitude but they are acting in opposite direction for resultant force to be zero

$$\frac{F_1}{L_1} = \frac{F_2}{L_2}$$

Assume that the charge Q is placed at a distance X from the wire carrying the



$$\frac{F_1}{L_1} = \frac{\mu_0 \times 6 \times 14}{2\pi x}$$

$$\frac{F_1}{L_1} = \frac{42\mu_0}{\pi x} \dots\dots\dots\dots\dots\dots.(1)$$

$$\frac{F_1}{F_2} = \frac{\mu_0 \times 6 \times 14}{2\pi(8-x)}$$

$$\frac{F_2}{L_2} = \frac{42\mu_0}{\pi(8-x)} \dots\dots\dots\dots\dots\dots..(ii)$$

$$\frac{F_1}{L_1} = \frac{F_2}{L_2}$$

$$\frac{42\mu_0}{\pi x} = \frac{42\mu_0}{\pi(8-x)}$$

$$\frac{1}{x} = \frac{1}{8-x}$$

$$x = 8 - x$$

$$2x = 8$$

$$x = 4m$$

The charge Q is placed 4m from the either wire.

CLASSIFICATION OF MAGNETIC MATERIALS

All substances are affected by magnetic field some attain weak magnetic properties and some acquire strong magnetic properties and some acquire strong magnetic properties.

The magnetic properties of the substances are explained on the basis of modern atomic theory.

The atoms that make up any substance contain electrons that orbit around the central nucleus.

Since the electrons are charged they constitute an electric current and therefore produce magnetic field .

Thus an atom behave as a magnetic dipole and possesses magnetic dipole moment.

The magnetic properties of a substance depend upon the magnetic moments of its atoms.

IMPORTANT TERMS USED IN MAGNETISM

The following terms are used in describing the magnetic properties of the materials:

(i) **Magnetic flux density (B)**

Is a measure of the number of magnetic field lines passing per unit area of the material.

The greater the number of magnetic field lines passing per unit are of the material

(ii) **Magnetic permeability**

Is a measure of its conductivity for magnetic field lines

The greater the permeability of the material the greater is its conductivity for the magnetic field line and vice versa

Since magnetic field strength B is the magnetic field lines passing per unit area of the material, it is a measure of magnetic permeability of the material.

Suppose magnetic flux density in air or vacuum is B_o. If vacuum/air is replaced by a material, suppose the magnetic flux density in the material becomes B

Then ratio **B/B_o** called the **relative permeability** μ_r. of the material.

(i) Relative permeability μ_r.
Is the ratio of magnetic flux density B in that material to the magnetic flux density B_o that would be if the material were replaced by vacuum/ air .

$$\mu_0 = \frac{B_0}{B_0} = 1$$

Clearly μ_r is a pure number and its value per vacuum/air is 1

$$\mu_r = \frac{B_0}{B_0} = 1$$

Relative permeability of a material may also be defined as the ratio of absolute permeability μ of the material to absolute permeability μ_o of vacuum/air.

$$\mu_0 = \frac{\mu}{\mu_0} = 1$$

(ii) Magnetizing force/ Magnetic intensity $\vec{H}$

Is the number of ampere - turns flowing per unit length of the toroid.

The SI Unit of magnetizing force H is Ampere – turns per meter (AT/m)

Consider a toroid with n turns per unit length carrying a current I. if the absolute permeability of toroid material is M, then magnetic flux density B in the material is

$B = \mu nI$

$B = \mu_0 \mu_r nI$

The quantity nI is called magnetizing force or magnetic intensity

Therefore, the ratio $\frac{B}{H}$ in a material I is from

$B = \mu H$; $\mu = \mu_0 \mu_r$; B=$\mu_0 \mu_r H$

$B_0 = \mu_0 H$

Thus if the some magnetizing force is applied to two identical air cored and iron cored toroid, then magnetic flux density produced inside the toroid is

$B_0 = \mu_0 \mu_r H$

(iii) Intensity of magnetization ($\vec{I}$) is the magnetic moment developed per unit volume of the material.

When a magnetic material is subjected to a magnetizing force, the material is magnetized

Intensity of magnetization is the measure of the extent to which the material is a magnetized and depends upon the nature of the material

$$\vec{I} = \frac{\vec{M}}{V}$$

where:

$\vec{M}$ = magnetic moment developed in the material

V= volume of the material

If **m** is the pole strength developed,

a is the area of X – section of the material and 2l is the magnetic length. Then

$$I = \frac{m \times 21}{a \times 21}$$

$$I = \frac{m}{a}$$

Hence Intensity of magnetization of a material may be defined as the pole strength developed per unit area of cross – section of the material.

Thus the SI unit of I is Am^{-1} which is the same as the SI unit of H

Magnetic susceptibility X_m is the ratio of intensity of magnetic on I developed in the material to the applied magnetizing force H

It is represented by $X_m = \frac{1}{H}$

The magnetic susceptibility of a material indicates how easily the material can be magnetized.

The unit of I is the same as that of H so that X_m is a number

Since I is magnetic moment per volume X_m is also called volume susceptibility of the material .

Consider a current carrying toroid having core material of relative permeability μ_r

The total magnetic flux density B in the material is given by

$B = B_0 + B_m$

Where

B_o = magnetic flux density due to current in the coils.

B_m = magnetic flux density due to the material (Magnetization of the material)

$B_0 = \mu_0 H$ ....(i)

$B_m = \mu_0 I$ ....(ii)

Here I is the intensity of magnetization induced in the toroid material

B = $B_0 + B_M$

$B = \mu_0 H + \mu_0 I$

$B = \mu_0 (H + I)$

Now,

$$\boldsymbol{X_m = \frac{I}{H}}$$

$$\mathbf{I} = \boldsymbol{X_m H}$$

$$\mathbf{B} = \boldsymbol{\mu_0 (H + X_m H) = \mu_0 H (1 + X_m)}$$

$$\mathbf{B} = \boldsymbol{\mu H = \mu_0 \mu H}$$

$$\boldsymbol{\therefore \mu_0 \mu_r H = \mu_0 H (1 + X_m)}$$

Or

$$\boldsymbol{\mu_r = 1 + X_m} \quad \text{..............................}\mathbf{(iii)}$$

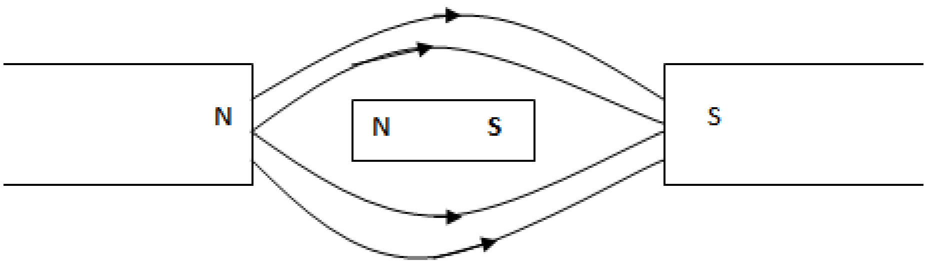
Equation (iii) give the relation between relative permeability (μ_r) and magnetic susceptibility (X_m).

CLASSIFICATION OF MAGNETIC MATERIALS

All materials or substances are affected by the external magnetic field. Some attain weak magnetic properties and acquire strong magnetic properties.

On the basis of their behavior in external magnetic field , the various substance classified into the following three categories
(i)Diamagnetic materials
(ii)Paramagnetic materials
(iii)Ferromagnetic materials

(i) DIAMAGNETIC MATERIAL
When a diamagnetic substance is placed in a magnetic field in the magnetic field lines prefer to passs through the surrounding air rather than through the substance.



Diamagnetic materials are materials which can not be affected by the magnetic field.

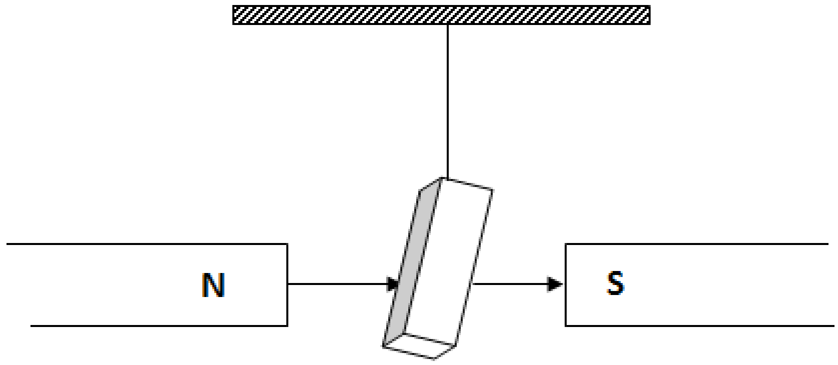
They are repelled by magnetic field e.g. lead, silver, copper, zinc, water, gold bismuth etc.

These substances when placed in a magnetic field are weakly magnetized in a direction opposite to that of the applied field.

PROPERTIES OF DIAMAGNETIC MATERIALS

1. A diamagnetic substance is feebly repelled by a strong magnet.
2. The magnetic susceptibility (X_m) of a diamagnetic substance has a small negative value.
3. The relative permeability (M_r) of a diamagnetic substances is slightly less than 1
4. When a rod of diamagnetic substances is suspended freely in a uniform magnetic field, the rod comes to rest with its axis perpendicular to the direction of the applied field.

   See figure below



This gives the relation between relative permeability μ_r and magnetic susceptibility X_m of the material.

(ii)PARAMAGNETIC MATERIALS

Are materials which when placed in a magnetic field are weakly magnetized in the direction of the applied field

The paramagnetic substances include the Aluminum antimony , copper sulphate, Crown grass etc

Since the weak induced magnetic field is in the direction of the applied field, the resultant magnetic field in the paramagnetic substance is slightly more than the external field

Hence the magnetic susceptibility of a paramagnetic substance is positive having

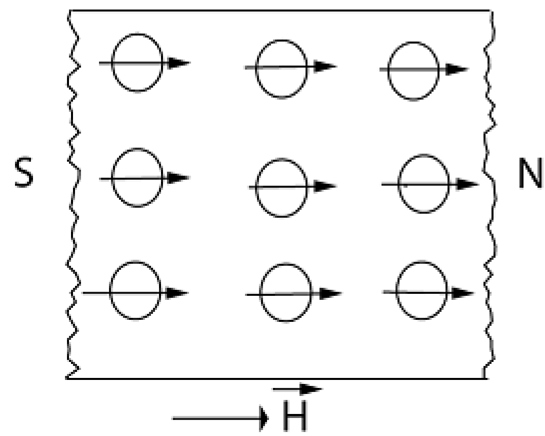
It clear that the relative permeability μ_r for such substances will be slightly more than 1

$\mu_r = 1 + X_m$

Paramagnetic substance loses its magnetism as soon as the external magnetic field is removed

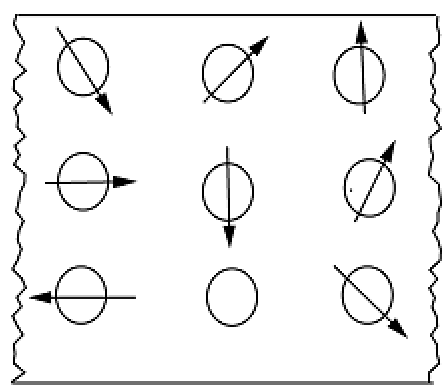
BEHAVIOR OF PARAMAGNETIC SUBSTANCES IN AN EXTERNAL MAGNETIC FIELD

When a paramagnetic substance is placed in an external magnetic field the dipoles are partially aligned in the direction of the applied field.



Therefore the substance is feebly magnetized in the direction of the applied magnetic field. This result into a weak attractive force on the substances.

In the absence of the external magnetic field the dipoles of the paramagnetic substances are randomly oriented and therefore the net magnetic moment of the substance is zero.



Hence the substance does not exhibit Para - magnetism

PROPERTIES OF PARAMAGNETIC SUBSTANCES

1. The relative permeability of a paramagnetic substance is always more than 1

$$\mu_r = \frac{B}{B_0}$$

The result field B inside a paramagnetic substance is more than the external field Bo
$B > B_0$

2. The magnetic susceptibility of the paramagnetic substance has small positive value

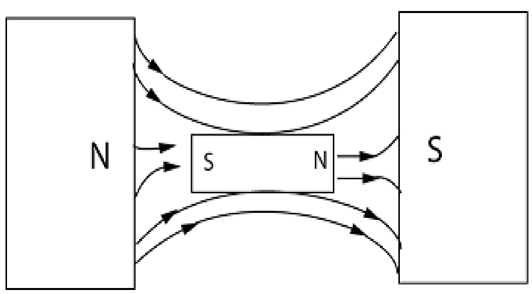
It is because $\mu_r = 1 + X_m$ and $\mu_r > 1$

3. The magnetic susceptibility of a paramagnetic material varies inversely as the absolute temperature

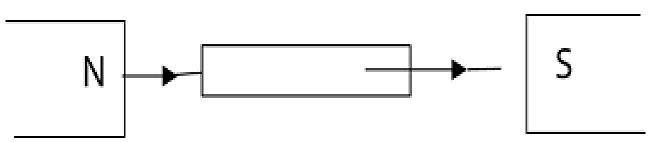
$$X_m \propto \frac{1}{T}$$

Paramagnetism is quite sensitive to temperature. The lower the temperature the stronger is the paramagnetism and vice versa

4. A paramagnetic substance is feebly attracted by the strong magnet. It is because a paramagnetic substance develops weak magnetization in the direction of the applied external magnetic field



5. When a paramagnetic substance is placed in a magnetic field, the magnetic field lines of force prefer to pass through the substance rather than through air.



Therefore the resultant field B inside the substance is more than the external field Bo

FERROMAGNETIC MATERIALS

Are the materials which when placed in a magnetic field are strongly magnetic in the direction of the applied field.

Ferromagnetic substances includes

- Iron
- Cobalt
- Nickel
- Fe_2O_3
- Gadolinium

Since the strong induced magnetic field is in the direction of the applied magnetic field, the resultant magnetic field inside the ferromagnetic substance is very large compared to external field

It is clear that ferromagnetism is very stronger form of magnetism. When external field (magnetic field) is removed some ferromagnetic substances retain magnetism

PROPERTIES OF FERROMAGNETIC SUBSTANCES

1. The relative permeability (μ_r) of the ferromagnetic substance is very large

Now

$$\mu_r = \frac{B}{B_0}$$

$$B \gg B_0$$

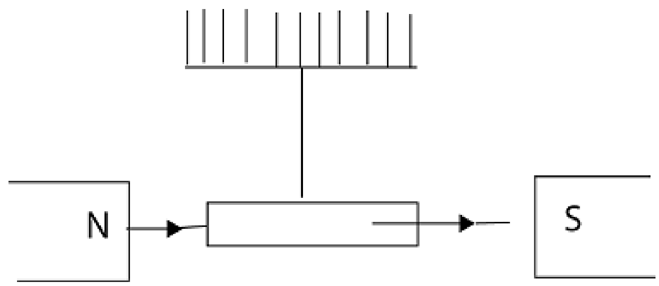
The resultant field B inside a ferromagnetic substance is very large as compared to the external filed Bo

2. The magnetic susceptibility (X_m) of a ferromagnetic substance is positive has a very high value.

It is because $\mu_r = 1 + X_m$ and $\mu_r \gg\gg 1$ for this reason, ferromagnetic substance can be magnetized easily and strongly.

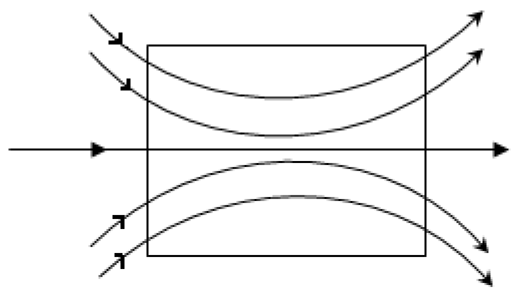
3. A ferromagnetic substance is strongly attracted by a magnet

4. When a rod of ferromagnetic substance is suspended in a uniform magnetic field, it quickly aligns itself in the direction of the field.



5. They retain their magnetization even when their magnetizing force is removed.

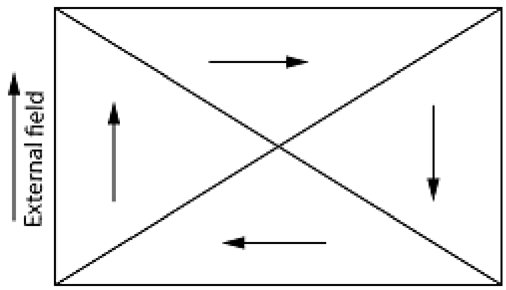
6. When a ferromagnetic substance is placed in a magnetic field the magnetic field lines tend to crowd into the substance



DOMAIN

Is the region of the space over which the magnetic dipole movements of the atoms are aligned in the same direction.

(i) In the absence of the external magnetic field the domain of the ferromagnetic materials are randomly oriented as shown below.



In other words, within the domain all the magnetic moments are aligned in the same direction but different domains are oriented randomly in different direction.

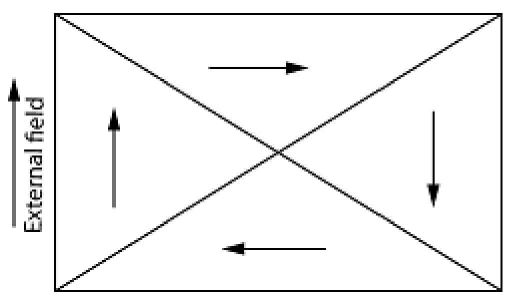
The result is that one domain cancels the effect of the other so that the net magnetic moment in the material is zero.

Therefore a ferromagnetic material does not exhibit magnetism in the normal state

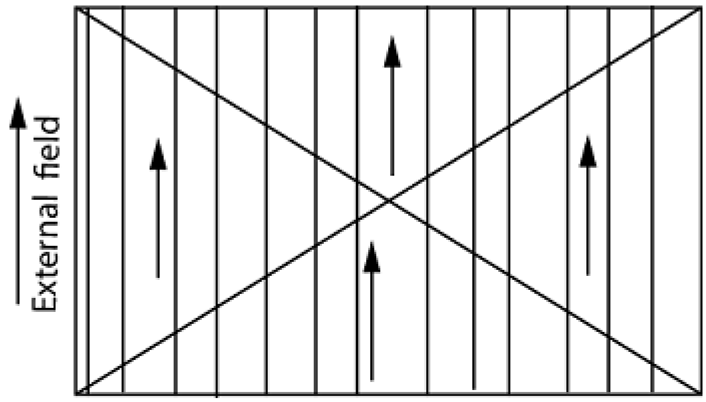
(ii) When a ferromagnetic substance is placed in an external magnetic field a net magnetic moment develops the substance.

This can occur in two ways

(a) By displacement of boundaries of the domains i.e. the domains that already happen to be aligned with the applied field may grow in size whereas those oriented opposite to the external field reduce in size.



(b) By the rotation of the domains i.e. the domains may rotate so that their magnetic moments are more or less aligned in the direction of the magnetic field.



The result is that there is net magnetic moment in the material in the direction of the applied field.

Since the degree of alignment is very large even for a small external magnetic field the magnetic field produced in ferromagnetic material is often much greater than the external field.

CURIE TEMPERATURE

Is the temperature at which the ferromagnetic substance becomes paramagnetic

It is also known as Curie point of the substance

Ferromagnetism decreases with the increases in temperature

When a ferromagnetic substance is heated magnetization decreases because random thermal motions tend to destroy the alignment of the domains

At sufficiently high temperature the ferromagnetic property of the substance suddenly disappears and the substance becomes paramagnetic.

In a ferromagnetic substance the atom appear to be grouped magnetically into what are called domains.

This occurs because the magnetic dipole moments of atoms of a paramagnetic substance exert strong force on their neighbor so that over a small region of space the moments are aligned with each other even with no external field.

Above Curie temperature these forces disappear and ferromagnetic substances become paramagnetic.

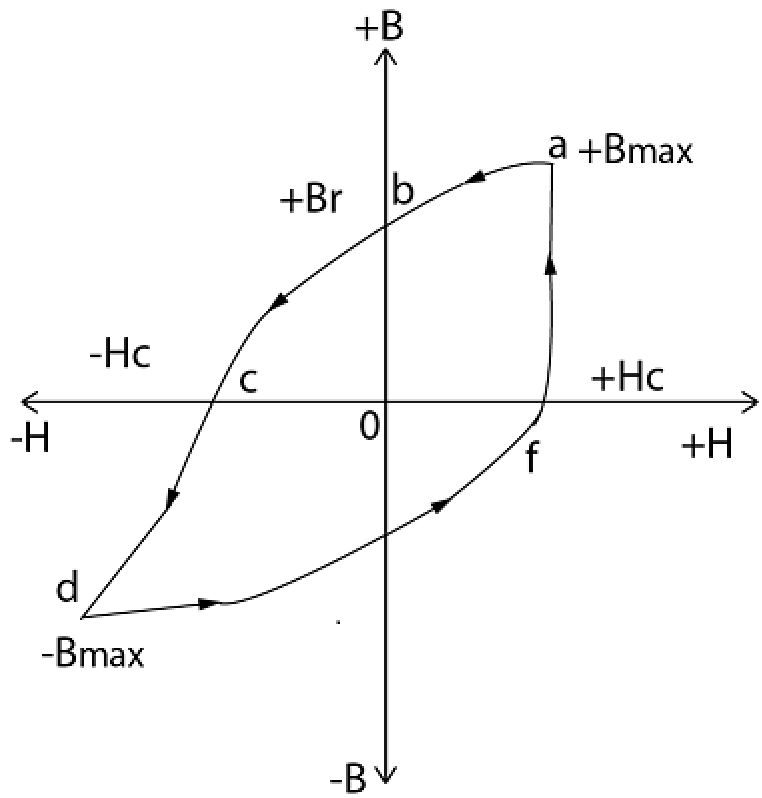
HYSTERESIS

Is the phenomenon of lagging of flux density (B) behind the magnetic force (H) in ferromagnetic materials subjected to cycles of magnetization.

When a ferromagnetic substance e.g. iron is subjected to cycle of magnetization (i.e. it is magnetized first in one direction and then in the other) it is found that flux density B in the materials lags behind the applied magnetizing force H.

This phenomenon is known as Hysteresis.

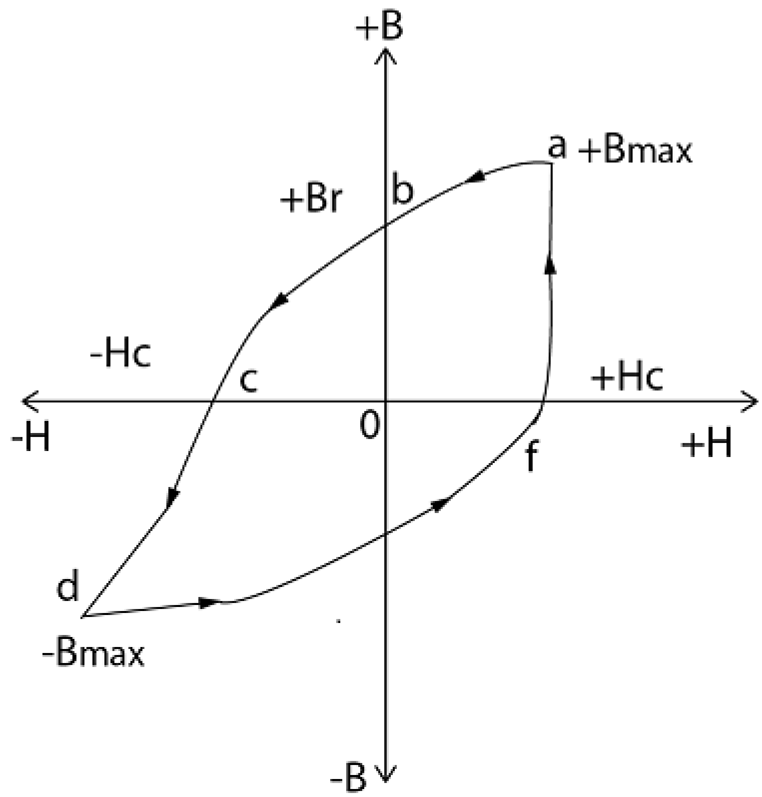
If a piece of ferromagnetic material is subjected to one cycle of magnetization the result B-H curve is a closed loop **a b c d e f a** called Hysteresis loop.



B always lags behind H, Thus at point b, H is zero but flux density B has a finite positive value ob similarly at point e, H is zero but flux density B has a finite negative value X .

HYSTERESIS LOOP

Consider an Iron cored toroid carrying current I



If N is the total number of turn and l the length of toroid, then magnetizing force is

$$H = \frac{NI}{l}$$

The value of H can be changed by varying current in the coil

Consider that when the Iron cored toroid is subjected to a cycle of magnetization the resultant B- H curve traces a loop **a b c d e f a** called hysteresis loop

(i) To start with the toroid is unmagnetised and its situation is represented by point O in graph

As H is increased (by increasing current I),B increases along OA and reaches its saturation value B_{max} at a this stage

(i) all the domains are aligned

(ii) If now H is gradually reduced by decreasing current in the toroid it is found that curve follows the path AB instead of AA

At point b, H = O but flux density in the material has a finite value of +Br called residual flux density

REMANENCE

Is the flux density left behind in the sample after the removal of the magnetizing force (H). It is also called Residual magnetism or retentively.

B lags behind H. This effect is called Hysteresis

(iii) In order to reduce flux density in the material to zero, it is necessary to apply H in the reverse direction

This can be done by reversing the current in the toroid

When H is gradually increased in the reverse direction the curve follows the path BC

At point C, B =O and H = -HC, the value of H needed to wipe out residual magnetism is called coercive force it.

COERCITIVITY OF THE SAMPLE

Is the value of reverse magnetizing force required to wipe out the residual magnetism in the sample

Now H is further increased in the reversed direction until point d is reached where the sample is saturated in the revision direction ($-B_{max}$).

If the H is now reduced to zero, point e is reached and the sample again retain magnetic flux density ($-B_r$)

The remaining part of the loop is obtained by increasing current to produce H in the original direction .

The hysteresis loop results became the domains do not become completely unaligned when H is made zero.

The area enclosed by the hysteresis loop represents loss in energy

This energy appears in the material as heat.

HYSTERESIS LOSS

This is the loss of energy in the form of heat when a ferromagnetic material is subjected to cycles of magnetization.

Hysteresis loss is present in all those electrical machines whose iron parts are subjected to cycles of magnetization.

The obvious effect of hysteresis loss is the rise in temperature of the machine.

HYSTERESIS LOOP

Is the loop traced by the resultant B-H curve when the Iron – cored toroid is subjected to a cycle of magnetization.

The shape and size of hysteresis loop largely depends upon the nature of the material

The choice of a ferromagnetic material per a particular application often depends upon the shape and size of the hysteresis loop

(i) The smaller the hysteresis loop area of a ferromagnetic material the smaller is the hysteresis loss

The hysteresis loop per silicon steel has a very small area.

For this reason, silicon steel is widely used per making transformer cores and rotating machines which are subjected to rapid reversals of magnetization

(ii) The hysteresis loop per hard steel indicates that this material has high retentivity and Coercivity.

Therefore hard steel is quite suitable for making permanent magnets.

But due to the large area of the loop there is a greater hysteresis loss

For this reason, hard steel is not suitable for the construction of electrical machines

(iii)The hysteresis loop for wrought iron shows that this material has fairly good residual magnetism and Coercivity.

Hence it is suitable for marking cores of electromagnets

APPLICATIONS OF FERROMAGNETIC MATERIALS

Ferromagnetic material (E.g. iron, steel nickel, cobalt etc) are widely used in a number of applications

The choice of ferromagnetic material for a particular for a particular application depends upon its magnetic properties such as

(i) Retentivity

(ii)Coercivity

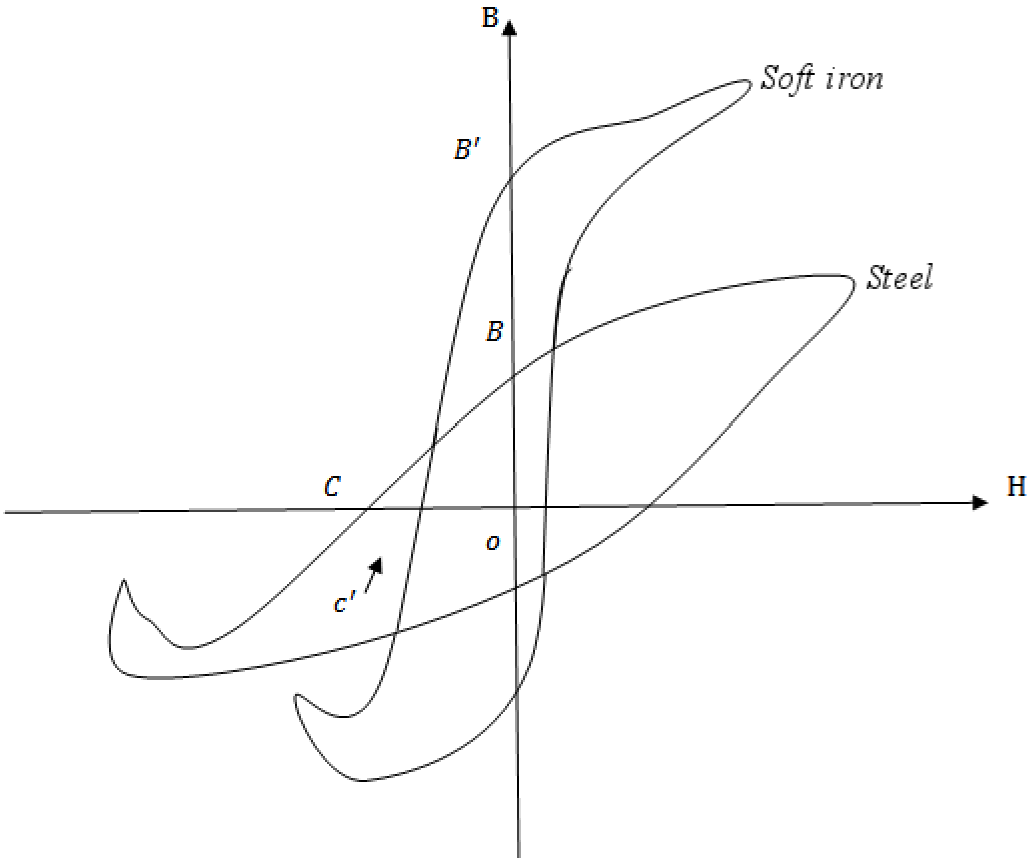
(iii) Area of the hysteresis loop

Ferromagnetic materials are classified as being either

(i) Soft (soft iron)

(ii) Hard (steel)

Figure below shows the hysteresis loop for soft and hard ferromagnetic materials



The table below gives the magnetic properties of hard and soft ferromagnetic materials

Magnetic property	Soft	Hard
Hysteresis loop	narrow	Large area
Retentivity	High	High
Coercivity	low	high

(i) PERMANENT MAGNETS

The permanent magnets are made for hard ferromagnetic materials (steel, cobalt, carbon steel)

Since these materials have high relativity the magnet is quite strong

Due to their high Coercivity, they are unlikely to be demagnetized by stray magnetic field

(ii) TEMPORARY MAGNETS / ELECTROMAGNETIC

The Electromagnets are made from soft ferromagnetic materials e.g. soft iron.

Since these materials have low coercively they can be easily demagnetized.

Due to high saturation flux density they make strong magnets

(iii) TRANSFORMER CORES

The transformer cores are made from soft ferromagnetic materials

When a transformer is in use, its core is taken through many cycles of magnetization

Energy is dissipated in the core in the form of heat during each cycle. The energy dissipated is known as hysteresis loss. And is proportional to the area of hysteresis loop

Since the soft Ferromagnetic materials have narrow hysteresis loop (smaller loop areas) they are used for making transformer cores.

WORKED EXAMPLE

1. (a) How does a permanent magnet attract an unmagnetised iron object?

(b) Show that the unit of magnetizing force is $Nm^{-2}T^{-1}$ or $Jm^{-1}wb^{-1}$

(C) Why is electromagnets made of soft iron?

(d) An Iron ring has a cross - sectional area of 400mm^2 and a mean diameter of 25cm. It is wound with 500trns. If the relative permeability of iron is 5000 find;

(i) The magnetizing force

(ii) The magnetic flux density set up in the ring.

The coil resistance is 474Ω and the supply voltage is 240V

Solution

(a) The magnet's field cause a slight alignment of the domains in the unmagnetised iron object so that the object becomes a temporary magnet with its north pole facing the south pole of the permanent magnet and viceversa. Therefore attraction results.

(b) The SI units of magnetizing (H) are Ampere/M(Am^{-I})

Now

$$H = \frac{B_0}{\mu_0} = \frac{F/_{QV}}{\mu_0}$$

$$H = \frac{F}{QV.\mu_0} = \frac{N}{C.Ms^{-1}.TMA^{-1}}$$

$$H = NM^{-2}T^{-1}$$

Also

$$H = NM^{-2}T^{-1}$$

$$H = \frac{N}{TM^2}$$

$$\therefore W_b = TM^2$$

$$H = NWb^{-1}$$

$$H = JM^{-1}Wb^{-1}$$

(c) The soft Iron has very small residual magnetism and coercive force. Therefore the material loses magnetism as soon as the magnetizing force is removed for this reason electromagnets are made of soft iron

(d) Current through the coil I

$$I = \frac{F}{R} = \frac{240}{474}$$

$$I = 0.506A$$

Mean length of the ring l

l = $2^{\pi r}$

l = $2^{\pi r}$ x (12.5 x $10^{-2)}$

l = 0.785m

(i) Magnetizing force H

$$H = nI$$

$$H = \frac{N}{L} \cdot I$$

$$H = \frac{500}{0.785} \times 0.506$$

$$H = 322\, AT/\mathrm{m}$$

(ii) Magnetic flux density B
From

$$H = \frac{B}{\mu}$$

$$B = \mu H$$

$$B = \mu_r \mu_0 H$$

$$B = 4\pi \times 10^{-7} \times 5000 \times 322$$

B = 2.02wb/m^2

2. (a) Diamagnetic is a property of a material. Discuss

(b) What is the magnetic susceptibility and permeability of a perfectly diamagnetic Substance?

(c) Why is Diamagnetism independent of temperature?

(d) The core of a toroid having 3000 turns has inner and outer radii of 11cm and 12cm respectively. The magnetic flux density in the core per a current of 0.70A is 2.5T

What is the relative permeability of the core?

Solution

(a) Diamagnetism is a natural reaction to the applied magnetic field. Therefore it is present in all materials but is weaker even than paramagnetism. As the result, diamagnetism is overwhelmed by paramagnetic and ferromagnetic effects in materials that display these other forms of magnetism

Solution

B = μ_0H + μ_0I

B = μ_0(H+I)

For perfectly diamagnetic substance

B = O

μ_0(H+I) = O

H +I = O
I = -H

Susceptibility

$X_m = \frac{I}{H}$

$X_m = \frac{-H}{H}$

$X_{m=} -1$

Also

μ_r = 1 + X_m

μ_0 = 1 – 1

μ_0 = O

(c) The induced magnetic moment in atoms of a diamagnetic substance is not affected by the thermal motion of the atoms. For this reason, diamagnetic is independent of temperature.

(d) **Solution**

Mean radius of toroid

$r = \frac{11+12}{2} = 11.5cm$

$r = 11.5 \times 10^{-2}m$

The magnetic flux density in the toroid is

$B = \mu H$

$B = \mu_r \mu_0 H$

$\mu_r = \dfrac{B}{\mu_0 n I}$

$n = \dfrac{3000}{2\pi \times 11.5 \times 10^{-2}}$

$n = 4.154 \times 10^3$

$\mu_r = \dfrac{2.5}{4\pi \times 10^{-7} \times 4.154 \times 10^3 \times 0.7}$

$\boldsymbol{\mu_r = 6.85 \times 10^2}$

3 .(a) (i) What is a non Magnetic material?

(ii) Can there be a material which is non -magnetic?

(b) What do you mean by the greater susceptibility of a material?

(c) An iron rod of $0.1m^2$ area of x-section is subjected to magnetic field of 1000AM^{-1}

Calculate its magnetic permeability. Given susceptibility of iron are 599.

(d) Which material is used to make permanent magnets and why?

Solution

(a) (i) A non –magnetic material is that which is not affected even by strong magnetic fields.

(ii) No, every material is at least diamagnetic.

(b) From $\boldsymbol{X_m = {}^{I}/_{H}}$

For a given H, $X_m \alpha\ I$ Thus the greater value of the susceptibility of the material the greater will be its intensity of magnetization i.e. more easily can be magnetized .Thus greater value of its susceptibility for iron means that it can be easily magnetized.

(c) **Solution**

$\mu_r = 1 + X_m$

But

$\mu = \mu_0\mu_r$

$\mu = \mu_0(1+X_m)$

$\mu = 4\pi \times 10^{-7}(1+599)$

$\mu = 7.54 \times 10^{-4} TA^{-4}M$

(d) Steel it is because has high coactivity. This ensures the stay of magnetism in steel for a longer period

4. (a) A toroid of mean circumference 50cm has 500 turns and carries a current of 0.15A

(i) Determine the magnetizing force and magnetic flux density if the toroid has an air core

(ii) Determine the magnetic flux density and intensity of magnetization if the core is filled with iron of relative permeability 5000

(b) Why do magnetic lines of force prefer to

(c) What is the SI unit of magnetic susceptibility?

Solution

(a) (i) given

$L = 50cm = 50 \times 10^{-2}$ m

N = 500 turns

I = 0.15A

Magnetizing force H

$\mathbf{H = nI}$

$= \frac{N}{L} \times I$

$= \frac{500}{50 \times 10^{-2}} \times 0.15$

$\mathbf{H = 150Am^{-1}}$

Magnetic flux density B

$\mathbf{B = \mu_0 H}$

$\mathbf{B = (4\pi \times 10^{-7}) \times 150}$

$\mathbf{B = 0.188 \times 10^{-3}T}$

(ii) Magnetic flux density

$B = \mu_0(H + I)$

$B = \mu_0 H + \mu_0 I$

$\mu_0 I = B - \mu_0 H$

$I = \frac{B}{\mu_0} - H$

$I = \frac{0.94}{4\pi \times 10^{-7}} - 150$

$I = 7.5 \times 10^5 Am^{-1}$

$\mathrm{B = \mu_0 \mu_r H}$

$\mathrm{B = 0.94T}$

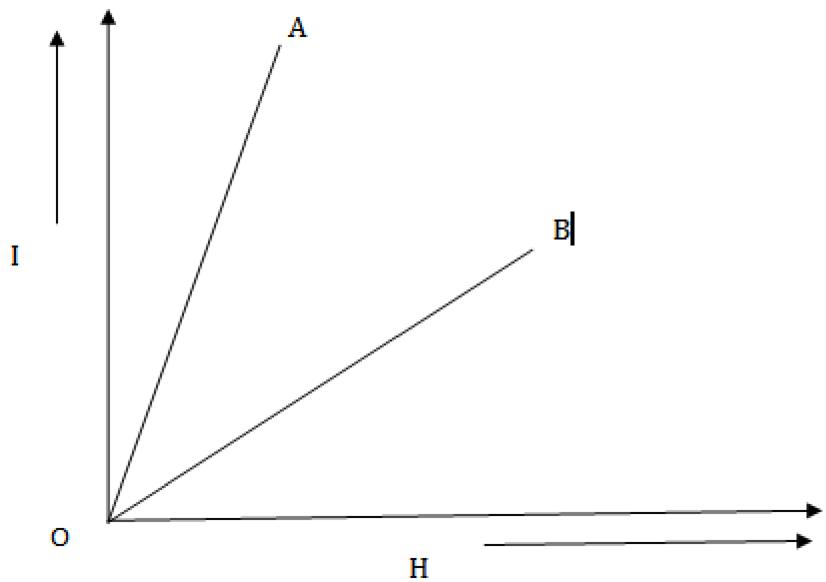
(b) It is because permeability of Iron (Ferromagnetic material) is very high as compared to that of air .

(c) $X_m = \frac{I}{H}$

$$X_m = \frac{AM^{-1}}{AM^{-1}} = 1$$

Therefore, X_m have no Units

5. (i) Figure below shows the variation of intensity of magnetization (I) versus the applied magnetic field intensity (H) for two magnetic material A and B



(a) Identify the materials A and B

(b) For the material A, plot the variation of I with temperature.

(ii) The relative permeability of a material is

(i) 0.999

(ii) 1.001

Solution

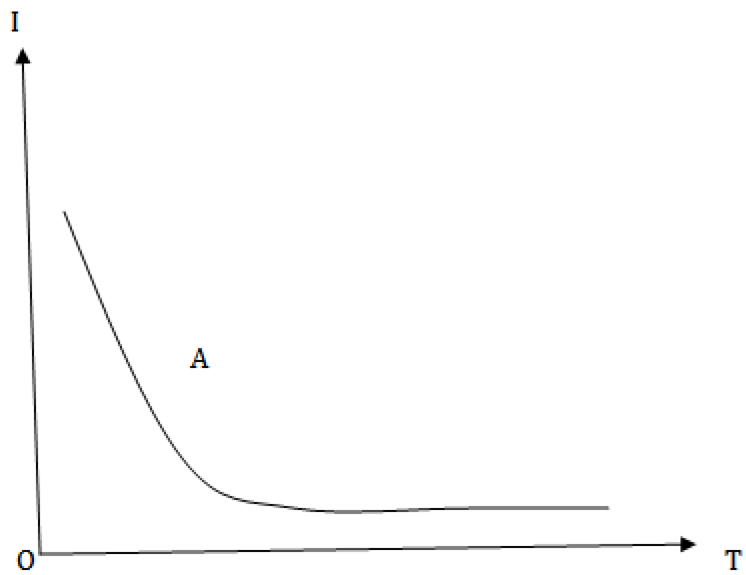
The slope I-H graph gives the magnetic susceptibility X_m of the material

$$slope = \frac{T}{H} = X_m$$

For material A the slope is positive and has a small value. Therefore, material A is paramagnetic.

For material B, the slope is position and has a large value. Therefore , material B is ferromagnetic

(a) The intensity of magnetization of a paramagnetic material A is inversely proportional to the absolute temperature $I \propto {}^{I}/_{T}$ therefore I – T graph for material A will be as shown in figure below .

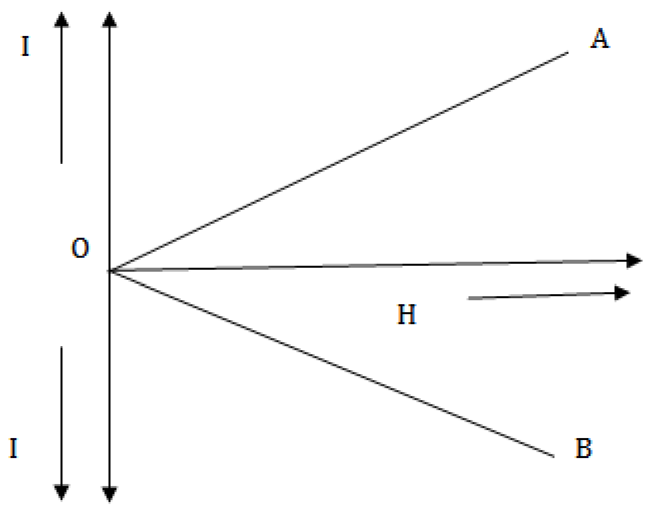


(ii) (i) Diamagnetic material

(ii) Paramagnetic material

6. (a) Graph below shows the variation of intensity of magnetization(I) versus the applied

Magnetic field intensity (H) for two magnetic material A and B



(i) identify the materials A and B

(ii) draw the variation of susceptibility with temperature for material B

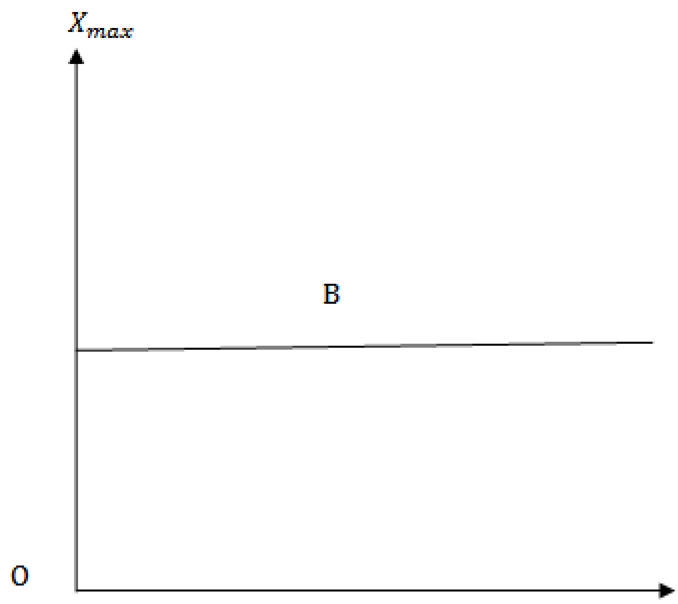
(b)A magnetizing for of 360 Am^{-1} produce a magnetic flux density of 0.6T in a ferromagnetic material. Calculate

(i) Permeability
(ii) susceptibility of the material

Solutions

(a) for material A the susceptibility X_m (= slope of I – H graph) is small and positive therefore material A is paramagnetic and for material B, Susceptibility is small and Negative. Therefore , material B is Diamagnetic

iii) The susceptibility of a diamagnetic material B is independent of temperature therefore X_m– T graph for material B will be shown in graph below.



(b) (i) Permeability of material

$$\mu = \frac{B}{H} = \frac{0.6}{360}$$

μ **= 1.67 X 10^{-3} A $^{-1}$Tm**

(ii) Susceptibility of the material

$\mu = \mu_0 (1 + X_m)$

$\mu = \mu_0 + \mu_0 X_m$

$\mu_0 X_m = \mu - \mu_0$

$$X_m = \frac{\mu}{\mu_0} - 1$$

$$X_m = \frac{1.67 \times 10^3}{4\pi \times 10^{-7}} - 1$$

X_m **= 1328.62 Am $^{-1}$**

6. (a) What is Magnetic solution?

(b) Name two materials which have

(i) Position susceptibility

(ii) Negative susceptibility

(c) Obtain the earth's magnetization , assuming that the earth's field can be approximated by a giant bar magnet of magnetic moment 8.0 x 10 ^{22} Am^2 ? Radius of earth = 6400km

(d) A bar magnet has Coercivity of $4x10^3$ A/m it is desired to demagnetize it by inserting it inside a solenoid 12cm long and having 60turns. With current should be sent through the solenoid

Solution

(a) Magnetic saturation

Is the maximum magnetization that can be obtained in the material when all the domains of a ferromagnetic material are in the direction of the applied magnetic filed

(b) (i) paramagnetic material e.g. Aluminum and Antimony

(ii) Diamagnetic materials e.g. Copper and Zinc

(c) magnetic moment M = 8 x $10^{22}Am^2$ radius of the earth Re = 6400km

Magnetization of the earth is given by

$$I = \frac{M}{V_0}$$

$$I = M \times \frac{3}{4\pi r_e^3}$$

$$I = 8 \times 10^{22} \times \frac{3}{4\pi \times (6400 \times 10^3)^3}$$

$$\boldsymbol{I = 72.8\ Am^{-1}}$$

The bar magnetic has a Coercivity of 4$\times 10^3\ \mathrm{A}/m$ i.e. it needs a magnetic intensity H = 4$\times 10^3\ \mathrm{A}/m$ to get magnetized

$$H = n\mathrm{I}$$

$$H = \frac{N\mathrm{I}}{l}$$

$$I = \frac{L \times H}{N}$$

$$I = \frac{(4 \times 10^3) \times 12 \times 10^{-2}}{60}$$

$$\mathbf{I = 8A}$$

7. (a) Define hysteresis loop

(b)What does the area of hysteresis loop indicate?

(c) What is the use of hysteresis loop?

(d) Why is soft iron preferred in making the core of a transformer?

Solution

(a) hysteresis loop

Is the resulting B – H curve (Closed loop) obtained when a ferromagnetic material is subjected to one circle of the magnetization

(b) the area of hysteresis loop is a measure of energy wasted in a sample when it is taken through s complete of magnetization

(c) The hysteresis loop of a material tells us about hysteresis loss retentively and Coercivity. This knowledge helps us in selecting materials for making electromagnetic permanent magnets cores of transformer

(d) The area of hysteresis loop for soft iron is small. Therefore energy dissipated in the core for cycle magnetization is small. For this reason, the core of a transformer is made of soft iron

8. (a) state curie law

(b) Give the graph between I and B/T

(c) What happens if an Iron magnet is melted?

(d) Copper Sulphate is paramagnetic with a susceptibility of 1.68×10^{-4} at 293K. What is the susceptibility of copper at 77.4K if it fellows curie law?

Solution

(a) Curie law

States that intensity of magnetization (I) if a paramagnetic substance is directly proportional to the external magnetic field (B) and inversely proportional to the absolute temperature(T) of the substance.

I $\propto B$

I $\propto \frac{I}{T}$

Combining these two factors, we have

$$I \propto \frac{B}{T}$$

$$I = C\frac{B}{T}$$

where

C is a constant of proportionality and is called curie constant

This law is physically reasonable As B increase the alignment of magnetic moments increases and therefore I increases

If the temperature is increased the thermal motions will make alignment difficult thus decreasing I

The curve law is found to hold good so long as $\frac{B}{T}$ does not become too large

Since

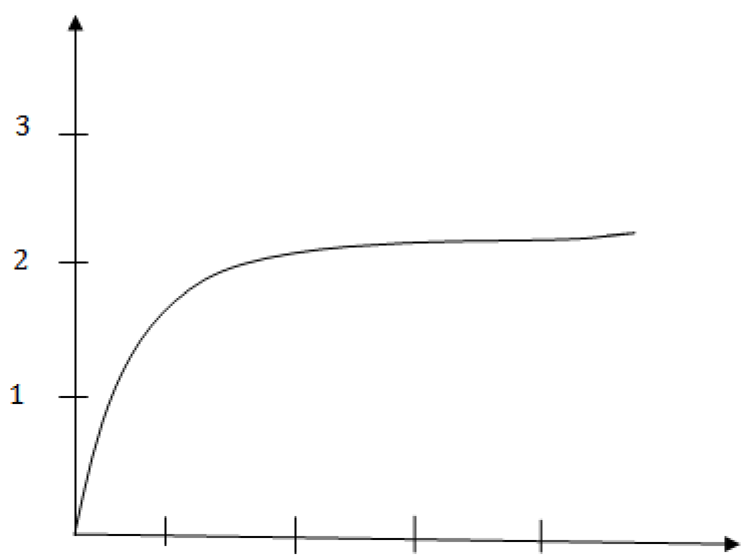
$B \propto H$

$I \propto \frac{B}{T}$

$I \propto \frac{H}{T}$

$\frac{I}{H} = \frac{1}{T}$

$X_m \propto \frac{1}{T} \quad curie's\ law$



(b) The temperature of molten iron 770^0C is above the Curie temperature i.e. On malting the iron becomes paramagnetic. Therefore it loses its magnetism

Solution

According to curie law the susceptibility depends inversely on the temperature

$$X_m = \frac{c}{T}$$

$$\frac{X_2}{X_1} = \frac{T_1}{T_2}$$

$$X_2 = X_1 \times \frac{T_1}{T_2}$$

$$X_2 = 1.68 \times 10^{-4} \times \frac{293}{77.4}$$

$$\boldsymbol{X_2 = 6.37 \times 10^{-4}}$$

10. (a) A solenoid 0.6m long is wound with 1800 turns of copper wire. An iron rod having a relative permeability of 500 is placed along the axis of the solenoid. What are the magnetic intensity H and field B when a current of 0.9A flows through the wire? What is the intensity of magnetization I in the iron? Find the average magnetic moment per iron atom. Density of iron is 7850 Kg/m^3.

(b) An Iron sample having mass 8.4Kg is repeadly taken over cycle of magnetization at a frequency of 50cyles per second it is found that energy equal to 3.2 x 10^4 J is dissipated as heat in the sample in 30Minutes if the density of the iron is 7200kg/m^3 find the energy dissipated per unit volume per cycle in the iron sample.

(c) A domain in ferromagnetic iron is in the form of a cube of side length 1μM. Estimate the number of Iron atom in the domain and the maximum possibility dipole moment and magnetization of the domain and the maximum possible dipole moment and magnetization of the domain. The molecular mass of Iron is 55g/mole and its density is 7.9g/cm^2 assume that each Iron atom has a dipole moment of 9.27×10^{-24} Am^2

Solution

$H = n\mathrm{I}$

$H = \frac{N}{L} \cdot \mathrm{I}$

$H = \frac{1800}{0.6} \times 0.9$

H = 2700A/m

$B = \mu_0 \mu_r H$

$B = (4\pi \times 10^{-7}) \times 500 \times 2700$

$\boldsymbol{B = 1.69T}$

$I = (\mu_r - 1)_{\mathrm{H}}$

$I = (500 - 1) \times 2700$

I = 1.35×10^{6} A/m

One Kilo mole (55.85Kg) of Iron has 6.2 $\times 10^{26}$ atoms. Therefore, number of atoms in 1 m^3 of Iron

$= \frac{7850 \times (6.02 \times 10^{26})}{55.85}$

$\boldsymbol{8.48 \times 10^{28} atoms}$

Average magnetic moments per iron atoms

$= \frac{Magnetic\ moments/M_3}{number\ of\ atoms/M^3}$

$$= \frac{1.35 \times 10^{6}}{8.48 \times 10^{28}}$$

$=1.59 \times 10^{-23}$ Am^2

Solution

$$Volume = \frac{mass}{density}$$

$$\frac{8.4}{7200}$$

$$volume = 1.17 \times 10^{-3} m^{3}$$

f =50 HZ

$$\text{energy dissipated}/_{\text{second}} = P$$

$$P = \frac{3.2 \times 10^{4}}{t} = \frac{3.2 \times 10^{4}}{30 \times 60}$$

$$P = 17.78\, J/S$$

$$\text{Energy dissipated/Unit volume}/_{\text{cycle}}$$

$$= \frac{P}{Vf}$$

$$= \frac{17.78}{1.17 \times 10^{-3} \times 50}$$

$$\mathbf{304\ J/m^{3} curie^{-1}}$$

Length of cubic domain l

l =1μM = 10^{-6}M

Volume of Domain

V = l_3

V = $(10^{-6})^3$

V = $10^{-18}M^3$

Mass of domain

=Volume X Density

=10^{-12}cm x7.9g

= 7.9 x10^{-12}g/cm

It is given that 55g of Iron contain 6.023 x10^{23} Iron atoms (Avogadro's no)

Number of atoms in the domain is

$$N = \frac{6.023 \text{ X } 10^{23} \text{ X } 7.9 \text{ X} 10^{-12}}{55}$$

N = 8.65 X 10^{10}atoms

The maximum possible dipole moment M_{max} is achieved per the case when all the atomic domains are perfect aligned (This condition is unrealistic)

M_{max} = (8.65 x10^{10}) x (9.27 x10^{-24})

= 8 x 10^{-13} AM^2

Maximum intensity of magnetization of the Domain is

$$I_{Max} = \frac{M_{max}}{V} = \frac{8\times10^{-18}}{10^{-18}}$$

$$I_{Max} = \underline{8 \times 10^{5}\ Am^{-1}}$$

NUMERICAL PROBLEMS

1. (1)The magnetic moment of a magnet (10cm x 2cm x 1cm) is 1AM? What is the intensity of magnetization?

   $I = 5 \text{ X } 10^{4}$ A/m

2. (2)An iron rod of cross sectional area 4cm^2 is placed with its length parallel to a magnetic flied of intensity 1600 A/M the flux through the rod is 4×10^{-4}Wb what is the permeability of the material of the rod?

   $\mu = 0.625 \times 10^{-3}$ Wb A^{-1} m^{-1}

3. (3) A toroid winding carrying a current of 5A is wound with 300turns/miter of core. The core is Iron which has a magnetic permeability of 5000Mo under the given conditions

   Find (i) the magnetic intensity H

   (ii) Flux density B

   (iii) Intensity of magnetization I

   i) 1500AT/m

   ii) 9.43T

   iii) $7.5 \text{ X} 10^{6}$A/m

4. (4)A specimen of Iron is uniformly magnetized by a magnetizing field of 500 A/m. if magnetic induction in the specimen is 0.2Wb/m^2, find the relative permeability and susceptibility

$X_m = 317.5$

$M_r = 318.5$

5. Consider a toroid of 1000 turns and mean radius 25cm. what is the B- field in the toroid if there is a current of 2A?

What will be the field when the toroid is filled with Iron per which $\mu = 100$H/m?

$B_0 = 1.6 \times 10^{-3}$

B = 0.16T

6. An Iron of volume 10^{-4}m^3 and relative permeability 1000 is placed inside a long solenoid wound with storms/cm. if a current of 0.5A is passed through the solenoid, find the magnetic moment of the rod.

M = 25Am^2

7. The flux through a certain toroid clangs from 0.65m Wb to 0.91M Wb when Air core is replaced by another material. What are

i) The relative permeability

ii) Absolute permeability of the material

$\mu_r = 1.4$

$\mu = 5.6 \times 10^{-7}$H/m

8. Answer the following Questions

a) Why does a paramagnetic sample display greater magnetization (per the same magnetizing field) when cooled?

b) Why is diamagnetism, in contrast almost independent of temperature?

c) Distinguish between a soft and a hard magnetic material, giving an example of soft magnetic materials are those which can easily be magnetized but do not retain their magnetism (retentively)

An example of soft magnetic material

Is soft Iron i.e. Iron in a reasonably pure state. It is otherwise known as wrought iron

Hard magnetic material

Are those which are difficultly to magnetic but once magnetized, can retain the magnetism per long

These are usually used making permanent magnetic

An example of hand magnetic material is steel which consists of iron and a small % of carbon

MOTION OF CHARGED PARTICLE IN UNFORM MAGNETIC FILED

Consider a charged particle of charge +Q and mass M moving with a velocity V in the plane of the paper.

Suppose this charged particle enters a uniform magnetic filed B which is perpendicular to the plane of the paper and directed outward

Clearly the entry of the charged particle is at right angles to the magnetic field

The force i.e. magnetic force F_m on the charged particle is given by

$$F_m = BQV$$

The magnetic force Fm acts at right angle to the plane containing V and B

On entering the magnetic field at M the charged particle experiences a force of magnitude F_m and is deflected in the direction shown

This force is at right angle to the direction of motion of the charge particle and therefore, cannot change the speed of charge particle it only charge its direction of motion

A moment later, then the particle reaches point N the magnitude of force Fm acting on it is the same as it was at M but the direction of force is different (Fm is still perpendicular to V)

Thus the force is perpendicular to the direction of motion of the charged particle at all times and has a constant magnitude

The magnetic force does not change the speed or kinetic energy of the charge particle it only charges the direction of the charged particle

When the moving charged particle is inside the uniform magnetic field, it moves along a circular path.

When the initial velocity of the particle is parallel to the magnetic field

$\theta = 0^0$

From

$F_m = BQV^{\sin\theta}$

$Fm = BQV^{\sin 0^\circ}$

Fm = 0

Thus in this case the magnetic field does not exert any force on the charge particle

Therefore the charged particle will continue to move parallel to the magnetic field then $\theta = 180^0$

$F_m = BQV\sin 180°$

$F_m = 0$

Therefore, the particle will continue to the move in the original direction.

When the initial velocity of the particle is perpendicular to the magnetic field $\theta = 90^0$

From

Fm = BQV$\sin 90^\circ$
Max. Value Fm = BQV

PARAMETERS OF MOTION

A force of constant magnitude F_m always acts perpendicular to the direction of motion of the charged particle.

Therefore , F_m provides the necessary centripetal force F_C to more the charged particle in a circular path in the circle of radius r perpendicular to the field

i) RADIUS OF PATH

The acceleration of a particle moving along a circular path of radius r is given by

$$a_c = \frac{V^2}{r}$$

$$F_M = F_C$$

$$F_M = Ma_c$$

$$F_M = \frac{MV^2}{r}$$

$$BQV = \frac{MV^2}{r}$$

$$r = \frac{MV}{BQ}$$

For a given charge mass and magnetic field r $\propto$ V. this means that fast particles move in large circles and slow ones in small circles.

ii) TIME PERIOD

The time taken by the charged particle to complete one circular revolution in the magnetic field is its Time period T

From

$$V = r\omega$$

$$V = r \cdot \frac{2\pi}{T}$$

$$T = \frac{2\pi r}{V}$$

$$T = \frac{2\pi}{V} \cdot r$$

$$T = \frac{2\pi}{V}\left[\frac{MV}{BQ}\right]$$

$$T = \frac{2\pi M}{BQ}$$

Thus Time period of the charged particle is independent of the speed (V) and the radius of the path

It only Depends on the magnitude of B and charge to mass ratio $Q/_{M}$ of the particle .

FREQUENCY

The number of circular revolutions made by the charged particle in one second is its frequency f

$$f = \frac{1}{T}$$

$$f = 1 \div \frac{2\pi M}{BQ}$$

$$f = \frac{BQ}{2\pi M}$$

$$T = \frac{2\pi M}{BQ}$$

There Frequency of the charged particle is also independent of speed (V) and radius (r) of the path

ANGULAR FREQUENCY

From

ω =2πf

But

$$f = \frac{BQ}{2\pi M}$$

Then

$$w = 2\pi.\frac{BQ}{2\pi M}$$

$$w = \frac{BQ}{M}$$

Again Angular frequency of the charged particle is independent of the speed (V) and radius (r) of the path..

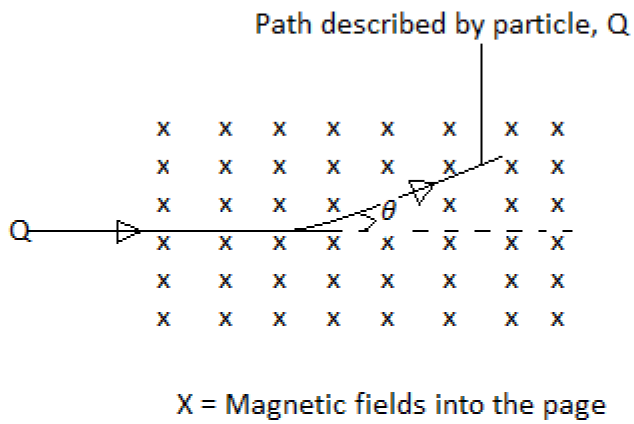
Since T, f and ω of a charged particle moving in a magnetic field are independent of its speed (V) and the radius (r) of the path.

In fact all the charged particles with same Q/M and moving in a uniform magnetic field B will have the same value of T, f and w

MOTION OF CHARGED PARTICLE ENTERING UNIFORM MAGNETIC FIELD AT AN ANGLE

Suppose the charged particle moving with velocity V enters a uniform magnetic field B making an angle θ to the direction of the field

Diagram



The velocity V can be resolved into two rectangular components

i) $V_1 = V\cos\theta$ Acting in the direction of the field

ii) $V_2 = V\sin\theta$ acting perpendicular to the direction

The perpendicular component V_2 moves the charged particle in a circular party while the horizontal component V_1 moves it in the direction of the magnetic field

In other words, the charged particle corers circular path as well as linear path. Consequently the charged particle will follow a helical path.

The charged particle rotates in a circle at speed V_2 while moving in the direction of the field with a speed V_I

PARAMETERS OF MOTION

The perpendicular component of velocity V_2 determines the parameters of the circular motions while the horizontal component of velocity V_I decides the pitch of helix

(i) Radius of path

From

$$F_M = Ma$$

$$BQV_2 = \frac{MV_2{}^2}{r}$$

$$BQ = \frac{MV_2}{BQ}$$

$$r = \frac{MV_2}{BQ}$$

$$r = \frac{MV\sin\theta}{BQ}$$

(ii) T, f and w

Since time period (T), frequency (f) and Angular frequency(ω) of a charged particle moving in a uniform magnetic field
are independent of speed V and radius (r) of the path, these values remain the same

$$T = \frac{2\pi M}{BQ}$$

$$f = \frac{BQ}{2\pi M}$$

$$w = \frac{BQ}{M}$$

(iii) Pitch of helix (d)

It is the linear distance covered by charged particle when it completes one circular revolution Or It is the linear distance covered by charged particle during time T

d = V_1T

d = V$\cos\theta$T.

d = V$\cos\theta \times \frac{2\pi M}{BQ}$

$$d = \frac{2\pi m}{BQ}[V\cos\theta]$$

The following points may be noted about the behavior of charged particle in a Uniform magnetic field

(i) If a charged particle is at rest V=0 in a magnetic field, it experiences no force

From

F_m= BQV$\sin\theta$

F_m= BQ$\times 0 \times \sin 0$

F_m= 0

(ii) If a moving charged particle enters a uniform magnetic field at right angles to the field it describes a *circular path*

(iii) If a moving charged particle enters a uniform magnetic field. Making an angle to the direction of the field it describes a *helical path*

(iv) A moving charged particle in a magnetic field experience maximum force when angle between V and B is 90^0

(v) Since magnetic force does not change the speed of a charged particle it means that K.E of the charged particle remains constant in the magnetic field.

(vi) Since magnetic force (Fm) is perpendicular to V, it does not work. Therefore work done by the magnetic force on the charged particle is zero

WORKED EXAMPLE

1. An electron and a proton moving in a circular path at $3x10^6$ Ms^{-1} in a uniform magnetic field of magnitude 2×10^{-4}T. Find the radius of the path

Solution

$$\mathrm{Bev} = \frac{\mathrm{mV}^2}{\mathrm{r}}$$

$$r = \frac{\mu_e V}{Be}$$

$$\frac{9 \times 10^{-31} \times (3 \times 10^6)}{(2 \times 10^{-4}) \times (1.6 \times 10^{-19})}$$

$$r = 8.5 \times 10^{-2}\mathrm{m}$$

2. An electron and a proton moving with the same speed enter the same magnetic field region at right angles to the direct of the field. For which of the two particles will the radius of circular path be smaller?

Solution

From

$$r = \frac{MV}{Be}$$

$r \propto m$

Since the mass of electron is less than that of the proton the radius of the circular path of electron will be smaller.

3. (a) What will be the path of a charged particle moving along the direction of a uniform magnetic field?

(b) A moving charged particle enters a magnetic

Solution

(a) When a charged particle moves along the direction of a uniform magnetic field ,it experiences no force $\theta = 0^0$ therefore the charged particle will more along its original straight path

(b) Helical path since the velocity of the charged particle can be resolved into two rectangular components one along the field, and the other perpendicular to the field. The velocity component perpendicular to the field causes the charged particle to more in a circular path while the velocity component along the field cause it to more it in the direction of the field. The combination of these two motions course the charged particle to move in a helical path.

4. A become of α-particles and of proton of the same velocity V, entries a uniform magnetic field at right angles to the field lines. The particle describes circular paths. What is the ratio of radii of the two paths?

Solution

Radius of α-particle path r_1

$$r_1 = \frac{M_1 V}{BQ_1} \dots\dots\dots\dots\dots . (i)$$

Of proton path r_2

$$r_2 = \frac{M_2 V}{BQ_2} \dots\dots\dots\dots\dots\dots\dots\dots\dots (ii)$$

Take equation (i) $\div$ equation (ii)

$$\frac{r_1}{r_2} = \frac{M_1 V}{BQ_1} \cdot \frac{BQ_2}{M_2 V}$$

$$\frac{r_1}{r_2} = \frac{Q_2}{Q_1} \cdot \frac{M_1}{M_2}$$

$$\frac{r_1}{r_2} = \frac{1 \times 4}{2}$$

$$\frac{r_1}{r_2} = 2$$

Therefore, radius of x - particle path is twice that of proton's path

Note

α – particle

Charge = 2e

Mass = mass of helium nucleus

5. A proton with charge – mass ratio of 10^8 CKg $^{-1}$ is moving in a circular orbit in a uniform magnetic field of 0.5T. calculate the frequency of revolution

$$f = \frac{BE}{2\pi M}$$

$$F = \frac{B.\,e}{2\pi\, M}$$

$$F = \frac{0.5\ .\ 10^8}{2\pi}$$

6. (a) What happens when a charged particle is projected perpendicular to a uniform magnetic field ?

(b) A beam of protons moving with a velocity of 4×10^5 M5^{-1} enters a uniform field of 0.3T at an angle of 60^0 to the direction of a magnetic field find

(i) The radius of helical path of proton beam

(ii) Pitch of helix

Given mass of proton = 1.67×10^{-27} Kg charge on proton = 1.6×10^{-19}C

Solution

(a) When a charged particle is projected perpendicular to a uniform magnetic field

(i) Its path is circular in plane perpendicular to B and V

(ii) Its speed and Kinetic energy remain the same

(iii) The magnitude of force remains the same Fm = BQV. Only the direction of velocity of the particle charges.

(iv) The force acting on the particle is independent of the radius of the circular path

(v) The time period of revolution of the particle is independent if V and r

(b) (i) Solution

$$r = \frac{MV\sin\theta}{BQ}$$

$$r = \frac{(1.67\ X10-27)X\ (4\ X105)\sin 60°}{(0.3)\ \mathrm{x}\ (1.6\ \mathrm{x}\ 10\ -19)}$$

r = 1.2 x10^{-2}M

(ii) Pitch of helix of d

$$d = \frac{VCOS\,\theta}{\frac{I}{T}}$$

$$d = V\ COS\,\theta°\ .T$$

$$= \frac{V\ COS\,\theta.2\pi M}{BQ}$$

$$d = \frac{4X10^5 XCOS\ 60°\ X\ 2\pi\ X\ 1.67\ x10^{-27}}{0.3\ \mathrm{x}\ 1.6\mathrm{x}\ 10^{-19}}$$

d = 4.37 x10^{-2}M

7. (a) A particle of charge ð□œ*f* moves in a circular path of radius r in a uniform is P = BQr

(b) An electron emitted by a heated cathode and accelerated through a potential difference of 2.0KV enters a region of magnetic field of 0.15 determine the trajectory of the electron if the field

(i) is traverse to its initial velocity
(ii) makes an angle of 30 with the initial velocity

c) A proton a deuteron and an α – particle whose kinetic energies are same enter perpendicularly to a uniform magnetic field. Compare the radii of their circular paths

(a) Solution

The magnetic force Fm provides the necessary centripetal force F_c

$F_m = F_c$

$$BQV = \frac{MV^2}{r}$$

$$BQ = \frac{MV}{r}$$

BQ = $\frac{momentum}{r}$

Momentum P = BQr

(b) Solution

When an electron (e) is accelerated through a p.d of V, it acquires energy eV. IF Vis the velocity gained by electron, then

$$eV = \frac{IM(v^2)}{2}$$

$$V = \sqrt{\left(2 \times \frac{eV}{IM}\right)}$$

$$V = \sqrt{(2x(1.6 \times 10^\wedge - 19)\, X\, (2X10^\wedge 3))}$$

V=9X10^{-31}

(i) Force on the electron due to transverse field is

Fm = BQr ð□œf = 90^0

Since magnitude of Fm is constant and Fm is perpendicular to both V and B the electron will move in a circle of radius r. The necessary centripetal force is provided by Fm

$$F_m = \frac{Mv^2}{r}$$

$$r = \frac{Mv^2}{BQv}$$

$$r = \frac{(9x10^{-31})x}{(10^{-15})x\,(1.6\,x10^{-19})}$$

r = 10^{-3}M

(ii) When electrons enters the magnetic field making an angle ð□œf = 30^0 with the field

$$BQBVSin\theta = \frac{M\,(rSin\,\theta^{\circ})^2}{r}$$

$$r = \frac{MVSin\theta}{BQ}$$

$$r = \frac{9x10^{-31} x \; Sin\; 30^o}{0.15 \; x1.6 \; x \; 10^{-19}}$$

r = 0.5 x10^{-3}M

(C) Solution

Let 1, 2 and 3 be the suffix force proton, deuteron and α - particle respectively

K .E, = K.E$_2$ = K.E$_3$

$\frac{1}{2} M_I V_1 = \frac{1}{2} M_2 V_2 = \frac{1}{2} M_3 V_3$

If MI then M$_2$ = 2M and M$_3$ = 4M

$MV^2_I = 2MV^2_2 = 4MV^2_3$

$V_I = \sqrt{2}V_2 = 2V_3$

$V_2 = \frac{VI}{\sqrt{2}}$

Also

$$V_3 = \frac{V_1}{2}$$

Radius of the path r

$r = \frac{MV}{BQ}$

If QI = Q, Then Q2 = Q and Q3 = 2Q

$r_I = \frac{M_I V_I}{BQ_I} = \frac{MV_I}{BQ}$

$r_2 = M_2V_2 \;=\; 2M. \;\; V_1$

$BQ_2 \qquad BQ \;\; \sqrt{2}$

$$r3 = \frac{M_3V_3}{BQ3} = \frac{4M}{2BQ} \cdot \frac{V_1}{2}$$

$r_{l:}\ r_2 : r_3 = \underline{MV_l} : \frac{\sqrt{2}}{}\underline{MV_l} : \underline{MV_l}$

$BQ \qquad BQ \qquad BQ$

$\mathbf{r_l : r_2 : r_3 = 1 : \frac{\sqrt{2}}{} : 1}$

NUMERICAL PROBLEMS

1. What is the radius of the path of an electron (mass 9×10^{-31})Kg and charge(1.6×10^{-16}C) moving at a speed of 3×10^7m/s in a magnitude field of 6×10^{-4}T perpendicular to it? What its frequency? Calculate its energy in KeV(1Ev=1.6×10^{-19} J)

r= 0.28m

f= 1.7×10^{-7} Hz

E = 2.53KeV

2. An electron after being accelerated through a p.d of 100V enters a uniform magnetic field of 0.004T perpendicular to its direction of motion. Find the radius of the path described by the electron

r = 8.4 10^{-3}m

3. An α-particle is describing a circle of radius of 0.45m in a field of magnetic Induction 1.2Wb/m^2. Find its speed, frequency of rotation and kinetic energy. What potential difference will be required which will accelerate the particle so as to this much energy to it? The mass of α-particle is 6.8×10^{-27}Kg and its charge is 3.2×10^{-19} C

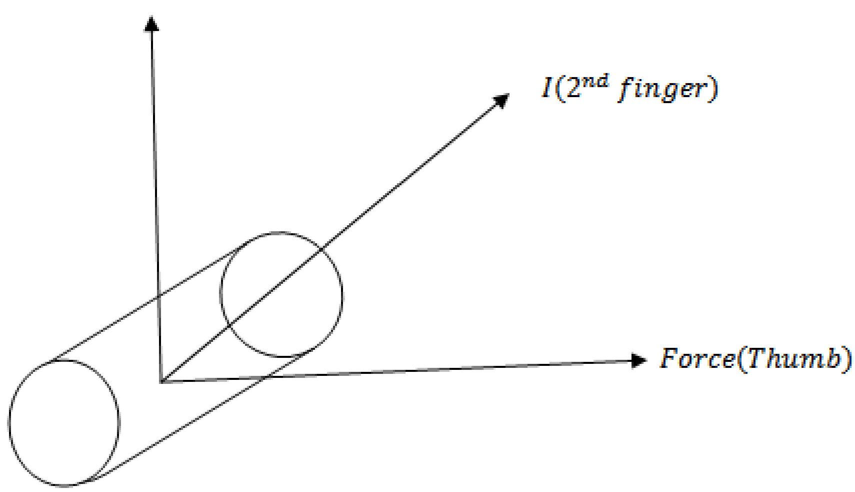
$V = 2.6 \times 10^{7}$ m/s
$f = 9.2 \times 10^{6}\ sec^{-1}$

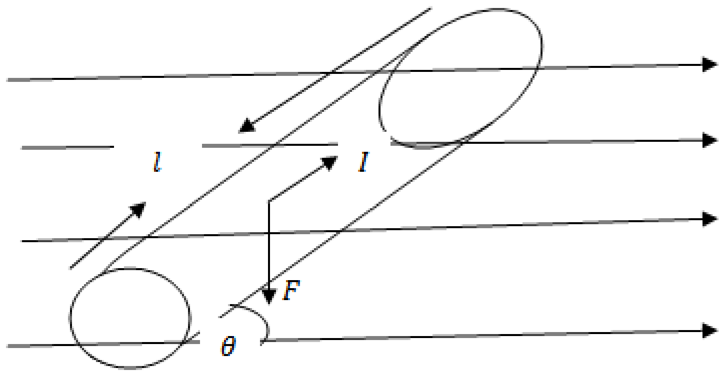
MAGNETIC TORQUE ON RECTANGULAR COIL IN UNIFORM FIELD
Consider a rectangular conductor ABCD having length L and N-turns carrying a current I and placed in a magnetic field between N and S-pole of bar magnets.

Since the velocities of the electrons in the sides side AB and BC are perpendicular to the magnetic induction B then these sides will experience the maximum magnetic forces equal to F.

F= BIL

The direction of the magnetic force in these two sides is given by Fleming's Left Hand Rule as shown in the figure below.

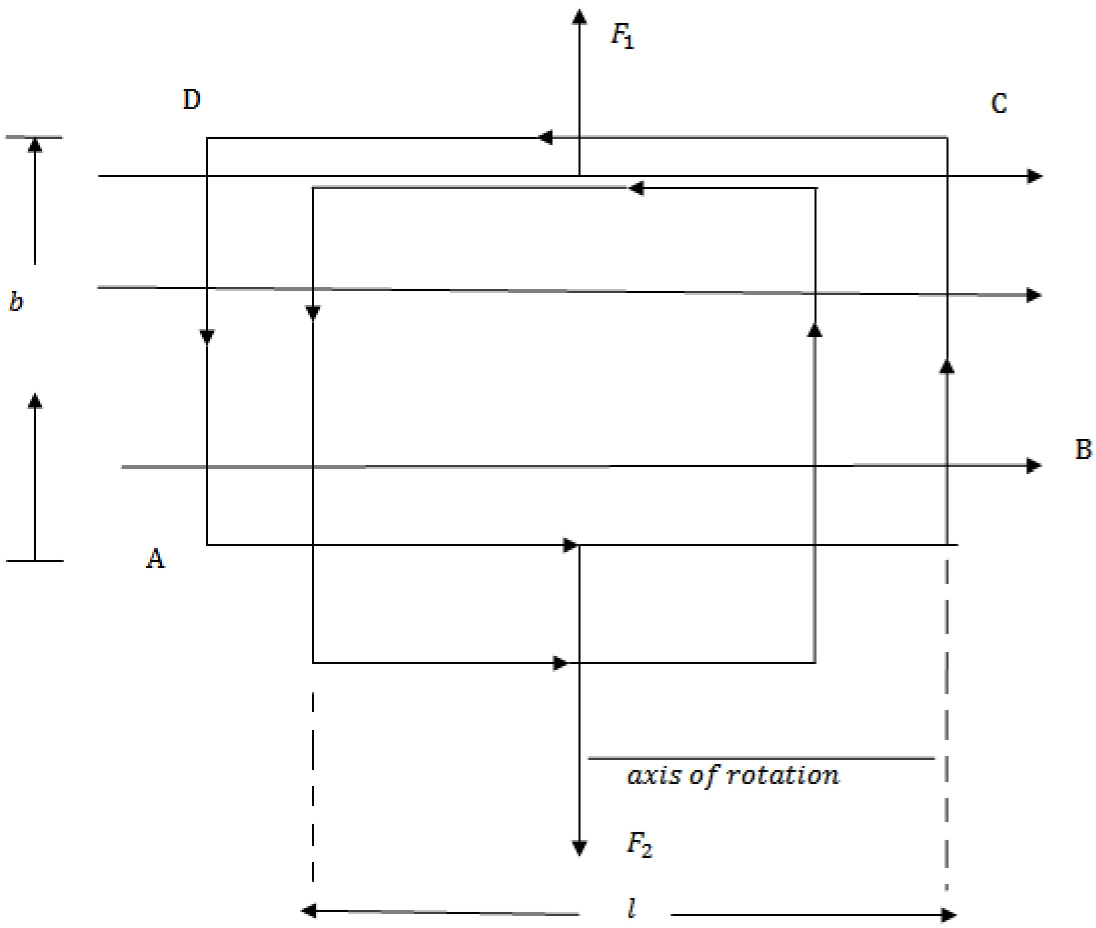




$F = IBL\sin\theta$ $\qquad$ $if\ \theta = 90°$

$F = IBL$

The two parallel and equal forces will constitute the turning of the rectangular called the magnetic torque.



Mathematically

The magnetic torque or couple is given by

τ = Total force x Perpendicular distance

τ=BIL x b

τ=BI (L x b)

$$\tau = \mathrm{BIA}$$

Lb = cross-sectional area=A

For the rectangular coil of N- turns

$$\tau = \mathrm{BANI}$$

ELECTROMAGNETIC MOMENT(M)

This is the magnetic torque acting on the coil when it is parallel to a uniform field whose flux density is one tesla.It is the property of the coil is defined as the couple required to hold the coil at right angles to field.

Thus, in equation τ=m when B=1I

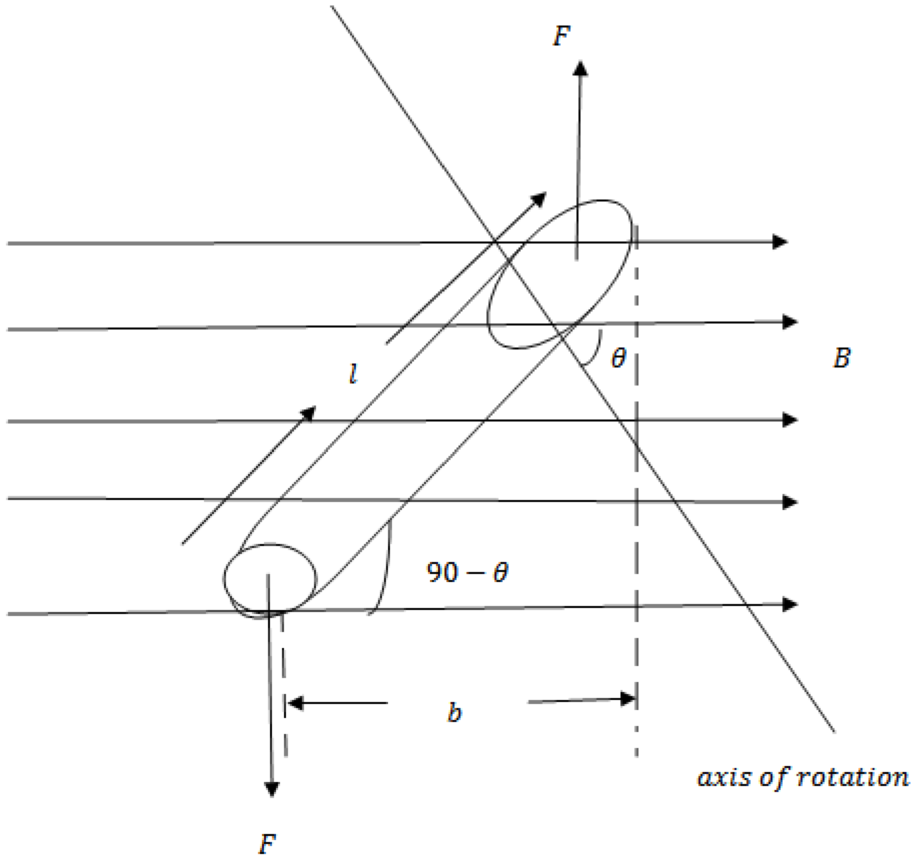
from

$$\tau = \mathrm{BANI}$$

Let magnetic moment $M = NIA$

$$Then, \tau = BM$$

Consider a rectangular coil ABCD placed at an angle θ to the magnetic field of flux density B



$F = BIL\sin(90° - \theta)$

The perpendicular distance between the parallel force is $b\cos\theta$ and not b. Then the magnetic force or compile is given

$\tau = BILb\cos\theta$

$but\ lb = area\ A$

$\tau = BIA\cos\theta$

$for\ a\ coil\ of\ N - turns\ \tau = BANI\cos\theta$

Special cases

(i) When $\theta = 0°$

Plane of the loop is parallel to the direction of magnetic field

$\tau = BANI\cos 0°$

Thus, the Torque on a current loop is maximum when the plane of the loop is parallel to the direction of magnetic field is given by

(ii) when ð□œf = 90^0

Plane of the loop is perpendicular to the direction of magnetic field

$\tau = BANI\cos 90°$

$\cos 90° = 0°$

$\tau = 0$

Thus , the torque on a current loop is minimum (zero) when the plane of the loop is perpendicular

(iii) When B = 1T

when B = 1T, then the magnetic torque is numerical equal to magnetic moment

From

$$\tau = BANI\cos\theta$$

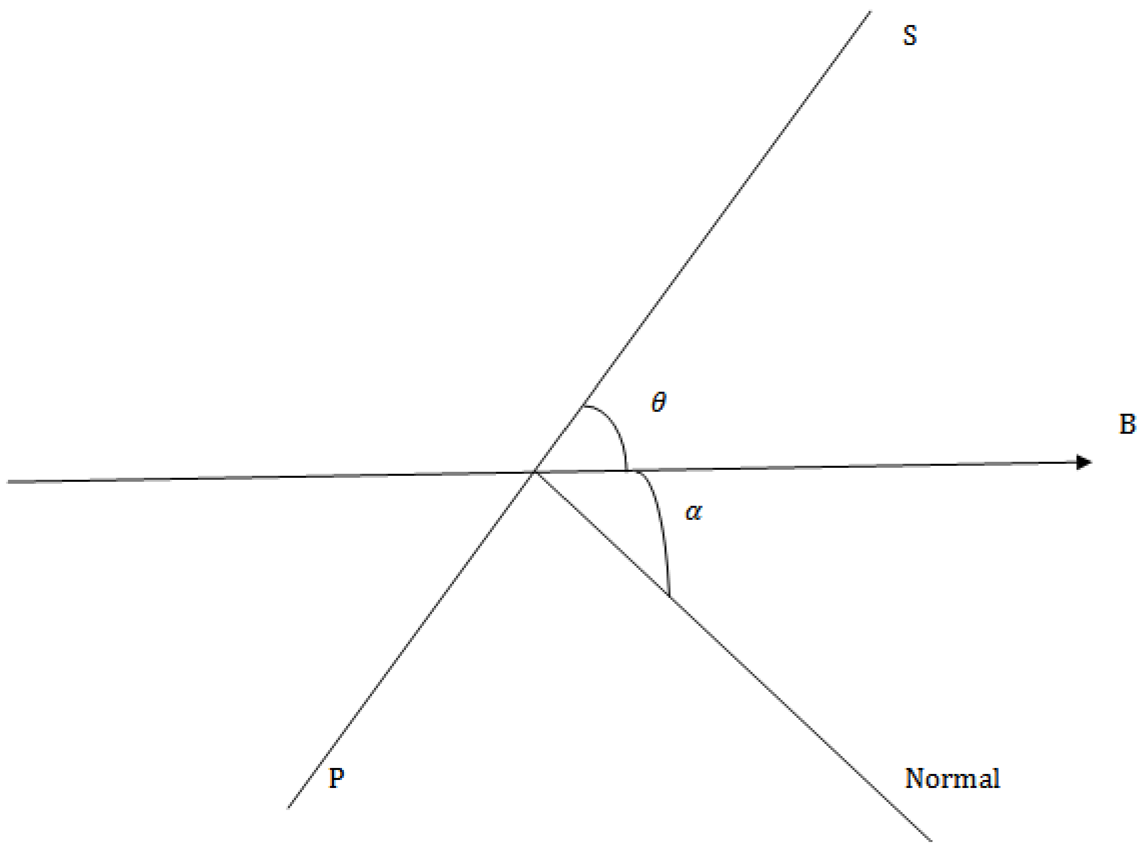
B = 1T

$$M = IAN\cos\theta$$

$$r = MB\cos\theta$$

MAGNETIC TORQUE AT AN ANGLE α BETWEEN THE AXIS OF THE COIL AND NORMAL TO THE PLANE OF THE COIL

We can also express torque in another useful form. If normal to the plane of coil makes an angle α with the direction of the magnetic field



$\alpha + \theta = 90°$

$\theta = 90° - \alpha$

From

$\tau = BANI \cos\theta$

τ = BANI cos (90 – α)

τ = ANIBSin α

Since M = IAN

$\bar{C} = MB \sin\alpha$

When a current carrying coil is placed in a uniform magnetic field, torque acts on it which tends to rotate the coil so that the plane of the coil is perpendicular to the direction of magnetic field.

WORK DONE BY TORQUE

If the magnetic torque displaces the coil through the small angular displacement $d\theta$

The work done by the torque is given by

$$dw = \tau d\theta$$

$$dw = Mb\cos\theta$$

The total work done is obtained by integration the above equation within limits

$$\int_{0}^{w} dw = \int_{\theta 3}^{\theta 2} mb\cos\theta d\theta$$

$$W = mb\int_{\theta 1}^{\theta 2} \cos\theta d\theta$$

$$W = mb[\sin\theta]_{\theta 2}^{\theta 1}$$

$$w = Mb(\sin\theta_2 - \sin\theta_1)$$

If the magnetic field displaces the coil through small angular displacement $d\alpha$the work done by torque is given by

dw = τdα

dw = mbSinαdα

$$\int_{0}^{w} dw = mb \int_{\propto 1}^{\propto 2} \sin \propto d \propto$$

$$W = -mb[\cos \propto]_{\propto 2}^{\propto 1}$$

$$W = -mb[\cos \propto_2 - \cos \propto_1]$$

$$dw = -Mb[\cos\theta_2 - \cos\theta_1]$$

WORKED EXAMPLES

1. A vertical rectangular coil of sides 5cm by 2cm has 10 turns and carries a current of 2A. Calculate the torque on the coil when it is placed in a uniform horizontal magnetic field of 0.1T with its plane.

(a) Parallel to the field

(b) Perpendicular to the field

(c) 60^0 to the field

Solution

The area of the coil

A= $(5 \times 10^{-2}) \times (2 \times 10^{-2})$

A= 10^{-3} m^2

(a) From

$\tau = BANI\cos\theta$

A=10^{-3}

I = 2A

N=10

B=0.1T

$\theta = 0°$

$\tau = 0.1 \times 10^{-3} \times 10 \times 2$

$\tau = 2 \times 10^{-3}$NM

(b)

$\theta = 90°$

$\tau = BANI \cos\theta$

$\tau = 0$

(c)

$\theta = 60°$

τ = BANICos ð□œ*f*

$\tau = 0.1 \times 10^{-3} \times 10 \times 2 \times \cos 60^0$

$\tau = 10^{-3}$NM

2. 2.Given a uniform magnetic field of 100T in East to West direction and a 44cm long wire with a current carrying capacity of at most 10A. what is the shape and orientation of the loop made of this wire which yields maximum turning effect on the loop?

Solution

A current carrying planar loop will experience maximum together if its area to the direction of the magnetic field for a given perimeter , a circle has the maximum area.

If 44cm wire is bent into a circular

$2\pi r = 44$

$\pi r = 22$

$\frac{22r}{7} = 22$

$r = 1$

7

R = 7cm

Area of loop = πr^2

$= \pi x 7^2$

$= 154cm^2$

$= 154x10^{-4}m^2$

Magnetic toque τ

$$\theta = 0°$$

$$\tau = NBIA\cos\theta$$

τ = BANI

$\tau = (154 X10^{-4}) X 100 X100$

τ = 150 T

3. A circular coil of wire of 50 turns and radius 0.05 carries current of 1A the wire is suspended vertically in a uniform magnetic field of 1.5T. the direction of magnetic field is parallel to the plane of the coil

(a) Calculate the Toque on the coil

(b) Would your answer charged if the circular coil is replaced by a plane coil of some irregular shape that has the same area (all other particulars are unaltered?)

(a) Solution

$$\tau = BANI\cos\theta$$

B = 1.5T

$A = \pi r^2 = \pi (0.05) M^2$

$A = 7.85 X 10^{-3}$

N = 50

I = 1A

$\theta = 0°$

$\tau = 1.5 \text{ X } (7.85 \text{ X}10^{-3}) \text{ X}50\text{X}1$

$\tau = 0.589\text{NM}$

(b) Since torque on the loop is independent of its shape provide area (A) remains the same the magnitude of the torque will remain unaltered.

4. A circular coil of 20turns and radius 10 cm is placed in a uniform magnetic field of 0.2T normal to the coil. If current in the coil is 5A find.

(i) Total torque on the coil

(ii) Total force on the coil

(iii) Average force on each electron in the coil due to the magnetic field. The coil is made of copper wire of cross-sectional area $10^{-5}m^2$ and force of electron density in the wire is $10^{29}m^{-3}$

Solution

(i) The toque on the coil is given by

$\tau = BANI\cos\theta$

Since
$\theta = 0°$

$\underline{\tau = 0}$

(ii) The net force on a planar current loop in a uniform magnetic field is always zero

(iii) Magnetic force on each electron

$F = BeV_d$

$$F = Be.\frac{I}{neA}$$

$$F = \frac{BI}{nA}$$

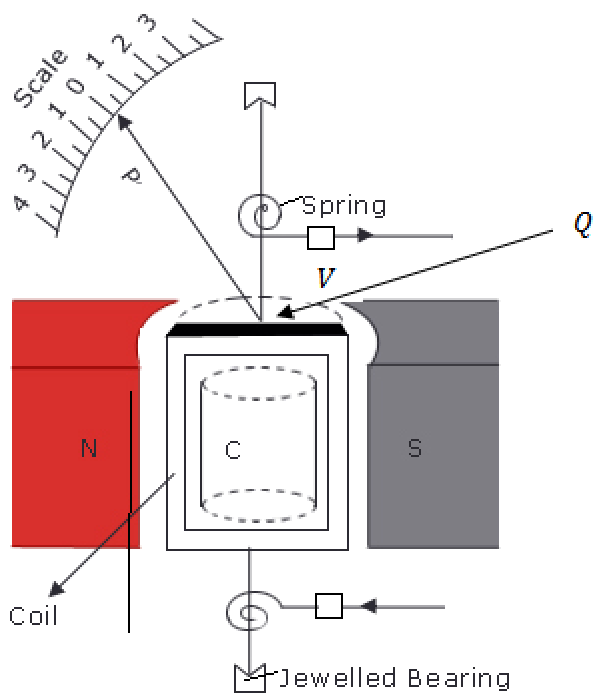
$$F = \frac{0.25 \times 5}{10^{29} \times 10^{-5}}$$

F= 10^{-24}N

MOVING COIL METERS

A galvanometer deflects or measures small amount of current passing through it and it gives the direction to which that current is flowing.

In these instruments a rectangular of fine insulated copper wire is suspended in an strong magnetic field as shown in the figure below. The field is set up between soft iron poler pierces Ns attacked to a powerful permanent magnet.



(i) Millimeter

The magnet field is radial to the core and pole pieces over the region which the coil can swing. In this case the deflected coil always comes to rest with the plane parallel to the field in which it then situated.

The moving coil galvanometer ha s has hair spring and jewel bearings. The coil is around in the rigid but light aluminium frame which also comes to carries a pivot. The current is led in and out of the springs.

Aluminum pointer P shows the deflection of coil, it is balanced by counter weight Q.

THEORY OF MOVING COIL GALVANOMETERS

The rectangular coil is situated in the radia field B when the current is passed into it the coil rotate to an angle Q which depend on the length of the spring.

No matter where the coil comes to rest, the field B in which it is situated always along the plane of the coil because the field is radial.

It is shown that the torque T of the coil is given by T = B A N I. ie.

In equilibrium the deflecting torque T is equal to the opposing torque due to the elastic forces in the springs. The opposing torque

Torque = CQ

Where C is the constant of the spring.

Example 1

A galvanometer coil has a coil has 100 turns which each turn having an area of $2.5cm^2$. If the coil is in the radial field of 2.0 10^{-9}Nm per degree what current is needed to give a deflection of 60^0?

Solution Formulae

N=100

$A= 2.5 \times 10^{-4}m^2$

$B= 2.0 \times 10^{-2}T$

$C= 5.0 \times 10^{-9}Nm$

$Q= 60^0$

$$BAIN = CQ$$

$$I = \frac{C}{BAN}Q$$

$$.\,:\, I = \frac{8\times10^{-9}\times60}{2\times10^{-2}\times2.5\times10^{-4}\times100} = 60 \times 10^{-5}$$

$$I = 6.0 \times 10^{-4}\ A$$

Example 2

A moving coil galvanometer with a coil of 15 turns and an area of $0.02m^2$ is suspended by a torsion wire which has restoring constant of 9.00 10^{-6}Nm per degree of twist if the current of â„¦MA is passed through the coil whose plane is parallel to a uniform magnetic field of 0.03T. What will be the deflection of the coil?

Solution: Formulae

N=15

$A=0.02m^2$

B=0.03T

$C=9\times 10^{-6}Nm$

$$Q=\frac{BAIN}{C}$$

$$= \frac{0.03 \times 0.02 \times 52 \times10^{-3} \times15}{9\times10^{-6}} = 52^0$$

The deflection of the coil is 52^0

CURRENT SENSITIVITY.

Therefore the greater the sensitivity is obtained with a stronger field B atom value of c that is a week springs and greater value of N and the value of A. However the size and number of turns of a mound increase the resistance of the meter which is not desirable

VOLTAGE SENSITIVITY

If the resistance of a moving coil meter is R the p.d.v across its terminals where a current I flows it. It is given by V = IR--------------- (1)

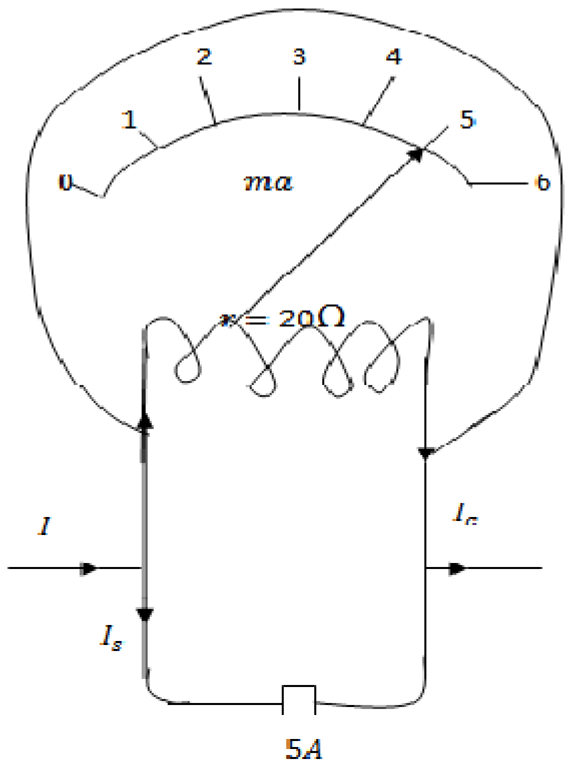
But we have BAIN = C ------------ (2)

Hence the voltage sensitivity depends on the resistance R of the meter.

Unlike the current sensitivity.

CONVERTING A MILLIAMETER TO AN AMMETER.

Moving coil meters give full scale deflection for current smaller than those generally met in laboratory. In order to measure the current of the order of an ampere or more we connect a low resistance s called the shunt across. The terminals of moving coil meter



The shunt turns most of the current to be measured I away from the coil.

Suppose the coil of meter has a resistance r of 20â,,¦ and full deflected by the current of 5MA. If we want to convert it so that its full scale deflection is 5A; Then the shunt s must be connected which will extra current that is (5- 0.005) A or 4.995A.

Potential difference across the shunt p.d across the coil ie.

Example

A milli ammeter has a full scale reading of 0.8MA and resistance of 75 ohms. What is the value of a single resistor which mould a current it into an ameter capable of reading 15amps at full scale

SOLUTION:

Given	Formula
$I_c = 0.8 \times 10^{-3}$	$I - I_c = I_s$
$r = 75\Omega$	
$I = 15A$	but $rI_c = rI_s$
$I_s = 14.9992$	$rI_c = 75 \times 0.8 \times 10^{-3} = 0.06v$

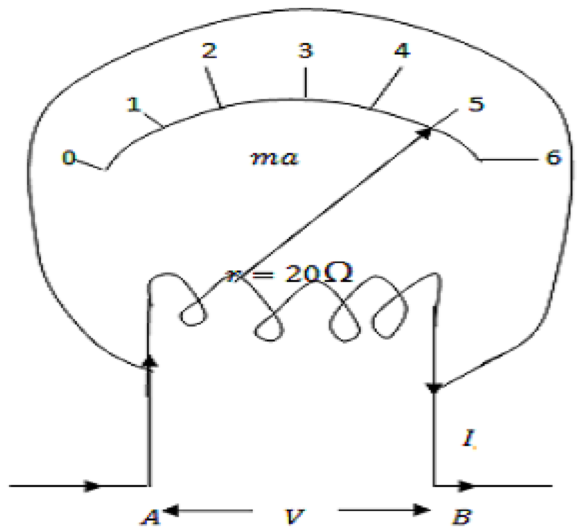
$$.:\ 0.006 = 14.992s$$

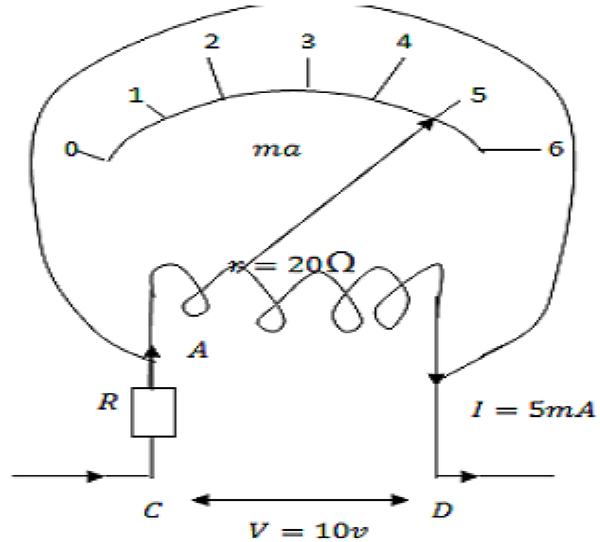
$$S = \frac{0.06}{14.992}\Omega \quad = 3.985 \times 10^{-3}$$

$$S = 4.0 \times 10^{-3}\Omega$$

CONVERTING A MILLIAMETER TO VOLTIAMETER

Suppose we have a moving coil meter which requires 5MA for full scale deflection and also let suppose that the resistance of its coil r is 20 â„¦





When this milliameter is full deflected the p.d across it is given by u = rI

= (20â„¦) (510^{-3})

=0.1u

If the coil resistance is constant the instrument can be used as a voltimeter giving a full scale deflection for p.d of 0.1 or 100mv so this milliameter can possess two scales for current and for voltage as shown in (1) above.

The p.d to be measured in the laboratory is usually greater than 100mv. Therefore to measure such a p.d we put resistor R in series of with the coil as shown in fig (2) above:

For example if we wish to measure to measure up 10v is applied between terminals CD then the scale current of 5MA flows through the moving coil. That is

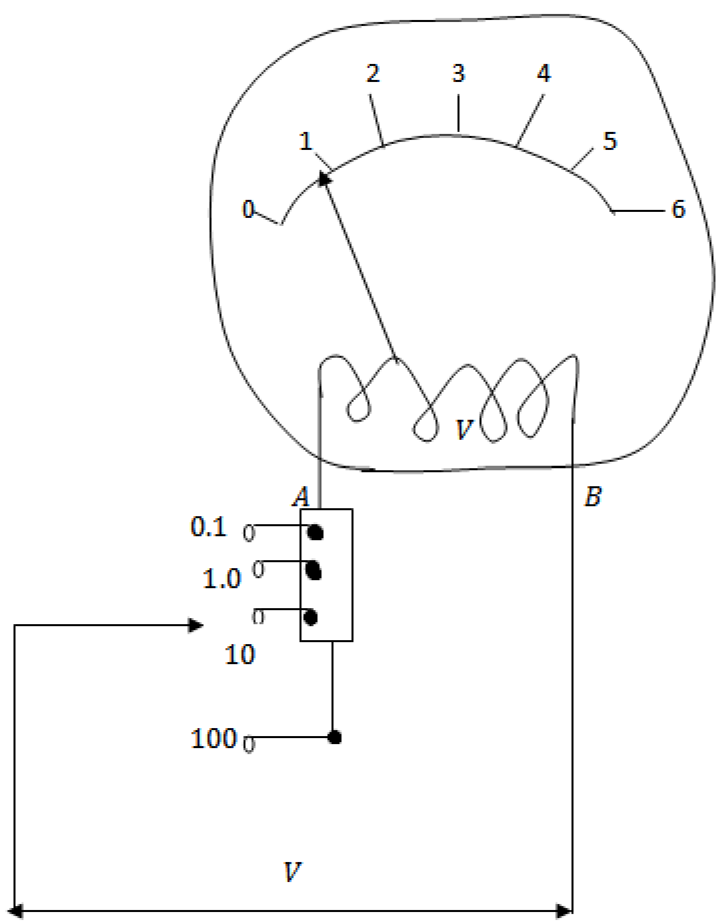
$V = (R + r) I$

$10 = (R + 20) 5 10^{-3}$

Or

$R = 2000 - 20$

=1980â„¦



The resistance r is called a multiplier.

Many voltmeters contain a series of multiplier contains a series of multiplier of different resistances which can be chosen by a switch…. and socket arrangement shown in fig (3) above

Example

A voltmeter whose range is 0-200v has resistance R of 1500 per volt.(fsd) what resistance should be converted in series with it to give a range of 0-2000v.

Solution

Given

But when the resistance in series they have the same current through it

R= 1500 So $I = \frac{200}{1500} \quad \frac{2}{15}$

V_R = 200v

V= 2000v

r= x

New solution:

Total resistance for this voltmeter is

=1500 200 =30 10^{-4}â,,¦

Total resistance for the new scale is

$=15002000 = 30\ 10^5$ â„¦

Extra resistance required is

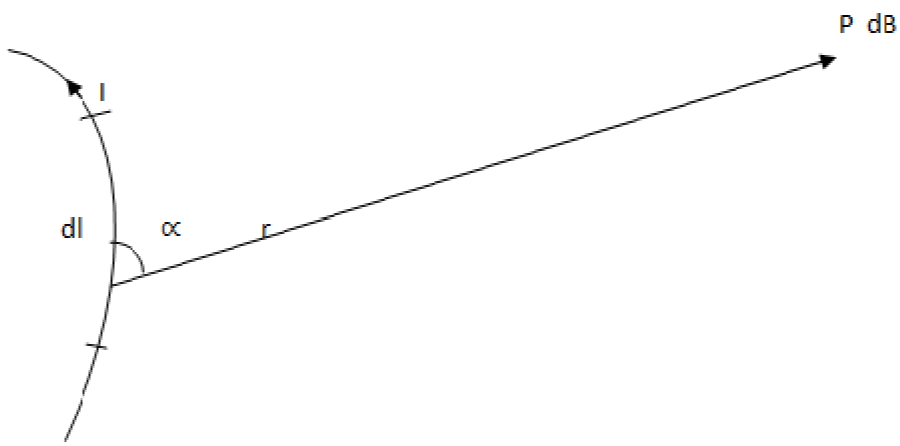
$(30-3)\ 10^5 = 27\ 10^5$ â„¦

MAGNITUDES FOR CURRENT CARRYING CONDUCTORS

Laws of Biot and Savant.

It state that the flux density dB= at point P due to a small element dl of a conductor carrying

Where r is the distance from the point P to the element is the angle formed it to P.



B: FLUX DENSITY = INDUCTION OR MAGNETIC INDUCTION

MAGNETIC FIELD

Equation (1) can be written as

dB=KIdlsinx

Where K is the constant of proportionality and it depends on the medium in which the

conductor is situated. Also K $=\frac{\mu_o}{4\pi}$

$$dB = \frac{\mu_0 Idlsinx}{4\pi r^2} \quad \text{---------------- (ii)}$$

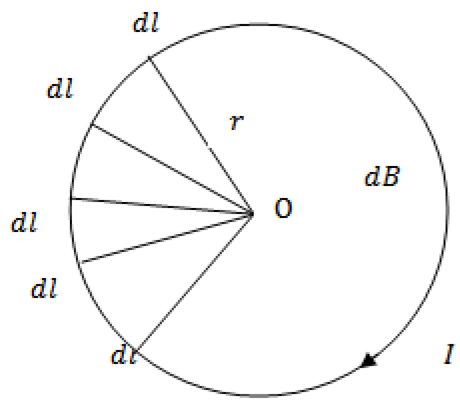
The formula is

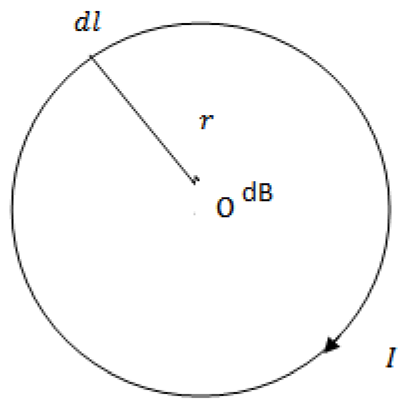
Note that $\mu_o = 4\pi \times 10^{-7}$ Henry m^{-1}

$$B = \frac{\mu_0 I}{2r}$$

B AT THE CENTRE OF A NARROW CIRCULAR COIL

Suppose the coil is in air has a radius of r carries a steady current I and it is considered to consist of current element of length dl. Each element is at the distance r from the centre o and it is at the right angles the line joining it to i.e. $\propto 90^0$





If the coil have N turns then the length of the wire = $2\pi rN$

From the equation (ii) $dB = \frac{\mu_0 Idlsinx}{4\pi r^2}$

Then the magnetic field at the centre is given by

$$B = \frac{u_oI}{4\pi r^2}\int_0^{2\pi rN} dt \;= \frac{u_oI}{4\pi r^2}\;[L]_0^{\;2\pi rN}$$

$$:B = \frac{u_0 IN}{2r} \qquad\qquad B = \frac{u_0 IN}{2r}$$

Example

A coil of wire with 15 turns of radius 6.0cm, has a current of 3.5A flowing through it. What is the magnetic flux density at the center of the coil?

Solution.

$$B = \frac{u_0 IN}{2r} \quad = \frac{\overset{2}{\cancel{4}}\pi \times 10^{-7} \times 3.5 \times \cancel{15}}{\cancel{2} \times \underset{2}{\cancel{6}} \times 10^{-2}} \quad = 175k \times 10^{-5}\text{ Tesla}$$

$$= 5.495 \times 10^{-4}\text{ Tesla}$$

Example

What is the magnitude of the flux density produced the center of a coil of radius 5cm carrying current of 4A in air.

Solution

Given

$I = 4A$

$r = 5 \times 10^{-2}m.$

$N = I$

Formula.

$B = \frac{u_0 IN}{2r} = \frac{4\pi \times 10^{-7} \times 4 \times 1}{2 \times 5 \times 10^{-2}} = 1.6\pi$

$B = 5.024 \times 10^{-5}$ Tesla

Example

A circular coil of radius 6cm consisting of 5 turns carries a current supplied from 2v accumulator of negligible internal resistance. If the coil has a total resistance of 2â„¦. Calculate the magnetic field induced at the centre

Solution.

$I = \frac{V}{R} = \frac{2}{2} = I = 1A$

$N = 5$

$r = 6 \times 10^{-2}m$

$u_o = 4 \times 10^{-7}$

Formula.

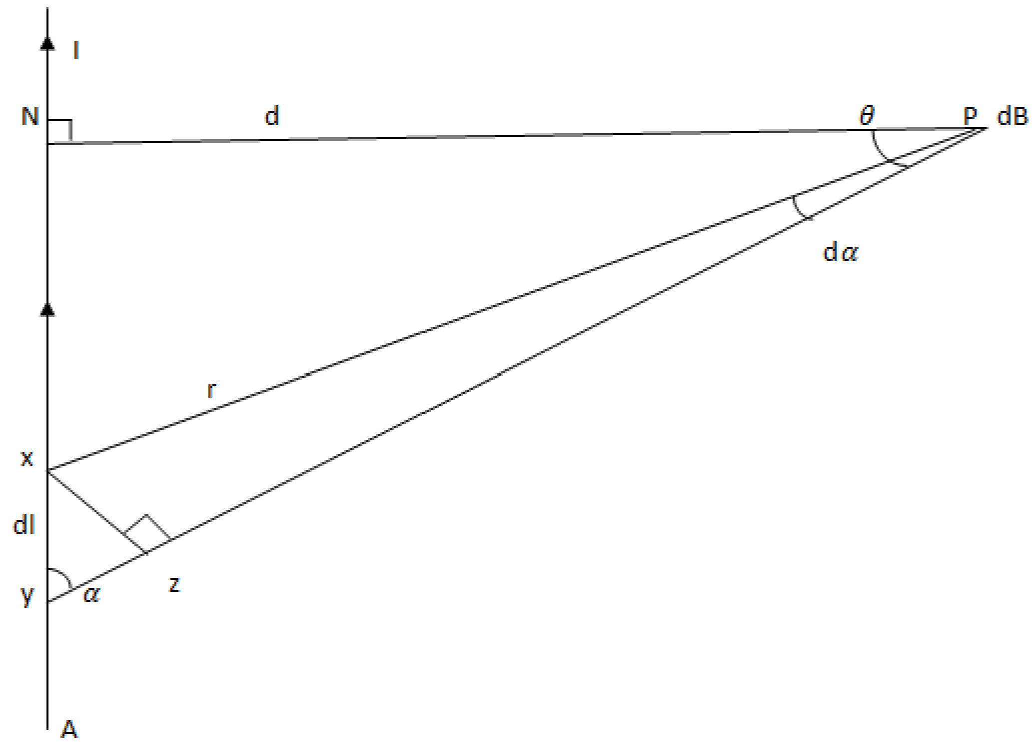
$\frac{u_0 IN}{2r} = \frac{\cancel{4} \times 10^{-7} \times 1 \times 5}{\cancel{2} \times \cancel{6} \times 10^{-2}}$
3

$B = \frac{1}{3} \times 10^{-5} = 3.335 \times 10^{-6}$ Tesla

$= 3.335 \times 10^{-6}$ Tesla

B DUE TO A LONG STRAIGHT WIRE AT A DISTANCE d SIDE THE WIRE

Consider a very long wire YN carrying a current I. Take P to be a point outside the wire but also this point is considered to be very near to this wire.



THE HALL EFFECT

Is the phenomenon where by $e.m.f$ or voltage is set up transversely or across a current carrying conductor when a perpendicular magnetic field is applied

Consider a piece of conducting material in a magnetic field of flux density B

Suppose that the field is directed (perpendicularly) into the paper and that there is a current flowing from right to left.If the material is a metal the current is carried by electrons moving from left to right

Consider the situation of one of these electrons and suppose that it has a velocity V

The electron feels a force F which by Fleming's left hand rule, is directed downwards .

Thus in addition to the electron flow from left to right electrons are urged away from face Y and towards face X. Anegative charged builds up on X, leaving a positive charged on Y so that a potential difference is established between X and Y. The buildup of charge continues until the potential difference becomes so large that it prevents any further increase . This maximum, potential difference is called the Hall voltage

Hall voltage

Is the potential difference created across a current carrying metal strip when the strip is placed in a magnetic field perpendicular to the current flow in the strip.

Actually, the magnetic field does not have to be totally perpendicular to the strip the magnetic field only needs to have a component that is perpendicular

The flow ceases when the e *.m .f* reaches a particular V_H called Hall voltage

MAGNITUDE HALL VOLTAGE

Suppose V_H is the magnitude of the Hall voltage and d is the width of the slab (the separation of x and y). Then the Electric field strength E set up across the slab is numerical equal to the potential gradient.

$$E = \frac{V_H}{d}$$

let F_v be the force exerted on an electron by the P.d between X and Y. Therefore when the buildup of charged on X and Y has ceased

$$F = F_v$$

$$BeV = eE$$

$$BV = E$$

$$BV = \frac{V_H}{d}$$

$$V_H = BVd \text{.......................(i)}$$

Where

E = The strength of the uniform electric field between X and Y due to the Hall voltage

V_H = Hall voltage

d = The separation of X and Y

HALL VOLTAGE

Is the potential difference created across a current carrying metal strip when the strip is placed in a magnetic field perpendicular to the current flow in the strip

Actually the magnetic field does not have to be totally perpendicular to the strip the magnetic field only needs to have a component that is perpendicular

The flow ceases when the e *.m. f* reaches a particular value V_H called Hall voltage. It has been shown that the current I in a material is given by I = neAV

Where

n = the number of electron per unit value

e = the charge on each electrons

v = the drift velocity of the electrons

A = the cross – sectional area of the material

V = $\frac{I}{neA}$ ----------------- (ii)

Sub equation (ii) into equation (i)

From

$$V_H = BVd$$

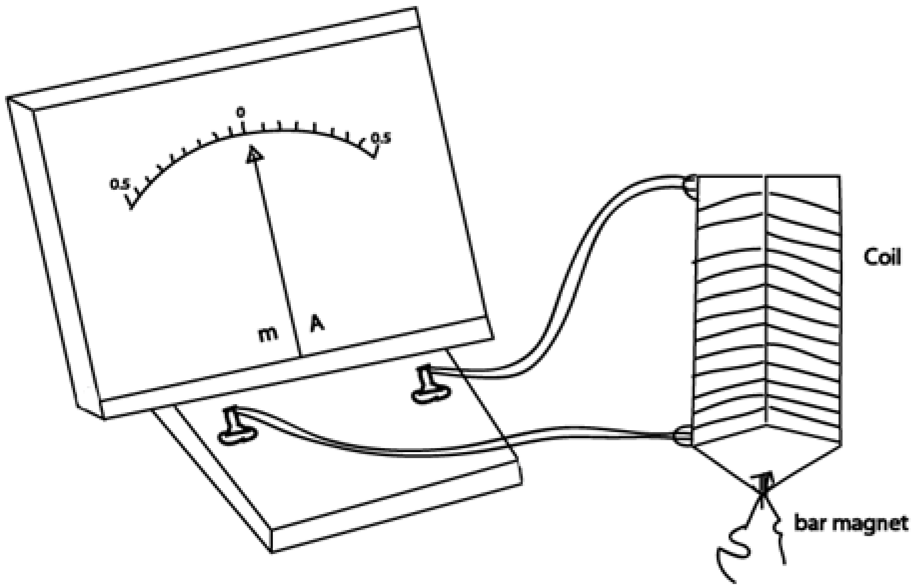
$$V_H = Bd \times \frac{I}{nedt}$$

$$V_H = \frac{BdI}{nedt}$$

In figure A = d $\times$ t and therefore

ELECTROMAGNETIC INDUCTION

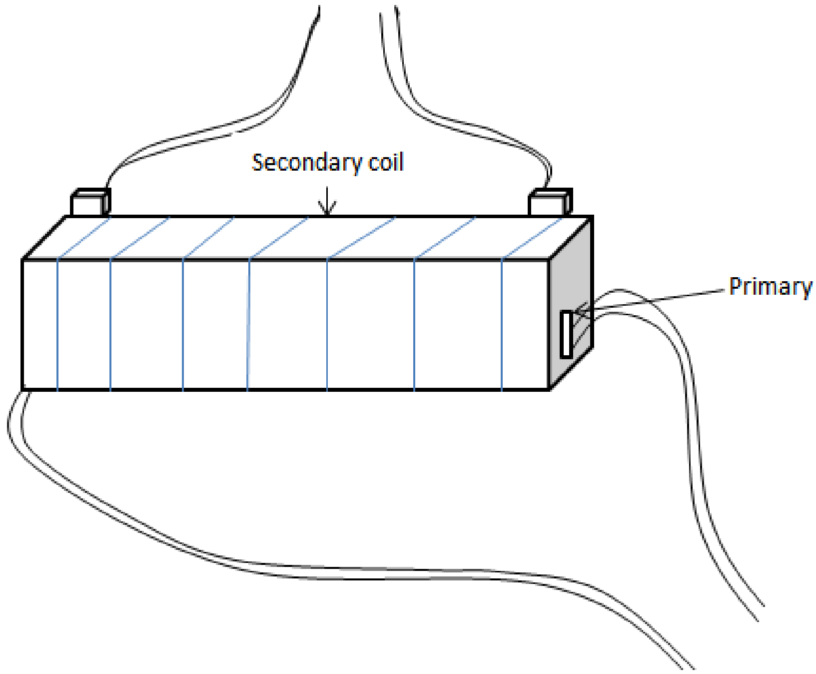
An electric current create magnetic field, the reverse effects of producing electricity by magnetism was discovered by Faraday and is called electromagnetic induction



Induced emf can be generated in two ways

(a) By relative moment (The generator effect)
if the bar magnet is moved in and out of a stationary galvanometer or small current is recorded during the motion but not at other time movement of the coil towards or away from the stationary magnet has the same results (figure above)
relative motion between the magnet and coil is necessary, the direction of the of induced current depends on the direction of the relative motion. And magnitude of current is produced increase with
i. the speed of the motion
ii. The number of turns in the coil
iii. The strength of the magnet used
(b) **By changing a magnetic field (transformer effects)**

In this case two coils are arranged one inside the other (Figure below) to galvanometer



Rheostat the other called secondary is connected to galvanometer, switching the current on or off in the primary causes impulse of emf and current to be induced in the secondary. Varying the primary current quietly altering the value of rheostat has the same effect.

Electromagnetic induction thus they occurs only when there is only change in primary current and also in magnetic fields it induces

LAWS OF ELECTROMAGNETIC INDUCTION

While the magnitude of the induced EMF is given by Faraday law. Its direction can be predicted by Lenz's Law

LENZ'S LAW

The direction of induced emf is such that it tends to oppose the flux change which causing it and does oppose it if induced current flows

Faraday or Newman's laws

The induced emf is directly proportional to the rate of change of the flux through the the coil.

If E = induced emf then

$$E \propto \frac{d\phi}{dt} \text{ for single coil}$$

$$E \propto \frac{d}{dt}(N\phi) \text{ For N coil (turns)}$$

$$E = -k\frac{d}{dt}(N\phi)$$

But the value of k was found to be equal to 1

Therefore,

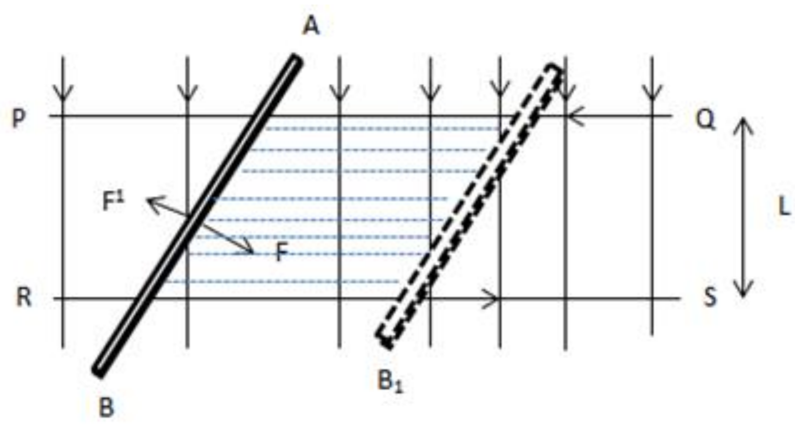
$$E = -\frac{d}{dt}(N\phi)$$

NOTE

I). The minus sign express Lenz's Law

II). NÉ¸ is the flux linkage in the coil

INDUCED *EMF* IN A MOVING ROD



Area swept in 1 second

AB is a wire which can be moved by a force F in a contact with a smooth metal rails PQ and RS. A magnetic field of flux density B acts downwards perpendicular to the plane of the system.

As the wire AB cuts the flux density the emf is produced by the current I and is in opposition to the motion

Therefore

F= BIL i

Where l is the distance between two rails

And I = $\frac{E}{\gamma}$ ii

Where γ is the resistance of the wire

If the wire is moving with a speed V then F' = F iii

F' = $\frac{BEL}{\gamma}$ 4

Power = $\frac{workdone}{time}$ = $\frac{F.d}{t}$ = Force x velocity

= $\frac{BELV}{\gamma}$ 5

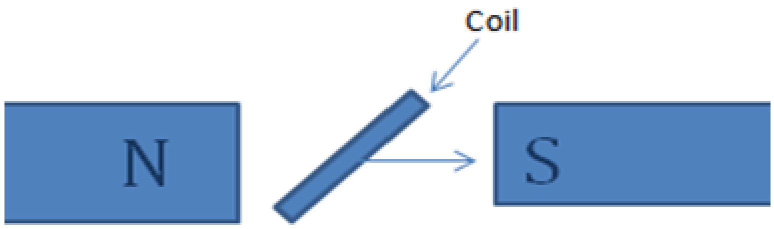
Also power = $T^2\gamma$ = $\frac{E^2}{\gamma}$ $\cdot\gamma$ = $\frac{E^2}{\gamma}$ 6

Equating equation 5 and 6

$$\frac{BELV}{\gamma} = \frac{E^2}{\gamma}$$

I.e. E = BLV (This is the induced emf in a moving coil)

INDUCED *EMF* IN A ROTATING COIL



Consider a coil of an area A and its normal makes an angle of θ with the magnetic field B_Y

The flux linkage with the coil of n turns is expressed as

$N\Phi = NAB\cos\theta$ ……………………………………………………1

The induced emf is given by

$E = \frac{d}{dt}(N\Phi) = -\frac{d}{dt}(N\Phi) = -\frac{d}{dt}(NAB\cos\theta) = -NAB\frac{d}{dt}\cos\theta$

$E = NABw\sin wt$ since $\theta = wt$

If the maximum value of emf is denoted by ε_o

Then

$E = E_o \sin wt$ where $E_o = NABw$

A gain $w = 2f\pi$

Therefore

1. $E = NAB(2\pi f)\sin(2\pi ft)$

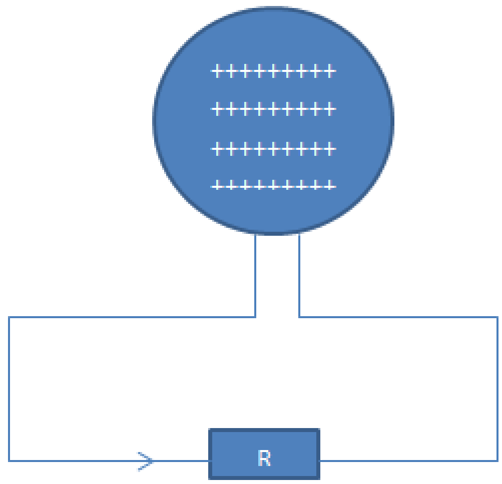
2. $E_o = 2f\pi NAB$

Exercise 1

The magnetic flux Q_B through the loop perpendicular to the plane of the coil and directed into the paper as shown in the diagram is varying according to the equation $Q_B = 8t^2 + 5t + 5$ where Q_B is measured in millimebers and t in seconds

i. What is the magnitude of induced emf in the loop when $t = 3seconds?$

ii. What is the direction of the current through R?



Solution

$$E = \frac{d}{dt}(8t^2 + 5t + 5)$$

E = 16t + 5

E = 53Mv

Exercise 2

What is the maximum emf induced in a coil of 500turns, each with an area of 4.0cm^2, which makes 50reflections per second in a uniform magnetic field of flux density 0.04T?

Solution

B = 0.04T

4.0cm^2

$\text{A} = 4.0\ \text{x}10^{-4}\text{m}^2$

$f = 50$

$E = 2f\pi NAB$

$= 2x3.14x500x4.0\ \text{x}10^{-4}x0.04x50$

$\text{E} = 25.1\ \text{x}10^{-1}$

2.5Volts

INDUCED EMF IN ROTATING DISC – DYNAMO

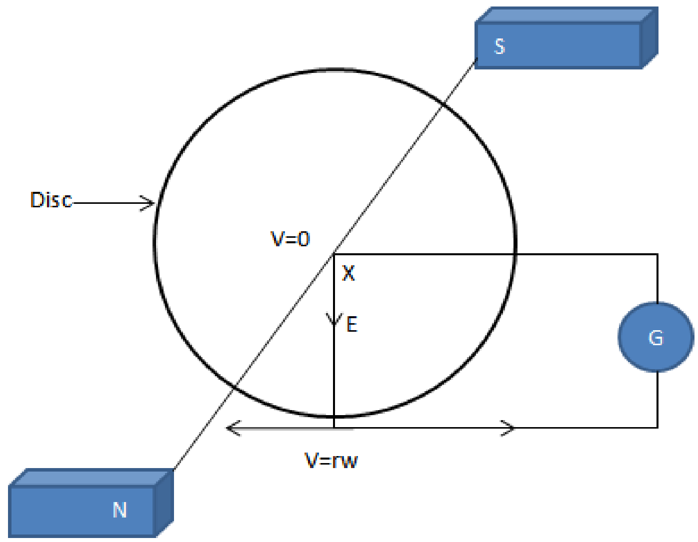
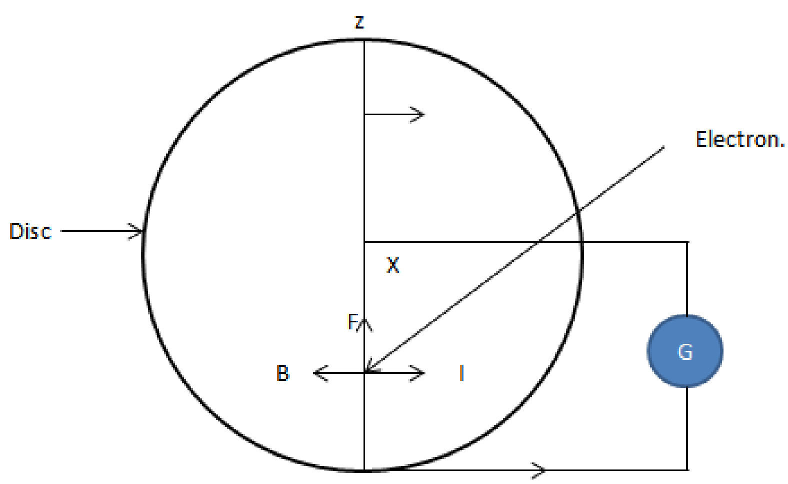


Fig i)



Consider a copper disc which rotates between poles of magnets. Connections are made to its circle and the circumference. An induced emf is obtained between the Centre of the disc and one edge. We assume that magnetic field is uniform over the radius xy

The radius xy continuously cuts the magnetic flux between the poles of the magnet. For this straight conductor, the velocity at the end of x is zero and that at the other end y xy where w is the angular velocity of the disk

Average velocity of xy is $v = \frac{1}{2}(0 + rw) = \frac{rw}{2}$

An induced emf in straight conductor is given by

$E = Blv$ In this case $l = r \; and \; v = \frac{rw}{2}$

$v = Br\left(\frac{rw}{2}\right) = \frac{1}{2}Br^2w$ ……………………………………………………i)

Since $E = B\pi \mathbf{r}^2 f$ ……………………………………...ii)

If the disc has the radius r_1 and an axle at the Centre of radius r_2 the area swept out by a rotating radius of the metal disc is $\pi r_1{}^2 - \pi r_2{}^2 = \pi(r_1{}^2 - r_2{}^2)$ in this case the induced emf would be

$E = B\pi(r_1{}^2 - r_2{}^2)f$

The direction of the E is given by Fleming's right hand rule

As the disc rotates clockwise the radius xy moves to the left at the same time as the radius xz moves to right

If the magnetic field covers the whole disk, induced emf in the two radii would be in opposite direction. So the resultant emf between yz would be zero. $B\pi \mathbf{r}^2 f$ The emf between the Centre and the rim of the disc is the maximum emf which can be obtained

Qn.

A circular metal disc with a radius of 10cm rotates at 10revolutions per seconds. If the disc is in a uniform magnetic field of 0.02T at a right angle to the plane of the disc. What will be the emf induced between the Centre and the rim of the disc?

Solution

B = 0.02T

$r = 10\ x10^{-2}m^2$

$f = 10$

$E = B\pi r^2 f$

$E = 0.02x3.14x(10x10^{-2})^2x10$

$E = 6.28x10^{-3}v$

SELF INDUCTANCE (L)

An induced emf appear in the coil if the current in that coil is changed is called self-induction and emf produced is called self-induced emf

For a given coil produced no magnetic materials nearly the flux linkage $N\emptyset$ proportional to the current I

$N\emptyset\ \alpha\ I$ Or $N\phi = LI \ldots \ldots \ldots \ldots \ldots \ldots \ldots .i$

Where L is a constant proportionality which is called self-inductance of a coil

From Faraday's law in such a coil emf the induced

$E = -\frac{d}{dt}(N\phi)^{-2}$

Substitute i) in ii)

$E = -\frac{d}{dt}(LI) = -L\frac{dI}{dt}$ or

$L = -\frac{dE}{\frac{dI}{dt}}$

Hence the unit of inductance vsA^{-1}. A special name the Henry has been given to this combination of units

Two coils A and B have 200 and 800turns respectively. A current 2Amperes in A produces a magnetic flux of $1.8x10^{-4}wb$ in each turn of A, compute:

i. Mutual inductance

ii. Magnetic flux through A when there is a current of 4.0 Ampere in B

iii. The emf induced when the current in A changes 3A to 1A in 0.2seconds

SELF INDUCTANCE (L) FOR THE COIL

The induced emf, $E = -\frac{d\phi}{dt} = L\frac{dI}{dt}$

$d\phi = LdI$ (By integrating the equation we have)

$\phi = LI$

$\phi = LI\frac{dI}{dt}$

Therefore $L = \frac{\phi}{I}$

The self-inductance may be defined as the flux linkage per unit current, when Φ is in wabers and I is in amperes then L is in henry:

Magnetic flux density for a long coil is given by μ_{oNI} with an iron core with a relative permeability of μ_r

The flux density is given by $B = \frac{\mu_0 \mu_r \mathrm{N}^2 \mathrm{AI}}{L}$ since $\mu_r = \frac{\mu}{\mu_0}$

Thus the flux linkage $\Phi = NAB = \frac{\mu_0 \mu_r \mathrm{N}^2 \mathrm{AI}}{L}$

$L = \frac{\Phi}{I} = \frac{\mu_0 \mu_r \mathrm{N}^2 \mathrm{A}}{L}$ (Unit for L is Henry)

ENERGY STORED IN AN INDUCTOR

Because of emf of the self-induction that act when the current in the coil change, electrical energy must be supplied in setting up the current against the f.

If L is the self-inductance of the inductor then the back emf across it is given by

$\boldsymbol{E = L\frac{dI}{dt}}$**……………………….i)**

Hence rate at which work is done against the backward emf.

Power = EI……………………….ii)

Substitute equation i) into ii)

Then equation ii) becomes

$p = LI\frac{dI}{dt}$

The work done to bring the current from zero to a steady state value I_o is

$\mathrm{W} = \int_0^t pdt = \int_0^I LIdt.dt = L\int_0^I IdI$

$[I^2]_0^I = \frac{L}{2}[I^2 - 0] = \frac{L}{2}I^2$

Therefore

$W = \frac{1}{2}LI_o^2$

MUTUAL INDUCTANCE (M)

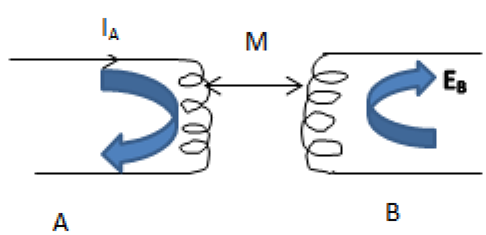
The emf may be induced by in one circuit by changing current in another. This phenomenon is often called mutual induction and the pairs of circuits which shows it are said to have mutual inductance

The mutual inductance m between the two circuits is defined by the following equation

emf Induced in B by changing = M (rate of change of current in A) i.e.

$E_B = M\frac{dI_A}{dt}$

The unit of mutual inductance is Henry the same as that of self-inductance



MUTUAL INDUCTION

Since $E_B =$ the rate of change in flux in B then

$E_B = \frac{d\phi}{dt} = \frac{mdI}{dt}$

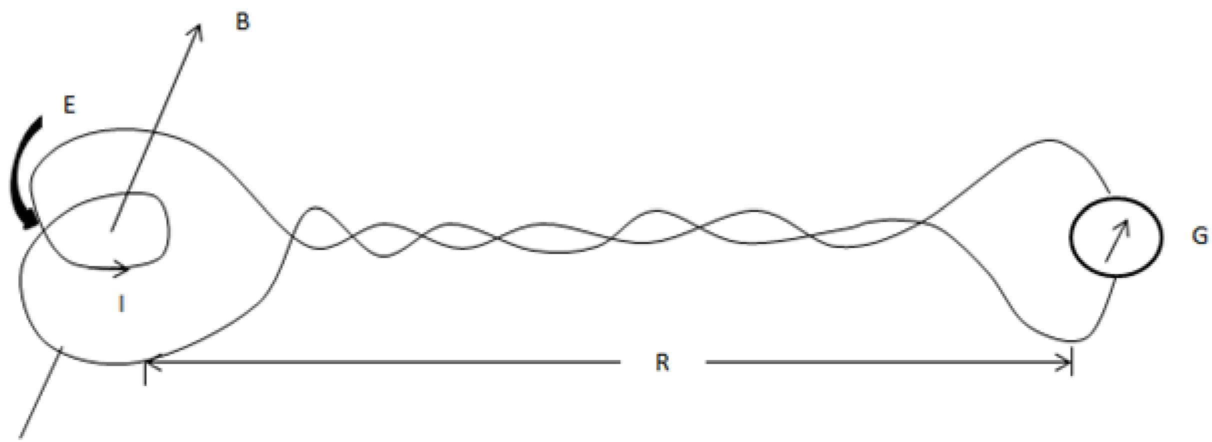
$$M = \frac{d\phi_B}{dI_A} = \frac{\textit{Flux change in B}}{\textit{current change in A}}$$

QUANTITY OF ELECTRICITY INDUCED

Consider a close circuit of total resistance R Ohms which has a total flux linkage Φ with magnetic field B. if the flux linkage starts to change

Induced emf, $E = \frac{d\Phi}{dt}$ but current

$I = \frac{E}{R} = \frac{1}{R}\frac{d\Phi}{dt}$



Flux linkage will not change at a steady rate and a current will not be constant. But throughout it changes. Its charge is being carried round the circuit. If a time t seconds is taken to reach a new constant value the charge carried round the circuit in that time is

$Q = \int_o^t Idt$. $from\ equation\ i)\ we\ have$

$Q = \frac{1}{R}\int_o^t \frac{d\emptyset}{dt} .dt = \frac{1}{R}\int_{\emptyset o}^{\emptyset t} d\emptyset$

Where $\emptyset_o$ is the number of linkage at t=o and $\emptyset_t$ is the number of linkage time t

Thus $\frac{(\emptyset_t - \emptyset_o)}{R} = \frac{(\emptyset_o - \emptyset_t)}{R}$

$$Q = \frac{Change\ of\ flux\ linkage}{R}$$

AC THEORY

When a battery is connected to a circuit the current flows steadily in one direction, this is called a Direct current (d.c).

The use of Direct currents is limited to a few applications e.g charging of batteries, electroplating etc.

Most of electrical energy is generated and used in the form of alternating current due to many reasons including,

i) Alternating voltages can be changed in value very easily by means of transformers.

ii) A.c motors are simpler in construction and cheaper than d.c motors.

ALTERNATING VOLTAGE AND CURRENT

i) Alternating voltage

An alternating voltage is one whose magnitude changes with time and direction reverses periodically. The instantaneous value (i.e value at any time t) of an alternating voltage is given by,

$$E = E_0 \sin wt$$

where,

E = Value of the Alternating voltage at time t

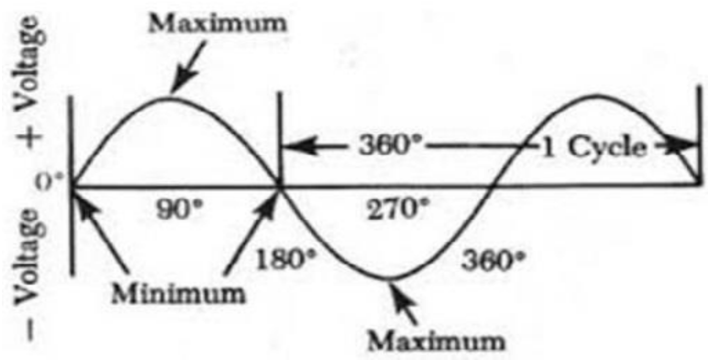
E_0 = Maximum value of the Alternating voltage

ω = Angular frequency of supply

From $\omega = 2\pi f$,where f is the frequency of the alternating voltage, If T is the time period of alternating voltage then

$$\omega = 2\pi f$$

$$\omega = \frac{2\pi}{T}$$



The voltage varies from zero to a positive peak ($^{+}E_0$) then back via zero to negative peak ($-E_0$) and so on.
In time period T, the wave completely cycle.

$$E_o = \frac{E}{\sin \omega t}$$

ii) (ii)Alternating current

This is one whose magnitude changes with time and direction reverses periodically.

The Instantaneous value ($i.e$ value at any time t) of sinusoidally varying alternating current is given by

$$I = I_0 \sin \omega t$$

where

I = value of alternating current at time t

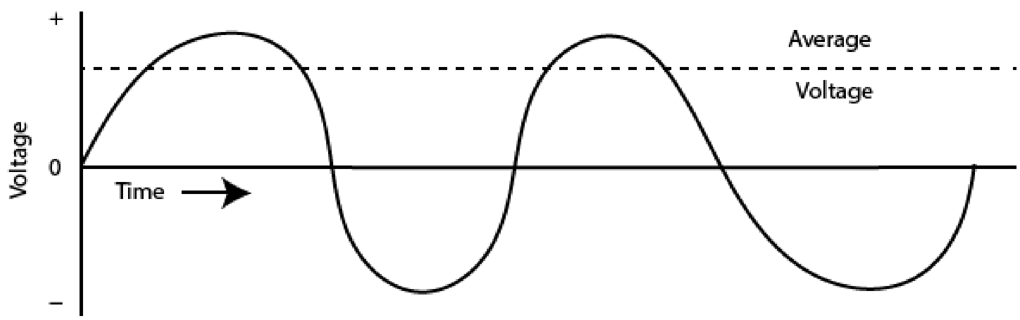
I_0 = maximum value (Amplitude) of alternating current.

ω = Angular frequency of supply.

$$\omega = 2\pi f$$

$$\omega = \frac{2\pi}{T}$$

Figure below shows the waveform of alternating current.



Current varies sinusoidally with time. The current increases gradually from zero to a positive peak ($^{+}I_o$), then back via zero to a negative peak ($^{-}I_o$) and so on.

In time period T, the wave completes a cycle.

$$E_0 = \frac{E}{\sin \omega t}$$

An Alternating voltage and current also be represented as a cosine function of time.

$$E = E_0 \cos \omega t$$

$$I = I_0 \cos \omega t$$

Both these representations give the same result as is given by the one containing sine functions.

MEASUREMENT OF ALTERNATING CURRENT

Since the average value of sinusoidal alternating current is zero, an ordinary (DC) ammeter or galvanometer will not show any deflection when connected in an AC circuit.

Due to inertia it will not be possible for the needle to oscillate with the frequency of the current.

Therefore to measure AC we use hot wire instruments because the heating effect of current is independent of the direction of current.

MEAN OR AVERAGE VALUE OF ALTERNATING CURRENT

The mean value or average value of alternating current over one complete circle is zero.

It is because the area of positive half cycle is exactly equal to the area of the negative half cycle.

However, we can find the average or mean value of alternating current over any half cycle.

Half Cycle Average Value of a.c

This is that value of steady current (d.c) which would send the same amount of charge through a circuit for half the time period of a.c $i.e\ \frac{T}{2}$ as it sent by the a.c through the same circuit in the same time.It is represented by I_m or I_{av}

The instantaneous value of alternating current is given by, $I\ = I_0\, sin\ \omega t$

Suppose current I remains constant for a small time dt. Then small amount of charge sent by alternating current in a small time dt is given by

$$dQ = Idt$$

$$dQ = I_0 \sin \omega t dt$$

If Q is the total charge sent by the positive half cycle of a.c ($i.e\ 0\ to\ {}^{T}/_{2}\ \ time), then$

$$\int_0^{Q/2} dQ = \int_0^{T/2} I_0 \sin wtdt$$

$$Q = I_0 \int_0^{T/2} \sin \omega t\ dt$$

$$Q = I_0 \left[\frac{-\cos \omega t}{\omega}\right]_0^{T/2}$$

$$Q = -\frac{I_0}{\omega}[\cos \omega t]_0^{T/2}$$

$$Q = -\frac{I_0}{\omega} \quad \left[\cos \frac{\omega t}{2} - \cos 0\right]$$

$$Q = -\frac{I_0}{\omega}\left[\cos\left(\frac{2\pi}{T}.\frac{T}{2}\right) - \cos 0\right]$$

$$Q = -\frac{I_0}{\omega}[\cos \pi - \cos 0]$$

$$Q = -\frac{I_0}{\omega}[-1-1]$$

$$Q = \frac{2I_0}{\omega} \dots \dots \dots \dots \dots \dots . . (i)$$

If the I_m is the half cycle average means values of positive half cycle of a.c then by definition

$$Q = I_m \times \frac{T}{2} \ldots\ldots\ldots\ldots\ldots\ldots\ldots..(ii)$$

From equation (i) and (ii)

$$I_m \times \frac{T}{2} = \frac{2I_0}{\omega}$$

$$I_m = \frac{2I_0}{\pi}$$

$$I_m = 0.637 I_0$$

Hence half cycle average value of a.c is 0.637 times the peak value of a.c

For positive ½ cycle $I_m = +0.637\ I_0$

For negative half cycle $I_m = -0.637\ I_0$

Obviously, average value of a.c over a complete cycle is zero.

MEAN OR AVERAGE VALUE OF ALTERNATIVE VOLTAGE

Half Cycle Average value of alternating $e.m.f$

This is that value of steady $e.m.f$, ($d.c\ \ e.m.f$) which would send the same amount charge through a circuit for half the time period of alternating $e.m.f$ $\left(\frac{T}{2}\right)$ as is sent by the alternating $e.m.f$ through the same circuit in the same time.It is denoted by $E_m\ and\ E_{av}$ The instantaneous value of alternating $e.m.f$ is given by $E = E_0\ sin\ \omega t$

Suppose this alternating $e.m.f$ is applied to a circuit of resistance R. Then by ohm's law the instantaneous value of alternating current is

$$I = \frac{E}{R}$$

$$I = \frac{E_0 \sin \omega t}{R}$$

If this current remains constant a small time dt, then small amount of charge send by alternating $e.m.f$ in small time dt is given by

$dQ = Idt$

$$dQ = \frac{E_0}{R} \sin \omega t \ dt$$

If Q is the total charge sent by positive half cycle of a alternating $e.m.f$ then

$$\int_0^{Q/2} dQ = \int_0^{T/2} \frac{E_0}{R} \sin wtdt$$

$$Q = \frac{E_0}{R} \int_0^{T/2} \sin \omega t \ dt$$

$$Q = \frac{E_0}{R} \left[\frac{-\cos \omega t}{\omega} \right]_0^{T/2}$$

$$Q = -\frac{E_0}{R\omega} [\cos \omega t]_0^{T/2}$$

$$Q = -\frac{E_0}{R\omega} \left[\cos \frac{\omega T}{2} - \cos 0 \right]$$

$$Q = -\frac{E_0}{R\omega} \left[\cos \frac{2\pi}{T} . \frac{T}{2} - \cos 0 \right]$$

$$Q = -\frac{E_0}{R\omega} [\cos \pi - \cos 0]$$

$$Q = -\frac{E_0}{R\omega} [-1 - 1]$$

$$Q = \frac{2E_0}{R\omega} \dots \dots \dots \dots \dots \dots \dots \dots . (i)$$

If E_m is the half cycle average or mean value of the positive half cycle of alternating $e.m.f$ then by definition

$$Q = \frac{E_m}{R} \times \frac{T}{2} \dots \dots \dots \dots \dots \dots \dots . . (ii)$$

From equation (i) and equation (ii)

$$\frac{E_m}{R} \times \frac{T}{2} = \frac{2E_0}{\omega R}$$

$$\frac{E_m}{R} \times \frac{T}{2} = \frac{2E_0}{R} \times \frac{T}{2\pi}$$

$$E_m = \frac{2}{\pi} E_0$$

$$E_m = 0.637E_0$$

Therefore, half cycle average value of alternating $em.f$ is 0.637 times the peak value of alternating $e.m.f$.

For positive half cycle

$$E_m = +0.637E_0$$

For negative half cycle

$$E_m = -0.637E_0$$

A d.c voltmeter or ammeter reads average (or d.c) value. Therefore, they can be used to measure alternating voltage on current.

It is because the average value of alternating voltage or current over a complete cycle is zero.

We use a.c meters to measure alternating voltage/current.

ROOT MEAN SQUARE VALUE OF ALTERNATING CURRENT

The average value cannot be used to specify an alternating current (or voltage).

It is because its value is zero over one cycle and cannot be used for power calculations.

Therefore we must search for more suitable criterion to measure the effectiveness of an alternating current or voltage.

The obvious choice would be to measure it in terms of direct current that would do work (or produce heat) at the same average rate as a.c under similar conditions.

This equivalent direct current is called the root mean square ($\mathrm{r.m.s}$) or effective value of alternating current.

Effective or value of Alternating Current

The root mean square (r.m.s) of alternating current is that steady current (d.c) which when flowing through a given resistance for given time produces the same amount of heat as produced by the alternating current when flowing through the same resistance for the same time.

- It is also called virtual value of a.c
- It is denoted by called $I_{r.m.s}$ or I_{eff} or I_v

For example, when we say that $r.m.s$ or effective value of an alternating current is 5A, it means that the alternating current will do the work (or produce heat) at the same rate as 5A direct current under similar conditions.

RELATION BETWEEN R.M.S. VALUE AND PEAK VALUE OF A.C

Let the alternating current be represented by

$$I = I_0 \sin \omega t$$

If this alternating current flows through a resistance R for a small time dt, then small amount of heat produced is given.

$$dH = I^2 R dt$$

$$dH = (I_0 \sin \omega t)^2 R dt$$

$$dH = I_0^2 R sin^2 \omega t \ dt$$

In one complete cycle $i.e$ for time 0 to T of alternating current the total amount of heat produced in R is given by

$$\int_0^H dH = \int_0^T I_0^2 \ R \sin^2 wt \ dt$$

$$H = I_0^2 R \int_0^T sin^2 \ \omega t \ dt$$

$$H = I_0^2 R \int_0^T \left(\frac{1 - Cos\ 2\omega t}{2}\right) dt$$

$$H = \frac{I_0^2 R}{2} \int_0^T (1 - Cos\ 2\omega t)\ dt$$

$$H = \frac{I_0^2 R}{2} \int_0^T dt - \int_0^T \cos 2\omega t\ dt$$

$$H = \frac{I_0^2 R}{2} \left[t - \frac{\sin 2\omega t}{2\omega}\right]_0^T$$

$$H = \frac{I_0^2 R}{2} \left[(T - 0) - \left(\frac{\sin 2\omega T}{2\omega} - \sin 0^\circ\right)\right]$$

$$H = \frac{I_0^2 R}{2} \left[T - \frac{\sin 2\omega T}{2\omega}\right]$$

$$H = \frac{I_0^2 R}{2} \left[T - \sin 2.\frac{2\pi}{T} T \Big/ 2\omega\right]$$

$$H = \frac{I_0^2 R}{2} \left[T - \frac{\sin 2 \times 2\pi}{2\omega}\right]$$

$$H = \frac{I_0^2 R}{2} [T - 0]$$

$$H = \frac{I_0^2 RT}{2} \dots\dots\dots\dots\dots\dots\dots\dots\dots (i)$$

If I_{rms} is the virtual or $r.m.s$ value of the alternating current, then heat produced in R in the same time $i.e$ (0 to T) is given by

$$H = I_{rms}^2\ RT$$

From equation (i) and (ii), we have

$$I_{rms}^2 = \frac{I_0^2 RT}{2}$$

$$I_{rms}^2 = \frac{I_0^2}{2}$$

$$I_{rms} = \frac{I_0}{\sqrt{2}}$$

$$I_{rms} = 0.707 I_0$$

Hence the $\mathrm{r.m.s}$ value of effective value or virtual value of alternating current is 0.707 times the peak value of alternating current.

The $\mathrm{r.m.s}$ value is the same whether calculated for half cycle or full cycle.

Alternative Method

Let the alternating current be represented by

$$I = I_0 \sin \omega t$$

If this current is passed through a resistance R, then power delivered at any instant is given by

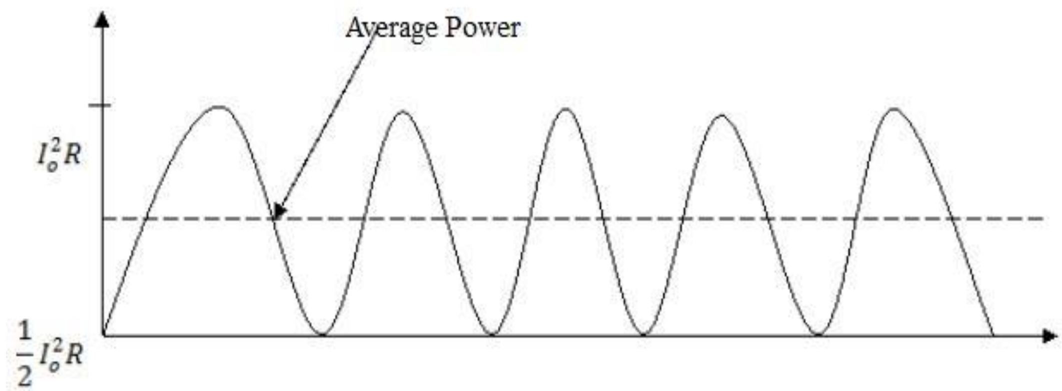
$$P = I^2 R$$

$$P = (I_0 \sin \omega t)^2 R$$

$$P = I_0^2 R sin^2 \ \omega t$$

Because the current is squared, power is always positive since the value of $sin^2 \omega t$ varies

between 0 and 1



Average power delivered

$$P = \frac{1}{2} I_0^2 R \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots. (i)$$

If I_{rms} is the virtual or $r.m.s$ value of the alternating current then by definition

Power delivery P

$$P = I_{rms}^2 R \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots (ii)$$

From equation (I) and (ii)

$$I_{rms}^2 \text{ R} = \frac{1}{2} I_0^2 R$$

$$I_{rms}^2 = \frac{I_0}{2}$$

$$I_{rms} = \frac{I_0}{\sqrt{2}}$$

$$I_{rms} = 0.707 I_0$$

ROOT MEAN SQUARE VALUE OF ALTERNATING E.M.F

The root mean square (r.m.s) value of alternating voltage is that steady voltage (d.c voltage when applied to a given resistance for a given time produces the same amount of heat as is produced by the alternative $e.m.f$ when applied to the same resistance for the same time.

It is also called virtual value of alternating e.m.f and is donated by $E_{r.m.s}$ of E_{eff} or E_v
The instantaneous value of alternating $e.m.f$ is given by

$$E = E_0 \sin \omega t$$

When this alternating voltage is applied to a resistance R, small amount of heat produced in a small time dt is

$$dQ = I^2 Rdt$$

$$dQ = I^2 Rdt$$

But

$$I = \frac{E}{R}$$

Then,

$$dH = \frac{E^2}{R^2}.Rdt$$

$$dH = \frac{E^2}{R}dt$$

$$dH = \frac{(E_0 \sin \omega t)^2}{R}dt$$

$$dH = \frac{E_0^2 \, sin^2 \, \omega t}{R}dt$$

$$dH = \frac{E_0^2}{R} sin^2 \, \omega t \, dt$$

In one complete cycle $i.e$ for time 0 to T the total amount of heat produced in resistance R is

$$\int_0^T dH = \int_0^T \frac{E_0^2}{R} \sin^2 \omega t \, dt$$

$$H = \frac{E_0^2}{R} \int_0^T \sin^2 \omega t \, dt$$

$$H = \frac{E_0^2}{R} \int_0^T \frac{(1 - \cos 2\omega t)}{2} dt$$

$$H = \frac{E_0^2}{2R} \int_o^T (1 - \cos 2\omega t) dt$$

$$H = \frac{E_0^2}{2R} \left[\int_0^T dt - \int_0^T \cos 2\omega t \, dt \right]$$

$$H = \frac{E_0^2}{2R} [t]_0^T$$

$$H = \frac{E_0^2 T}{2R} \ldots \ldots \ldots \ldots \ldots \ldots \ldots \ldots \ldots . (i)$$

If E_{rms} is the $r.m.s$ value of the alternating $e.m.f$, then heat produced in the same resistance R for the same time T is

$$H = \frac{E_v^2}{R} T \ldots \ldots \ldots \ldots \ldots \ldots \ldots \ldots (ii)$$

From the equation (i) and equation (ii)

$$E_{rms}^2 = \frac{E_0^2 T}{2R}$$

$$E_{rms}^2 = \frac{E_0^2}{R}$$

$$E_{rms} = \frac{E_0}{\sqrt{2}}$$

$$E_{rms} = 0.707E_0$$

Hence the $r.m.s$ value of alternating $e.m.f$ is 0.707 times the peak value of the alternating $e.m.f$

Therefore, $r.m.s$value is the same whether calculated for half cycle or full cycle.

Importance of r.m.s values

An alternating voltage or current always specified in terms of $r.m.s$ values.

Thus an alternating current of 10A is the one which has the same heat effect as 10A d.c under similar conditions. The following points may be noted carefully.

i) The domestic a.c supply is 230V, 50H. It is the $r.m.s$ or effectively value. It means that the alternative voltage available has the same effect as 230 V d.c under similar conditions.

The equation of this alternative voltage is

$$E = E_0 \sin \omega t$$

$$E = \sqrt{2}\, E_{rms} \sin wt$$

$$E = 230\sqrt{2} \sin 2\pi \times 50 \times t$$

$$\therefore 230\sqrt{2} \sin 314t$$

ii) When we say that alternating current in a circuit is 5A, we specifying the $r.m.s$ value.

- It means that the alternating current flowing in the circuit has the same heating effect as 5A d.c under similar conditions.

iii) A.C ammeters and voltmeters record $r.m.s$ values of current and voltage respectively.

The alternating voltage/current can be measured by utilizing the heating effect of electric current.

Such Instruments are called hot wire instruments and measure the $r.m.s$ value of the voltage/current since r.m.s value is the same for half cycle or complete cycle.

WORKED EXAMPLES

1. An a.c main supply is given to be 220V what is the average $e.m.f$ during a positive half cycle?

Solution

$$E_{rms} = \frac{E_0}{\sqrt{2}}$$

$$E_{rms} = \sqrt{2}\,.E_v$$

$$E_{rms} = \sqrt{2}.220$$

$$\therefore E_0 = 311V$$

Average $e.m.f$ during positive half is given by

$$E_{mean} = \frac{2E_0}{\pi} = \frac{2 \times 311}{\pi}$$

$$\therefore\ E_{mean} = 198V$$

2. An alternating current I is given by

$$I = 141.4 \sin 314t$$

Find

(i) The maximum value

(ii) Frequency

(iii) Time period

(iv) The instantaneous value when t= 3ms

Solution

Comparing the given equation of the alternating current with the standard form

$$I = I_0 \sin \omega t$$

(i) Maximum value, $I_0 = 141.4\ A$

(ii) Frequency f

$\omega = 314$

$2\pi f = 314$

$f = \dfrac{314}{2\pi}$

f = 50 HZ

(iii) Time period T

$T = \dfrac{1}{f} = \dfrac{1}{50}$

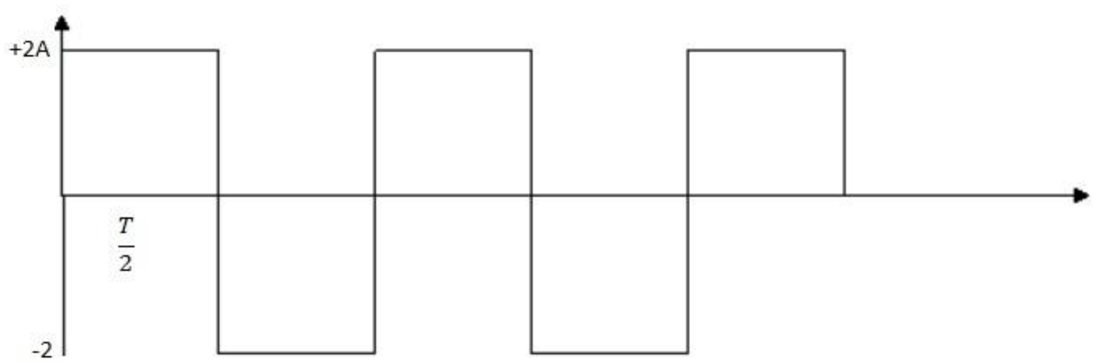
$\therefore T = 0.02\ s$

(iv) $I = 141.4 \sin 314t$

When $t = 3 \times 10^{-3}\ s$

$I = 141.4 \sin(314 \times 3 \times 10^{-3})$

3. Calculate the $r.m.s$ value of the current shown in figure below



Solution

$$I_{r.m.s} = \sqrt{\frac{Area\ of\ half\ cycle\ of\ squared\ wave}{Half\ cycle\ base}}$$

$$I_{r.m.s} = \sqrt{\frac{2^2 \times {}^{T}/_{2}}{{}^{T}/_{2}}}$$

$$I = \sqrt{4}$$

$$\therefore I_{r.m.s} = 2A$$

4. An a.c voltmeter records 50V when connected across the terminals of sinusoidal power source with frequency 50Hz. Write down the equation for the instantaneous voltage provided by the source.

Solution

An a.c voltmeter records $r.m.s$ values

Now

$$E = E_0 \sin \omega t$$

Here

$$E_0 = \sqrt{2}\, E_{rms} \qquad \omega = 2\pi f$$

$$E_0 = \sqrt{2}.50 \qquad \omega = 2\pi \times 50$$

$$E_0 = 70.7V \qquad \omega = 314$$

$$\boldsymbol{E = 70.7 \sin 314t}$$

5. An alternating voltage of 50 Hz has maximum value of $200\sqrt{2}$ volts. At what time measured from a positive maximum value will the instantaneous voltage be 141.4 volts.

Solution

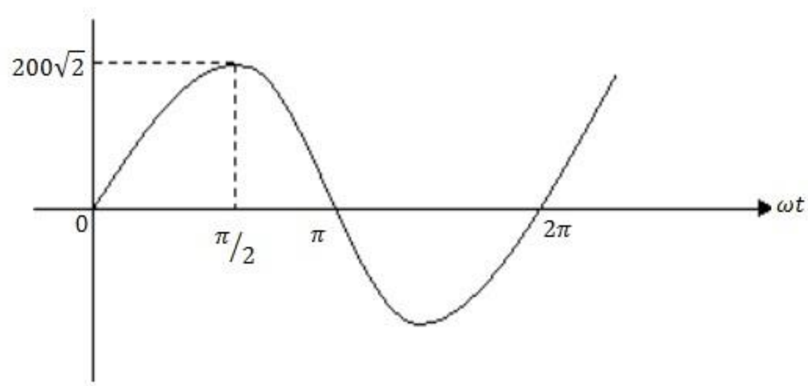
$$E = E_0 \sin \omega t$$

$E = E_0 \sin 2\pi f t$

$E = 200\sqrt{2} \sin 314t$

$E = 282.8 \sin 314t$

This equation is valid when time is measured from the instant the voltage is zeroi.e point O. Since the time is measured from the positive maximum value at point A.



The above equation is modified to

$E = 282.8 \sin\left(314t + \frac{\pi}{2}\right)$

$E = 282.8 \cos 314t$

Let the value of voltage become 141.4 volts t second after passing then the maximum positive value.

$141.4 = 282.8 \cos 314t$

$\cos 314t = \frac{141.4}{282.8}$

$\cos 314t = 0.5$

$314t = cos^{-1}\,(0.5)$

$314t = \frac{\pi}{3}$

$t = \frac{\pi}{3 \times 314}$

$t = 3.33 \times 10^{-3}\ s$

$\therefore \boldsymbol{t = 3.33\ ms}$

6. An alternating current of frequency 60 Hz has a maximum value of 120 A.

i) Write down the equation for the instantaneous value.

ii) Recording time from the instant the current is zero and becoming positive. Find the instantaneous value after 1/360 second.

iii) Time taken to reach 96 A for the first time.

Solution

i) The instantaneous value of alternating current is given by

$I = I_0 \sin \omega t$

$I = I_0 \sin 2\pi f t$

$\mathrm{I} = 120 \sin(2\pi \times 60)\,\mathrm{t}$

$\mathrm{I} = 120\ sin\ 120\pi t$

ii) Since point 0 has been taken as the reference the current equation is

$\mathrm{I} = 120\ sin\ 120\pi t$

$96 = 120 \sin 120\pi t$

When

$t = {}^{1}/_{360}\ second$

$I = 120 \sin 120\pi \times \frac{1}{360}$

$I = 120 \sin \frac{\pi}{3}$

$\therefore I = 103.92\ A$

iii) To reach the current 96A for the first time

$96 = 120 \sin 120\pi t$

$\sin 120\pi t = \frac{96}{120}$

$\sin 120\pi t = 0.8$

$120\pi t = sin^{-1}\, 0.8$

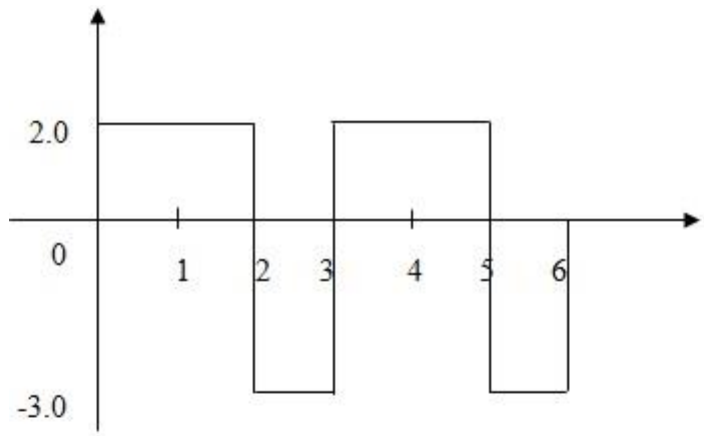
$120\pi t = 0.9273$

$t = \frac{0.9273}{120\pi}$

$t = 2.459 \times 10^{-3}\ s$

$\therefore t = 2.459 ms$

7. Find the r.m.s value of the current after current I shown in figure below



Solution

Consider the current variation over one complete cycle.

Interval	I/A	I^2/A^2
0-1	+2.0	4.0
1-2	+2.0	4.0
2-3	-3.0	9.0

Average value of I^2

$$I^2 = \frac{4+4+9}{3}$$

$$I^2 = 5.67A$$

R.M.S value of current

$$\sqrt{I^2} = \sqrt{5.67}$$

$$= 2.4A$$

$$\therefore R.M.S\ value\ of\ current = 2.4\ A$$

A.C CIRCUIT

An A.C circuit is the closed path followed by alternating current.
When a sinusoidal alternating voltage is applied in a circuit, the resulting alternating current is also sinusoidal and has the same frequency as that of applied voltage.

However there is generally a phase difference between the applied voltage and the resulting current.

As we shall see, this phase difference is introduced due to the presence of inductance (L) and capacitance (C) in circuit.

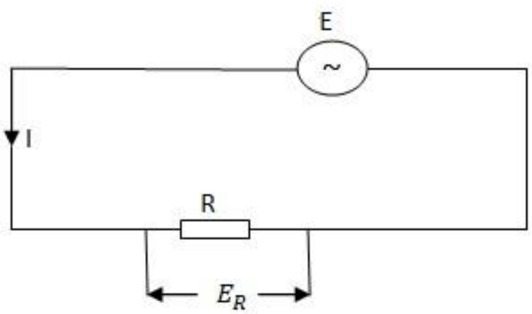
While discussing A.C circuits our main points of interest are;

(i) Phase difference between the applied voltage and circuit current
(ii) Phasor diagram. It is the diagram representation of the phase difference between the applied voltage and the result circuit current.
(iii) Wave diagram
(iv) Power consumed.

A.C CIRCUIT CONTAINING RESISTANCE ONLY

When an alternating voltage is applied across a pure resistance, then from electrons $i.e$ current flow in one direction for the first half cycle of the supply and then flow in the opposite direction during the next half cycle, thus constitute alternate current in the circuit.

Consider a pure resistor of resistance R connected across an alternating source of $e.m.f$



Suppose the instantaneous value of the alternating $e.m.f$ is given by

$$E = E_0 \sin \omega t \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots (i)$$

If I is the circuit current at that instant, then by ohm's law

$$I = \frac{E}{R}$$

$$I = \frac{E_0 \sin \omega t}{R}$$

$$I = \frac{E_0}{R} \sin \omega t \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots (ii)$$

$$I = I_0 \sin \omega t \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots (iii)$$

The value of I will be maximum I_o when

$$\sin \omega t = 1$$

Therefore equation (ii) becomes

From $I = \frac{E_0}{R} \sin \omega t$

Since sin ωt = 1 then I = I_o

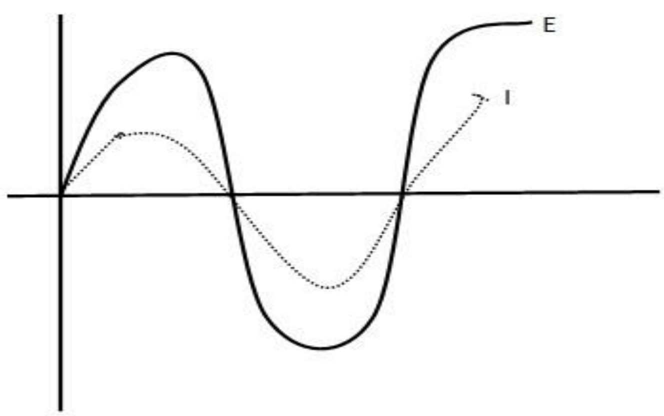
$$I_0 = \frac{E_0}{R}$$

1) Phase Angle

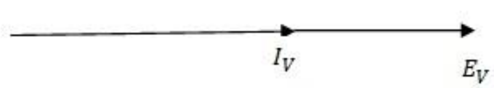
It is clear from equation (i) and (iii) the applied $e.m.f$ and circuit current are in phase with each other $i.e$ they pass through their zero values at the same instant and attain their peak value both positive and negative peaks at the same instant.

This is indicated in the wave diagram shown in figure below.

The $phasor$ diagram shown in figure below also reveals that current is in phase with the applied voltage.



a) Wave diagram



(b) Phasor diagram

Hence in an a.c circuit, current through R is in phase with voltage across R.

This means that current in R varies in step with voltage across R. If voltage across R is maximum current in R is also maximum, if voltage across R is zero, current in R is also zero and so on.

2) Power Absorbed

In a.c circuit, voltage and current vary from instant to instant. Therefore power at any instant is equal to the product of voltage and current at that instant.

Instantaneous power P

$$P = EI$$

$$P = (E_0 \sin \omega t).(I_0 \sin \omega t)$$

$$P = E_0 I_0 [sin^2 \omega t]$$

$$P = E_0 I_0 \left[\frac{1 - \cos 2\omega t}{2}\right]$$

$$P = \frac{E_0 I_0}{2} - \frac{E_0 I_0}{2} \cos 2\omega t \ldots \ldots \ldots \ldots \ldots \ldots \ldots \ldots \ldots (iv)$$

Since power varies from instant to instant the average power over a complete cycle is to be considered.

This is found by integrating equation...........(iv) with respect to time for 1 cycle and dividing by the time of 1 cycle. The time per one cycle is T.

Average power P

$$P = \frac{1}{T}\int_0^T \frac{E_0 I_0}{2}\,dt - \frac{1}{T}\int_0^T \frac{E_0 I_0}{2}\cos 2\omega t\,dt$$

Now

Then

$$\int_0^T \cos 2\omega t\,dt = 0$$

$$P = \frac{1}{T}\int_0^T \frac{E_o I_o}{2}\,dt$$

$$P = \frac{E_o I_o}{2T}\int_0^T dt$$

$$P = \frac{E_o I_o}{2T}[t]_0^T$$

$$P = \frac{E_o I_o}{2}$$

$$P = \frac{E_o}{\sqrt{2}} \times \frac{I_o}{\sqrt{2}}$$

$$P = E_{rms} I_{rms}$$

Therefore, average power absorbed by a resistor in an a.c circuit is equal to the product of virtual voltage (E_{rms}) across it and virtual current (I_{rms}) through it.

Obviously, this power is supplied by the source of alternating $e.m.f$.

Since

$$I_{rms} = \frac{E_{rms}}{R}$$

$$\text{P}= \frac{E_{rms}^2}{R}$$

Or

$$\text{P}= I_{rms}^2\text{R}$$

WORKED EXAMPLES

1. An a.c circuit consists of a pure resistance of 10Ω and is connected across an a.c supply of 230V, 50 Hz.

Calculate

i) Circuit current

ii) Power dissipated and

iii) Equations for voltage and current

Solution

$E_{rms} = 230V$ $R = 10\Omega$ $f = 50HZ$

i) Circuit current

$$I_{rms} = \frac{E_{rms}}{R} = \frac{230}{10}$$

$$I_{rms} = 23A$$

ii) Power dissipated P

$$P = E_{rms} I_{rms}$$

$= 230 \times 23$

$\therefore$ P = 5290 W

iii)Equations for voltage and current

$$E_0 = \sqrt{2}\, E_{rms}$$

$= \sqrt{2} \times 230$

E_0 = 325. 27V

$I_0 = \sqrt{2} x I_{rms}$

= 2 x 23

∴ I_0 = 32.52A

ω = 2πf

= 2$\pi \times 50$

ω = 314 S^{-1}

∴ The equations of voltage and current

E = 325.27 sin 314t and I = 3.52 sin 314t

2. In a pure resistive circuit, the instantaneous voltage and current are given

E = 250 sin 314t

I = 10 sin 314t

Determine

i) Peak power

ii) Average power

Solution

In a pure resistive circuit

i) Peak power = $E_0 I_0$

= 250 x 10

∴ Peak power = 2500W

ii) Average power P

$$P = \frac{E_0 \times I_0}{2}$$

$$P = \frac{2500}{2}$$

$\therefore$ P = 1250 W

3. Calculate the resistance and peak current in a 1000 W hair dryer connected to 120V, 60Hz supply. What happens if it is connected to 240V line?

Solution

$E_{rms} = 120\text{V}$

$\text{P} = 1000\text{W}$

$I_{rms} = ?$

$\text{P} = E_{rms}\, I_{rms}$

$$I_{rms} = \frac{P}{E_{rms}}$$

$I_{rms} = 8.33\text{A}$

Peak current I_o

$$\text{I}_o = \sqrt{2}\; I_{rms}$$

$$= 2 \text{ x } 8.33$$

$$\therefore\ I_O = 11.8A$$

Resistance of hair dryer R

$$R = \frac{E_{rms}}{I_{rms}}$$

$$R = \frac{120}{8.33}$$

$$\therefore\ \text{R} = 14.4\ \Omega$$

When connected to 240V line, the average power delivered would be

$$P = \frac{E_{rms}^2}{R}$$

$$P = \frac{240^2}{14.4}$$

$$\therefore \ P = 4000W$$

This would undoubtedly melt the heating element or the coils of the motor.

4. A voltage E = 60sin 314t is applied across a 20 Ω resistor. What will;

i) An a.c ammeter

ii) Ordinary moving coil ammeter in series with resistor read?

Solution

i) E = 60 sin 314t

An a.c ammeter will read the r.m.s value.

$$E_V = \frac{E_0}{\sqrt{2}} = \frac{60}{\sqrt{2}} = 42.4 \text{ V}$$

$$I_V = \frac{E_v}{R} = \frac{42.4}{20} = 2.12A$$

Therefore a.c a meter will read 2.12A

ii) An ordinary moving coil ammeter will read average value of alternating current. Since the average value of a.c over one cycle is zero, this meter will record zero reading.

5. What is the peak value of an alternating current which produces three times the heat per second as a direct current of 2A in a resistor R?

Solution

Heat per second by 2A

$\frac{H}{t} = I^2$R

H = 2^2R

$\frac{H}{t} =$ 4R

Three times heat per second

3 $\frac{H}{t} = 3 \times 4R$

3 $\frac{H}{t}$ = 12R

If I_v is the *r .m .s* value of the a.c heat per second in R

$\frac{HR}{t} = I_v{}^2$ R

12R = $I^2{}_V$R

$I_v{}^2$ = 12

$\sqrt{I^2{}_V} = \sqrt{12}$

Peak value is given by, $I_0 = \sqrt{2}\, I_v$

$I_0 = \sqrt{2} \times \sqrt{12}$

∴ Peak value = 24 = 4.9A

6. An a.c voltage of 4V peak (maximum) is connected to a 100Ω resistor R

a) What is the phase of the current and voltage?

b) Calculate the current in R in mA

c) What is the power in R in mW

Solution

a) The current and voltage are in phase

b) Current is R

$E_0 = \sqrt{2} \times E_V$

$E_0 = 4\text{V}$

$E_V = ?$

$E_V = \frac{E_0}{2} = \frac{4}{\sqrt{2}}$

$E_V = 2.83\text{V}$

From,

$I_V = \frac{E_V}{R}$

$I_V = \frac{2.83}{100}$

$= 0.0283A$

Or

$I_V = 28mA$

c) Power in R

P = $I_V{}^2$R

P = 0.028^2 x 100

P = 0.078W

P = 78mW

A.C CIRCUIT CONTAINING INDUCTANCE

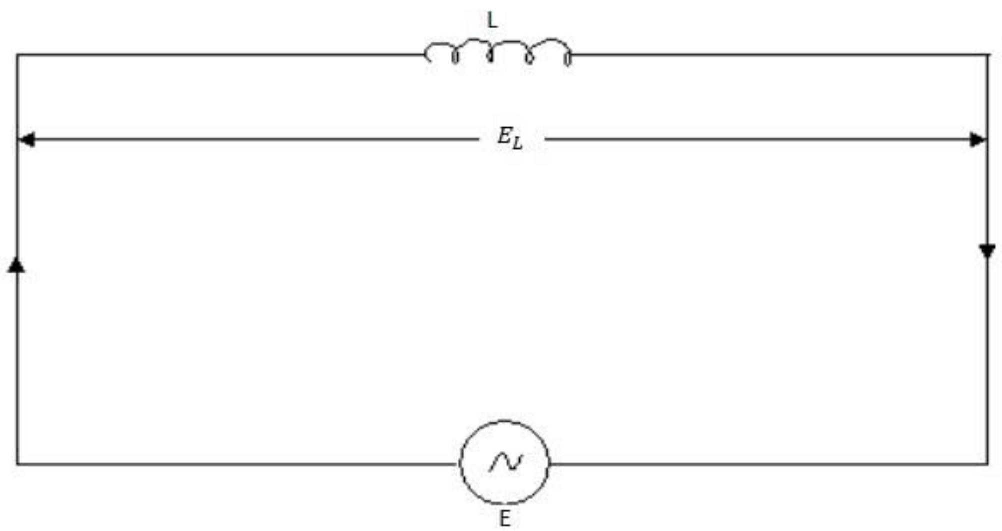
When alternating voltage is applied across a pure inductive coil a back $e.m.f$ (E) is induced in the coil due to its self inductance.

$$E = -L\frac{dI}{dt}$$

The negative sign indicates that induced *e. m .f* opposes the change in current.

In order to maintain the flow of current the applied voltage must be equal and opposite to induced voltage at every instant.

Consider a pure inductor of inductance L connected across an alternating source of $e.m.f$



$E = E_o \sin \omega t$

Suppose the instantaneous value of the alternating *e. m. f* is given by

$E = E_0 \sin wt$ .... (i)

If I is the current in the circuit and $\frac{dI}{dt}$ is the rate of change of current at that instant, then *e. m. f* induced in L is given by

$$E = -L\frac{dI}{dt}$$

As applied voltage is equal and opposite to induce *e .m .f* at every instant

$$E = -(-L\frac{dI}{dt})$$

$$E = L\frac{dI}{dt} \text{ .... (ii)}$$

From, *Equation (i)* and *Equation (ii)*

$$E_0\ Sin\ wt\ =\ L\frac{dI}{dt}$$

$$dI\ =\ \frac{E_0}{L}\ Sin\ \omega tdt$$

Integrating both sides we get

$$\int_0^I dl = \int_0^t \frac{E_0}{L}\sin\omega tdt$$

$$I = \frac{E_0}{L}\int_0^t \sin\omega tdt$$

$$I = \frac{E_0}{L}\left[-\frac{\cos\omega t}{\omega}\right]_0^t$$

$$I = \frac{E_0}{\omega L}[-\cos\omega t]$$

$$I = \frac{E_0}{\omega L}\left[\sin\left(\omega t - \frac{\pi}{2}\right)\right]$$

The value of I will be maximum I_o when, Sin$(wt - \pi/2)$ = 1

Then,

$$I_0 = \frac{E_0}{\omega L}$$

Substituting the value of $\frac{E_0}{\omega L} = I_0$ in equation (iii)

$$I = \frac{E_0}{\omega L}\sin(wt - \pi/2)$$

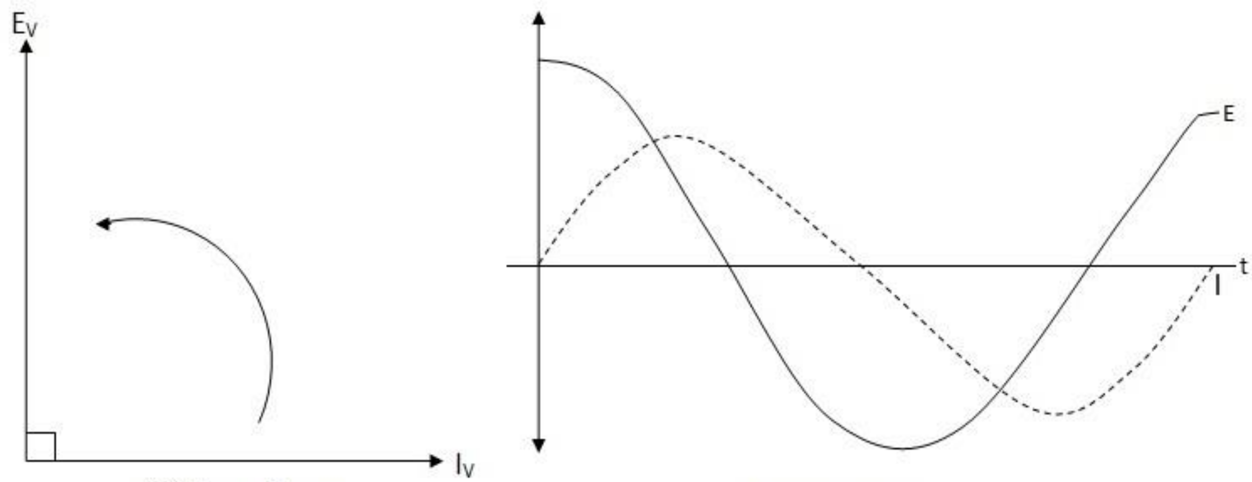
$$I = I_o \sin(wt - \pi/2) \text{ .... (iv)}$$

I) Phase Angle

It is clear from equation (i) and (iv) that circuit current lags behind the applied voltage by (π/2) radians or 90^0.

This fact is also indicated in the wave diagram.

The phase diagram in figure below also reveals the fact that I_V lags behind E_V by 90º



(a) Phasor diagram (b) Wave diagram

Hence in an a.c circuit current through L lags behind the voltage across L by 90º

This means that when voltage across L is zero, current through L is maximum and vice versa.

From

$$E = L\frac{dI}{dt}$$

Now, $\frac{dI}{dt}$ is maximum when circuit current is zero and $\frac{dI}{dt}$ is zero when circuit current is maximum.

INDUCTIVE REACTANCE

Inductive reactance is the opposition in which an inductor offers to current flow. It is denoted by X_L

Inductance not only causes the current to lag behind the voltage but it also limits the magnitude of current in the circuit.

We have seen above that,

$I_o = \frac{E_0}{\omega L}$

$\omega L = \frac{E_0}{I_0}$

Clearly the opposition of inductance to current flow is ωL. This quantity ωL is called inductive reactance XL of the inductor.

Inductive reactance X_L

$$X_L = \omega L$$

Or

$$X_L = 2\pi f L$$

Or

$$X_L = \frac{E_0}{I_0}$$

a) From

$X_L = \frac{E_0}{I_0}$

But $E_0 = \sqrt{2}\, E_V$ and $I_o = \sqrt{2}\, I_V$

Then,

$$X_L = \frac{E_V}{I_V}$$

$X_L = \frac{\sqrt{2}}{\sqrt{2}} \times \frac{E_V}{I_V}$

b) For d.c

f = 0

so that,

$X_L = 2\pi f L$

$X_L = 2\pi \times 0 \times L$

$X_L = 0$

Therefore a pure inductance offers zero opposition to d.c

$X_L \propto f$

c) $X_L = 2\pi f L$

Therefore,the greater f,the greater is X_L and vice versa.

d) We can show that the units of X_L are that of ohm

$X_L = \omega L = \frac{1}{sec}$ x Henry = $\frac{1}{sec} \times \frac{volt}{amp/sec}$ and

$X_L = \frac{volt}{ampere} = ohm$

III) Average Power consumed

E = $E_0 \sin \omega t$

I = $E_0 \sin (\omega t - \frac{\pi}{2})$ I = $-I_0 \cos \omega t$

Instantaneous power P

P = EI

P = ($E_0 \sin \omega t$) ($-I_0 \cos \omega t$)

P = $-E_0 I_0 \sin \omega t \cos \omega t$

P = $-\frac{E_0 I_0}{2} \sin 2\omega t$

Average power P is equal to average of power over one cycle.

$$P = \frac{I}{T}\int_0^T -\frac{E_0 I_0}{2} \text{ Sin } 2\omega t dt$$

$$P = 0$$

Hence average power absorbed by pure inductor is zero

During one quarter cycle of alternating source of *e .m .f.* energy is stored in the magnetic field of the inductor this energy is supplied by the source.

During the next quarter cycle the stored energy is returned to the source. For this reason average power absorbed by a pure inductor over a complete cycle is zero.

NUMERICAL EXAMPLES

1. A pure inductive coil allows a current of 10A to flow from a voltage of 230V and frequency 60Hz supply.

Find

i) Inductive reactance

ii) Inductance of the coil

iii) Power consumed

Write down equations for voltage and current:

Solution

E_V = 230V

I_V = 10A

f =50Hz

i) Inductive reactance X_L

$$X_L = \frac{E_V}{I_V} = \frac{230}{10} = 23\Omega$$

ii) Inductance of the coil L

From

$X_L = 2\pi f L$

$L = \frac{X_L}{2\pi f} = \frac{23}{2\pi} \times 50$

L = 0.073 H

iii) Power absorbed = 0

Also

E_0= 230 x $\sqrt{2}$ I_0= 10 x $\sqrt{2}$

E_0 = 325.27V I_0= 14.14A

ω = 2$\pi \times 50$

ω = 314

Since in pure Inductive circuit current lags behind the applied voltage by $\frac{\pi}{2}$ radians. The equation for voltage and current are,

E = 325 .27 sin 314t, I = 14.14 sin (314t)

2. Calculate the frequency at which the inductive reactance of 0.7H inductor is 220Ω

Solution

X_L = 220 Ω

L = 0.7H

f =?

$X_L = 2\pi f L$

f = $\frac{X_L}{2\pi \times L}$

∴ f = 50 Hz

3. A coil has self inductance of 1.4H. The current through the coil varies sinusoidally with amplitude of 2A and frequency 50 Hz

Calculate

i) Potential difference across the coil

ii) *r .m .s* value of P.d across the coil.

Solution

$I = I_0 \sin \omega t$

(i) P.d across the coil

E = L$\frac{dI}{dt}$

E = L$\frac{d(I_0 \sin \omega t)}{dt}$

E = L$\frac{d(I_0 \sin 2\pi f t)}{dt}$

E = L $I_0 \omega$ cos ωt,

E = L $I_0 2\pi f$ cos2$\pi f t$

E = 2$\pi \times 1.4 \times 2 \times \pi \times 50 \times Cos\ 2\pi \times 50t$

E = 880 cos 100 πt

ii) r.m.s value of potential different across the coil

$E_V = \frac{E_0}{\sqrt{2}}$

$E_V = \frac{880}{\sqrt{2}}$

E_V = 622.2V

4. How much inductance should be connected to 200V, 50 Hz a.c supply so that a maximum current of 0.9A flows through it?

Solution

E_v= 200V

I_o = 0.9A

f = 50 Hz

Peak value of voltage E_0

$E_{0}= \sqrt{2}\,E_v$

$E_{0}= \sqrt{2} \times 200$

Inductive reactance L

$X_L = \frac{E_0}{I_0}$

$X_L = \frac{\sqrt{2} \times 200}{0.9}$

X_L = 314.27 Ω

Inductance L,

$L = \frac{X_L}{2\pi f}$

$L = \frac{314.27}{2\pi \times 50}$

L = 1 H

5. An Inductor of 2H and negligible resistance is connected to 12V, 50Hz supply. Find the circuit current, what current flows when the inductance is changed to 6H?

Solution

* For the First case X_L

$X_L = 2\pi f L$

$X_L = 2\pi \times 50 \times 2$

$X_L = 628\ \Omega$

Circuit current

$I_V = \frac{E_V}{X_L}$

$= \frac{12}{628}$

$I_V = 0.019A$

* For the second case X_L'

$X_L' = 2\pi f L$

$X_L' = 2\pi \times 50 \times 6$

$X_L' = 1884\ \Omega$

Circuit current

$Iv = \frac{E_v}{X_L}$

$I_V = \frac{12}{1884}$

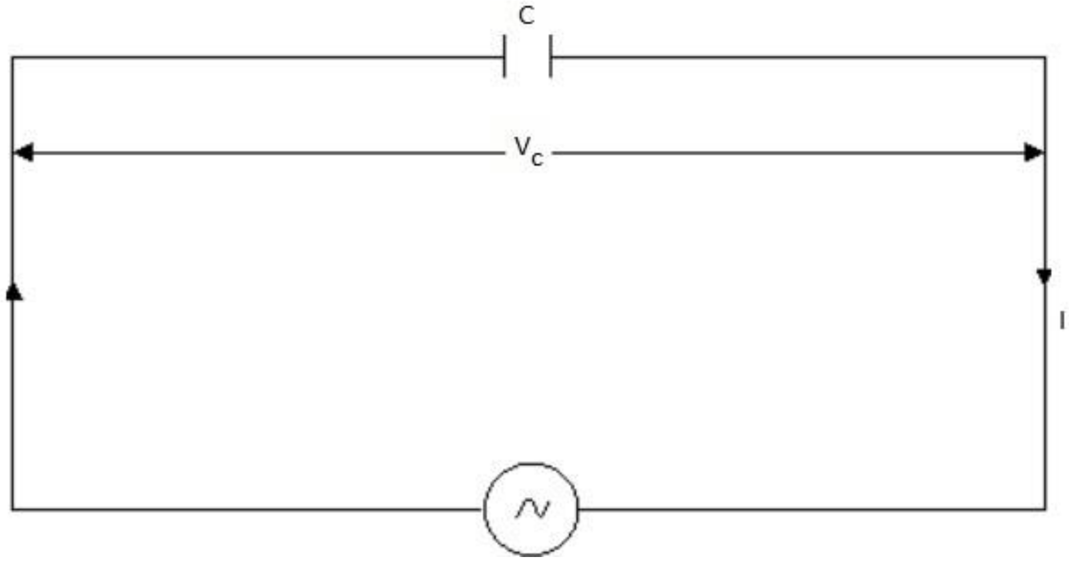
$I_V = 0.0063A$

A.C CIRCUIT CONTAINING CAPACITANCE ONLY

When an alternating voltage is applied to a capacitor, the capacitor is charged first in one direction and then in the opposite direction.

The result is that electrons move to and fro round the circuit connecting the plates, thus constituting alternating current.

Consider a capacitor of capacitance C connected across an alternating source of *e .m. f* .



Suppose the instantaneous value of the alternating *e.m.f* E is given by

E = E_0sin ωt …….. (i)

If I is the current in the circuit and Q is the charge on the capacitor at this instant, then the Potential difference across the capacitor V_C

$$V_C = \frac{Q}{C}$$

At every instant the applied e.m.f E must be equal to the potential difference across the capacitor.

E = V_C

$$E_0 Sin\ \omega t\ = \frac{Q}{c}$$

Q = CE_0 sin ωt

From,

I = $\frac{dQ}{dt}$

I = $\frac{d(CE_0 Sin\ \omega t)}{dt}$

I = CωE_0cos ωt

I = Cω E_0sin (ωt +$\pi/2$)

$$I = \frac{E_0}{1/wc}\sin \omega t + \pi/2$$

The value of I_0 will be maximum Io when sin (wt+$\pi/2$) = 1

$$I_0 = \frac{E_0}{1/\omega C}$$

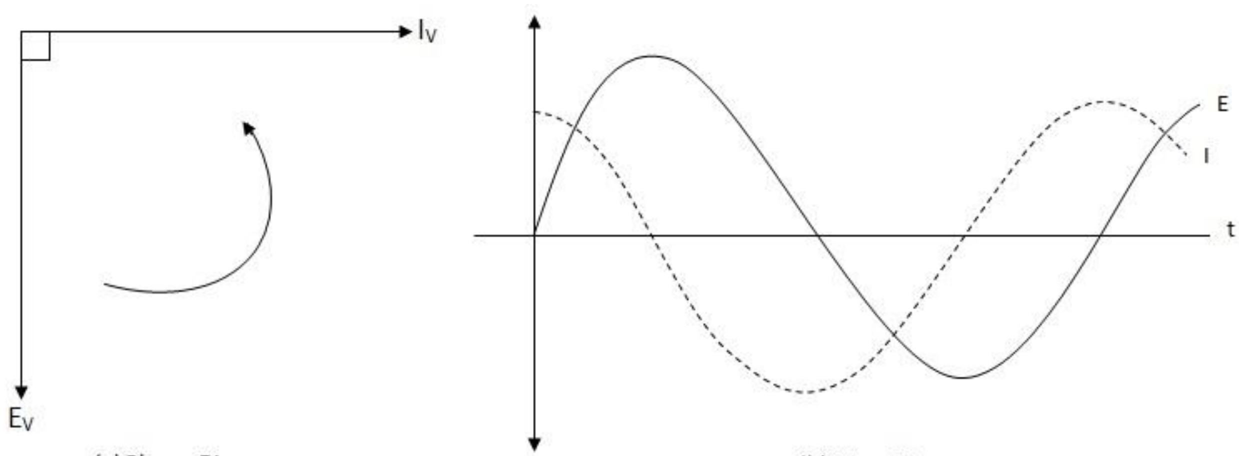
Substituting the value of I_o in equation (ii)

$$I = I_0 \sin(w + \pi/2)$$

1) Phase Angle

It is clear from equation (I) and (ii) that circuit current leads the applied voltage by π/2 radians or 90Ëš.

This fact is also indicated in the wave diagram.



(a) Phasor Diagram (b) Wave Diagram

It also reveals that I_v leads E_v by 90°, hence in a.c circuit current in capacity leads the voltage.
This means that when voltage across capacitor is zero, current in capacitor is maximum and vice versa.
When P.d across capacitor is maximum E_0, the capacitor is fully charged, i.e circuit current is zero.

Since the rate at which a sinusoidally varying p.d falls is greater as it reaches zero, the current has its maximum value when P.d across capacitor is zero. Hence, current and voltage are out of phase by 90°.

CAPACITIVE RESISTANCE

Capacitive resistance is the opposition which a capacitor offers to current flow. It is denoted by X_C.

Capacitance not only causes the voltage to lag behind the current but it also limits the magnitude of current in the circuit.

We have seen above that

$$I_0 = \frac{E_0}{1/\omega C}$$

$$\frac{I_0}{E_0} = \omega C$$

Then

$$\frac{E_0}{I_0} = \frac{1}{\omega c}$$

Clearly, the opposition offered by capacitance to current flow is 1/ωC

The quantity 1/ωC is called capacity reactance X_C of the capacitor

$$XC = \frac{E_0}{I_0}$$

Or

$$XC = \frac{1}{\omega c}$$

X_C will be in Ω if f is in Hz and C in Farad

a) From

$$\frac{E_0}{I_0} = \frac{E_V}{I_V}$$

Then

$\frac{E_V}{I_V} = \frac{1}{\omega C}$

$$XC = \frac{E_V}{I_V}$$

b) For d.c

f = 0

$X_C = \frac{1}{2\pi f C} = \frac{1}{2\pi \times 0 \times C}$

$X_C = \frac{1}{0}$

$X_C = \infty$

Therefore a pure capacitance offers infinite opposition to d.c. In other words, a capacitor blocks d.c.

c. From

$X_C = \frac{1}{2\pi f C}$

$$X_C \propto \frac{1}{f}$$

Therefore the greater the f the smaller is X_C and vice versa

d) From

$$X_C = \frac{1}{2\pi f C}$$

$$X_C = \frac{sec}{farad} = \frac{sec}{coulomb/volt}$$

$$X_C = \frac{volt \times sec}{ampere \times sec} = ohm$$

Therefore the units X_C are Ohm

(iii) **AVERAGE POWER ABSORBED**

From

$E = E_0 \sin \omega t$

$I = I_0 \sin (\omega t + \pi/2)$

$I = I_0 \cos \omega t$

Instantaneous power P

$P = EI$

$P = E_0 \sin \omega t . I_0 \cos \omega t$

$P = E_0 I_0 (\sin\omega t \cos\omega t)$

$P = \frac{E_0 I_0}{2} \sin 2\omega t$

∴ Average power P = Average of P Over one cycle

$P = \frac{1}{T}\int_0^T \frac{E_0 I_0}{2} \sin 2\omega t \, dt$

$P = 0$

Hence average power absorbed by pure capacitance is zero.
During one quarter cycle of the alternating source of $e.m.f$ energy is stored in the electric field of the capacitor.

This energy is supplied by the source during the next quarter cycle, the stored energy is returned to the source

WORKED EXAMPLES

1. A 318μF capacitor is connected 230V, 50Hz supply. Determine

i) The capacitive reactance

ii) r.m.s value of circuit current

iii) Equations for voltage and current

Solution

C = 318μF = 318 x 10^{-6} F

$E_v = 230V$

f = 50 Hz

i) Capacitive reactance X_C

$X_C = \frac{1}{2\pi f C}$

$X_C = \frac{1}{2\pi \times 50 \times 318 \times 10^{-6}}$

$X_C = 10\ \Omega$

ii) r.m.s value of current I_V

$I_V = \frac{E_V}{X_C} = \frac{230}{10}$

$I_V = 23A$

iii) $E_0 = \sqrt{2}.E_V$ $I_0 = \sqrt{2} \times 23$

$E_0 = \sqrt{2} \times 230$ $I_0 = 32.53A$

$E_0 = 325.27\ V$

$\omega = 2\pi F$

$\omega = 2\pi \times 50$

$\omega = 314$

$E_0 = 325.27\sin 314t$ and $I = 32.53 \sin 314t$

2. A coil has an inductance of 1H

a) At what frequency will it have a reactance of 3142Ω?

b) What should be the capacitance of a condenser which has the same reactance at that frequency?

Solution

a) L = 1H

X_L = 3142Ω

f = ?

$X_L = 2\pi f L$

f = $\frac{X_L}{2\pi L}$ = $\frac{3142}{2\pi \times 1}$;

f= 500Hz

b) X_C = 3142 , f=500 Hz

C =?

$X_C = \frac{1}{2\pi f C}$

C = $\frac{1}{2\pi f X_C}$

C = $\frac{1}{2\pi f X_C}$

C = 0.11×10^{-6}

C = $0.11 \mu F$

3. A $50 \mu F$ capacitor is connected to a 230V, 50Hz supply. Determine

i) The maximum charge on the capacitor

ii) The maximum energy stored in the capacitor

Solution

The charge and energy in capacitor will be maximum when p.d across the capacitor is maximum.

i) Maximum charge on the capacitor

Q = CE_0

Q = C$E_V\sqrt{2}$

Q = (50x10^{-6}) x (230 x$\sqrt{2}$)

Q = 16.26 x 10^{-3}C

ii) Maximum energy stored in the capacitor U

U = $\frac{1}{2}$CE^2

U = 1/2 x (50x10^{-6}) x (230 x $\sqrt{2}$)^2

U = 2.65J

4. The Instantaneous current in a pure inductance of 5H is given be

I = 10sin (314t- $\pi/2)$amperes

A capacitor is connected in parallel with the inductor. What should be the capacitance of the capacitor to receive the same amount of energy as inductance at the same terminal voltage?

Solution

The current flowing through pure inductor is

I = 10 Sin (314t$-\pi/2$)

I_0= 10A

ω = 314s^{-1}

Maximum energy stored in the inductor

$$U_L = \frac{1}{2} LI_0{}^2$$

$$U_L = \frac{1}{2} \text{ x } 5 \text{ x } 10^2$$

U_L = 250J ………. (i)

Now

$E_0 = \omega L I_0$

= 314 x 5 x 10

E_0 = 15700v

Max energy stored in the capacitor of capacitance C

$$U_C = \frac{1}{2} CE_o{}^2$$

$U_C = \frac{1}{2} \times 5 \times 15700^2$ ……….(ii)

Equate the equation (I) and (ii)

$= \frac{1}{2} \times C \times 15700^2 = 250$

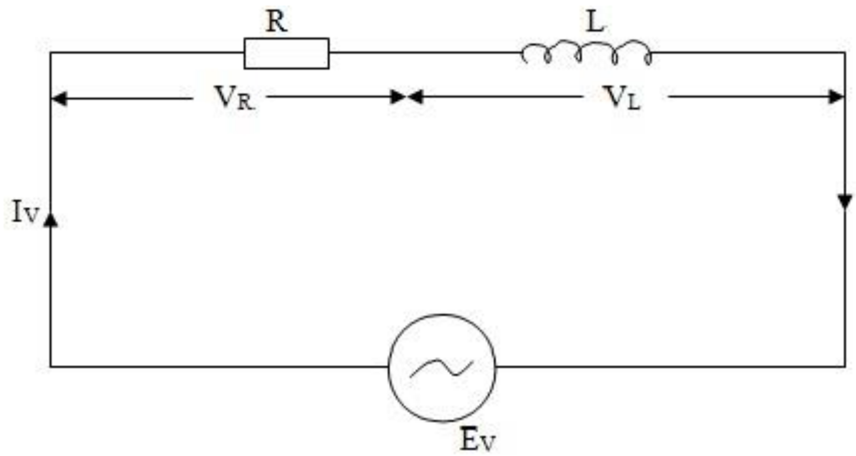
C = $\frac{250 \times 2}{15700^2}$

C = 2.03 x 10^{-6}F

C = 2.03 μF

A.C CIRCUIT CONTAINING R AND L IN SERIES

Consider a resistor of resistance R ohms connected in series with pure inductor of L Henry.



Let

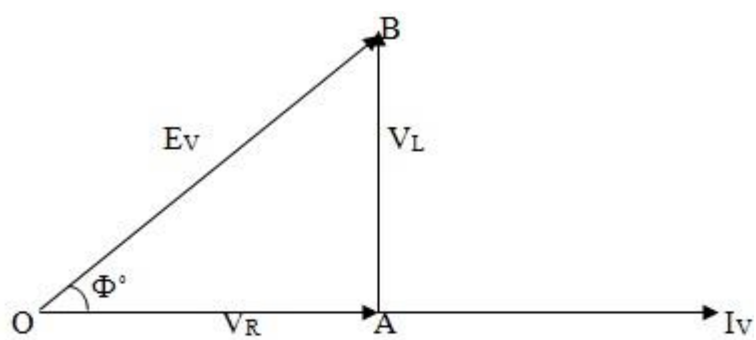
E_V = r.m.s value of applied alternating e.m.f

I_v=r.m.s value of the circuit current

V_R = I_V R when V_R is in phase with I_V

V_L = I_V X_L where V_L leads I_V by 90º

Taking current as the reference phasor, the phasor diagram of the circuit can be drawn as shown in figure.



The voltage drop V_R is in phases with current and is represented in magnitude and direction by the phase OA.

The voltage drop V_L leads the current by 90º and is represented in magnitude and direction by the phase AB.

The applied voltage E_V is the phasor sums of these two voltage drops

$$E_V{}^2 = V_R{}^2 + V_L{}^2$$

$$E_V{}^2 = (I_V R)^2 + (I_V X_L)^2$$

$$E_V = \sqrt{I_V{}^2\left(R^2 + I_V{}^2\right)}$$

$$E_V = I_V\sqrt{R^2 + X_L{}^2}$$

$$I_V = \frac{E_V}{\sqrt{R^2 + X_L^2}}$$

1) **Phase Angle**

It is clear from the phasor diagram that circuit current I_V lags behind the applied voltage E_V by Φº.

Therefore we arrive at a very important conclusion that in an inductive circuit current lags behind the voltage.

NUMERICAL EXAMPLE

1. Three impedance are connected in series across a 200V, 50Hz supply. The first impedance is a 10â„¦ resistor and the second is a coil or 15 â„¦ inductive reactance and 5 â„¦ resistance while the third consists of a 15 â„¦ resistor in series with a 25â„¦ capacitor

Calculate

i) Circuit current

ii) Circuit phase angle

iii) Circuit power factor

iv) Power consumed

Solution

i) Total circuit Resistance

R = 10 +5 + 15

R = 30 â„¦

Total Circuit reactance

$X = X_L - X_C$

X = 15 – 25

X = -10 â„¦ **(capacitive)**

Circuit impedance Z

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

$$Z = \sqrt{30^2 + (-10)^2}$$

Z = 31. 6 â„¦

Circuit current I_V

$I_V = \frac{E\ v}{Z}$

$I_V = \frac{200}{31.6}$

$\mathbf{I_V = 6.\ 33\ A}$

ii) Circuit phase Angle

$$\tan\emptyset = \frac{X_{L-}X_C}{R}$$

$$\tan\emptyset = -\frac{10}{30} = -0.33$$

$$\emptyset = \tan^{-1}(-0.33)$$

$$\emptyset = -18.26°$$

$$\emptyset = 18.26°\ lead$$

iii) Circuit power factor

Power factor = COS θ

= COS 18 .26º

Power factor = 0. 949

iv) Power consumed P

$P = E_V\ I_V \cos\emptyset$

P = (200 x 6.33) x 0. 949

P = 1201. 4 W

Alt

$P = I_v^2 R$

2. A 230V, 50Hz supply is applied to a coil of 0.06 H inductance and 2.5 Ω resistance connected in series with 6.8 μF capacitor

Calculate

i. Circuit impedance

ii. Circuit current

iii. Phase angle between E_V and Iv

iv. Power factor

v. Power consumed

Solution

i) Inductive reactance X_L

$X_L = 2\Pi fL$

$X_L = 2\Pi x\ 50\ x\ 0.60$

$X_L = 18\ .85\ \Omega$

Capacitive reactance X_C

$$X_C = \frac{1}{2\pi f C}$$

$$X_C = \frac{1}{2\pi \times 50 \times 6.8 \times 10^{-6}}$$

$X_C = 468\ \Omega$

Circuit Impedance

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

$$Z = \sqrt{(2.5)^2 + (18.85 - 468)^2}$$

Z = 449. 2 Ω

ii) Circuit current I_V

$$I_v = \frac{E_v}{Z}$$

$$I_v = \frac{230}{449.2}$$

I_V = 0.152 A

iii) Phase angle between E_V and I_V

$$Tan\,\emptyset = \frac{X_L - X_c}{R}$$

$$\tan\ \emptyset^\circ = \frac{18.85 - 468}{2.5}$$

$$\tan\ \emptyset^\circ = -179.66$$

$$\emptyset^\circ = \tan^{-1}(-179.66)$$

$$\emptyset = -89.7^0\ lead$$

iv) Power factor

$$Cos\,\Phi^0 = \frac{R}{Z}$$

$$\cos\Phi^0 = \frac{2.5}{449.2}$$

Cos Φ = 0.0056

v) Power consumed P

$P = E_V I_S \text{ Cos } \Phi$

P = 230.x 0.512 x 0.0056

P = 0.66 W

3. A resistance R, and inductance L = 0.01H and a capacitance C are connected in series. When an alternating voltage E = 400Sin ($3000t- 20^0$) is applied to the series combination, the current flowing is 10√2 Sin (3000t - 650). Find the value of R and C

Solution

The circuit current lags behind the applied voltage by θ

$\theta = 65^0 – 20$

$\theta = 45^0$

This implies that the circuit is inductive i.e.

$X_L > X_C$

The net circuit reactance X

$X = X_L – X_C$

Now

$X_L = \omega L$

$X_L = 3000 \times 0.01$

$X_L = 30\ \Omega$

Also

From $\tan\phi = \frac{X}{R}$ $\phi = 45^{\circ}$

$Tan\ 45 = \frac{X}{R}$

$$1 = \frac{X}{R}$$

Circuit Impedance Z

$$Z = \frac{E_o}{I_o} = \frac{400}{10\sqrt{2}}$$

Z = 28.3 Ω

Now

$Z^2 = R^2 + X^2$

$Z = R^2 + R^2$

$Z^2 = 2R^2$

Z = R √2

$R = \frac{Z}{\sqrt{2}} = \frac{28.3}{\sqrt{2}}$

R = 20Ω

Now

$X = X_L – X_C$

$20 = 30 – X_C$

X_C= 10Ω

From

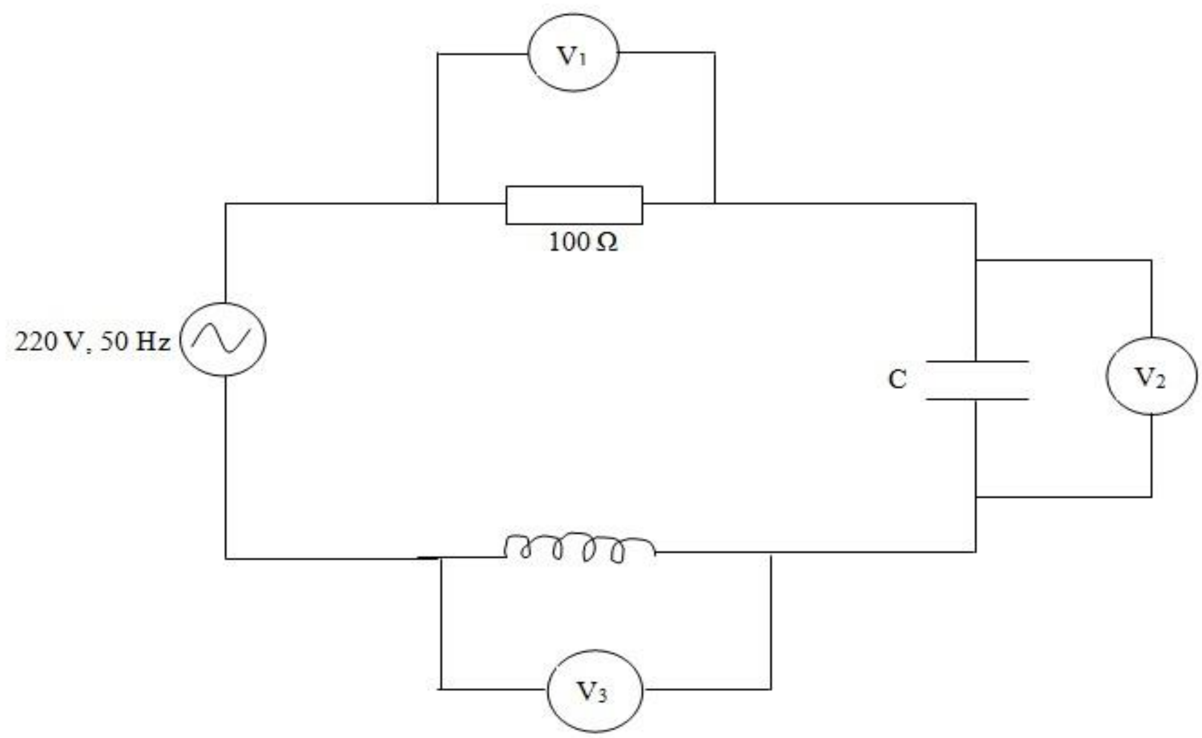
$$X_C = \frac{1}{2\pi f C}$$

$$C = \frac{1}{2\pi f X_c}$$

$$C = \frac{1}{3000 \times 10}$$

C = 33.3 x 10 $^{-6}$ F

4. A series RLC circuit is connected to an a.c (220V, 50 H) as shown in the figure below



If the reading of the three volt meter V_1 V_2 and V_3 are 65V, 415V and 204V respectively.

Calculate

i. The current in the circuit

ii. The value of inductor L

iii. The value of capacitor C

Solution

Here voltmeters are considered ideal i.e. having infinite resistance.

Therefore, it is a series RLC circuit

i) Circuit current I_V

$$I_V = \frac{V_R}{R} = \frac{65}{100}$$

I_V = 0. 65 A

ii) Inductive reactance X_L

$$X_L = \frac{V_L}{I_v} = \frac{204}{0.65}$$

X_L = 318. 85 Ω

Inductance L

$$L = \frac{X_L}{2\pi f} = \frac{313.85}{2\pi \times 50}$$

L = 1H

iii) Capacitive reactance X_C

$$X_C = \frac{V_C}{I_C} = \frac{415}{0.65}$$

X_C = 638. 46Ω

Capacitance C

$$C = \frac{1}{2\pi f X_C}$$

$$C = \frac{1}{2\pi \times 50 \times 638.46}$$

C = 5 x 10^{-6} F

5. A coil of resistance 8Ω and inductance 0.03H is connected to an a.c supply of 240V, 50 Hz.

Calculate

i) The current the power and power factor.

ii) The value of a capacitance which when connected in series with the above coil causes no change in the value of current and power taken from the supply

Solution

i) Reactance of the coil X_L

$$X_L = 2\pi f L$$

$$X_L = 2\pi \times 50 \times 0.03$$

X_L=9.42 Ω

Impedance of the coil Z

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

$$= \sqrt{8^2 + 9.42^2}$$

$Z = 12.46\ \boldsymbol{\Omega}$

Circuit Current I_V

$$I_V = \frac{E_V}{Z}$$

$$I_V = \frac{240}{12.46}$$

$\mathbf{I_V = 19.42\ A}$

Power consumed

$$P = I_V^2 R$$

$P = (19.42)^2 \times 8$

P= 3017 W

Power factor Cos Φ

$$Cos\,\Phi = \frac{R}{Z}$$

$$Cos\,\Phi = \frac{8}{12.46}$$

Cos Φ = 0.65 lag

ii) To maintain the same current and power, the impedance of the circuit should remain unchanged. Thus the value of capacitance in the series circuit should be such so as to cause the current to lead by the same angles as it previously lagged.

This can be achieved if the series capacitor has a capacitive reactance equal to twice the inductive reactance.

$X_C = 2X_L$

$X_C = 2 \times 9.42$

X_C = 18.84 **Φ**

Now

$$X_C = \frac{1}{2\pi f C}$$

$$C = \frac{1}{2\pi f X_c}$$

$$C = \frac{1}{2\pi \times 50 \times 18.84}$$

C = 169 x 10^{-9} F

RESONANCE IN R – L – C SERIES A.C CIRCUIT

The R- L – C series A.C circuit is said to be in electrical resonance when the circuit power factor is unity

$X_L = X_C$

This is called series resonance.

Resonant frequency

This is the frequency at which the reactance of the coil has the same magnitude and so giving the current in a circuit its maximum value.
$X_L = X_C$

where

X_C = capacitive reactance

X_L= inductive reactance

The frequency at which resonance occurs is called the resonant frequency

The resonance in R – L – C series circuit can be achieved by changing the supply frequency because X_L and X_C are frequency dependent

At a certain frequency, called the resonant frequency f_o, X_L becomes equal to X_C and resonance occurs.

At resonance

$X_L = X_C$

$$2\pi f_o L = \frac{1}{2\pi f_o C}$$

$$f_o^2 = \frac{1}{4\pi^2 LC}$$

$$f_o = \frac{1}{2\pi\sqrt{LC}}$$

If L and C are in Henry and farad respectively, then f_o will be in Hz

EFFECTS OF SERIES RESONANCE

When series resonance occurs, the effect on the circuit is the same as though neither inductance nor capacitance is present.

The current under this condition is dependent solely on the resistance of the circuit and voltage across it.

(i)The impedance of the circuit is minimum and equal to the resistance of the circuit.

From $Z = \sqrt{(R^2 + (X_L + X_C)^2}$

When $Z=Z_R$, $X_L = X_C$

$$Z_R = \sqrt{R^2 + 0}$$

$$Z_R = \sqrt{R^2}$$

$Z_R = R$

At series resonance

(ii) The circuit current is maximum as it is limited by the resistance of the circuit alone.

$$I_R = \frac{E_V}{Z_R}$$

(iii) Since at series resonance the current flowing in the circuit is very large, the voltage drops across L and C are also very large.

In fact, these drops are much greater than the applied voltage.

However, voltage drop across L – C combination as a whole will be zero because these drops are equal in magnitude but 180^0 out of phase with each other.

RESONANCE CURVE

Resonance curve is the curve between the circuit current and the supply frequency.

Figure below shows the resonance curve of a typical R-L-C series circuit.

Current reaches the maximum value at the resonant frequency f_o, falling off rapidly on either side at that point.

It is because if the frequency is below f_o, $X_C > X_L$ and the net reactance is no longer zero.

Net reactance X

$X = X_C - X_L$

If the frequency is above fo, then $X_L > X_C$ and the net reactance is again not zero.

Net reactance X

$X = X_L - X_C$

In both cases, the circuit impedance will be more than the impedance Z_R at resonance.

The result is that magnitude of circuit current decreases rapidly as the frequency changes from the resonant frequency.

The effect of resistance in the circuit. The smaller the resistance, the greater is the current at resonance and sharper the resonance curve.

On other hand, the greater the resistance, the lower is the resonant peak.

Q – FACTOR OF SERIES RESONANT CIRCUIT

At series resonance, the p.d across L or C (the two voltage drop being equal and opposite) builds up to a value many times greater than the applied voltage E_V.

The voltage magnification produced by series resonance is termed as Q – Factor of the series resonant circuit.

Q – Factor of a resonant.

R-L circuit

This is the ration of voltage across L or C to the applied voltage.

Or

This is the ratio of power stored to power dissipated in the circuit reactance and resistance respectively.

$$Q - Factor = \frac{Power\ stored}{\text{Power dissipated}}$$

$$Q - Factor = \frac{Voltage\ across\ L\ or\ C}{Applied\ voltage}$$

$$Q - Factor = \frac{I_R^2 X}{I_R^2 R} = \frac{I_R X}{I_R R}$$

$$Q - Factor = \frac{X}{R}$$

Where

X = Capacitive or inductive reactance at resonance

R = Series resistance

The Q – factor of a series resonant circuit can also be expressed in terms of L and C.

From,

$$f_o = \frac{1}{2\pi\sqrt{LC}}$$

$$2\pi f_o = \frac{1}{\sqrt{LC}}$$

$$\omega_o = \frac{1}{\sqrt{LC}}$$

Also

$$Q\ \ Factor = \frac{X_L}{R} \quad or \quad \frac{X_C}{R}$$

$$Q - Factor = \frac{\omega_o L}{R}$$

$$Q - Factor = \frac{L}{R\sqrt{LC}}$$

$$Q - Factor = \frac{1}{R}\sqrt{\frac{L}{C}}$$

The value of Q-factor depends entirely upon the design of the coil i.e. R – L part of the R – L – C circuit because resistance arises in this rather than.
With a well designed coil, the quality factor can be 200 or more.

PHYSICAL MEANING OF Q – FACTOR

The Q- Factor of series a.c circuit indicates how many times the p.d across L or C is greater than the applied voltage at resonance.

For example, consider on R-L- C series circuit connected to 240v a.c source.

If Q – factor of the coil is 20, then voltage across L or C will be,

$$Q - Factor = \frac{Voltage\ across\ L\ or\ C}{\text{Applied voltage}}$$

$$Voltage\ \ across\ L\ or\ C\ = Q - Factor\ \ x\ \ Applied\ Voltage$$

$V_C = V_L = QV_R = 20 \times 240$

$V_C = V_L = 4800$ V

Q – FACTOR AND RESONANCE CURVE

At series resonance the circuit current is maximum ($I_R = E_V/R$) and is limited by circuit resistance only.

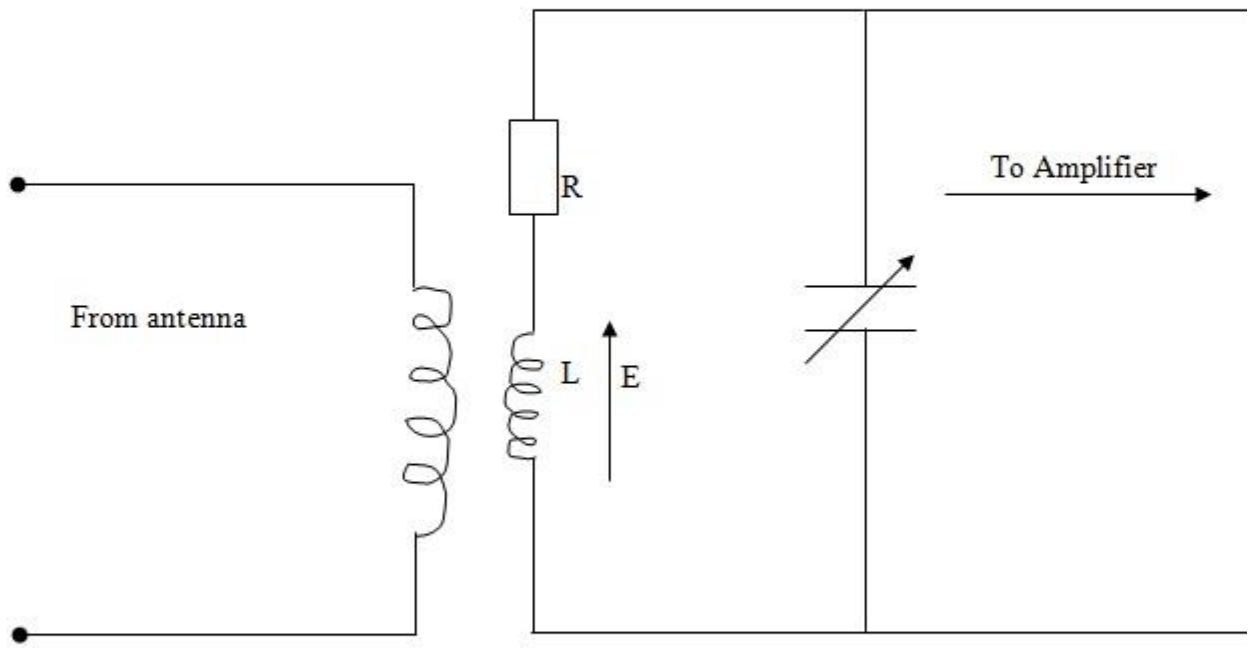
The smaller the circuit resistance, the greater is the circuit current and sharper will be the resonance curve.

Smaller circuit resistance means large value of Q – Factor. Therefore, the greater the Q – Factor of resonant R-L-C circuit, the sharper is the resonance curve.

$$Q - Factor = \frac{X_L}{R}$$

APPLICATION OF SERIES RESONANCE

One important applicant of series resonance is to tune radio and TV receivers.



The input signal comes from the antenna and induces a voltage E in L of the series resonant circuit.

$V_C = QE$

The voltage across the capacitor becomes V_C.

where:

Q = Quality factor of the circuit

As the value of Q is generally large, the original signal received by the antenna increases many times in value and appears across C.

The value of V_C is much more than that could have been obtained by direct transformer ratio.

Thus amplifier receives a greatly increased signal.

Bandwidth of a series resonant circuit.

This is the range of frequencies over which circuit current is equal to or greater than 70.7% of maximum current. (I_R, current at resonance).

The two frequencies are cut off frequency f_1 and upper off frequency f_2.

Bandwidth (BW) = $f_2 - f_1$

It can be shown by using mathematical equation:-

$$Q = \frac{f_o}{BW}$$

The circuit has a general appearance of a parallel circuit but actually it is a series circuit.

It is because no separate volt is applied to L but instead a voltage E is induced in it which is co as a voltage in series with LC.

Q – Factor of a resonant Circuit

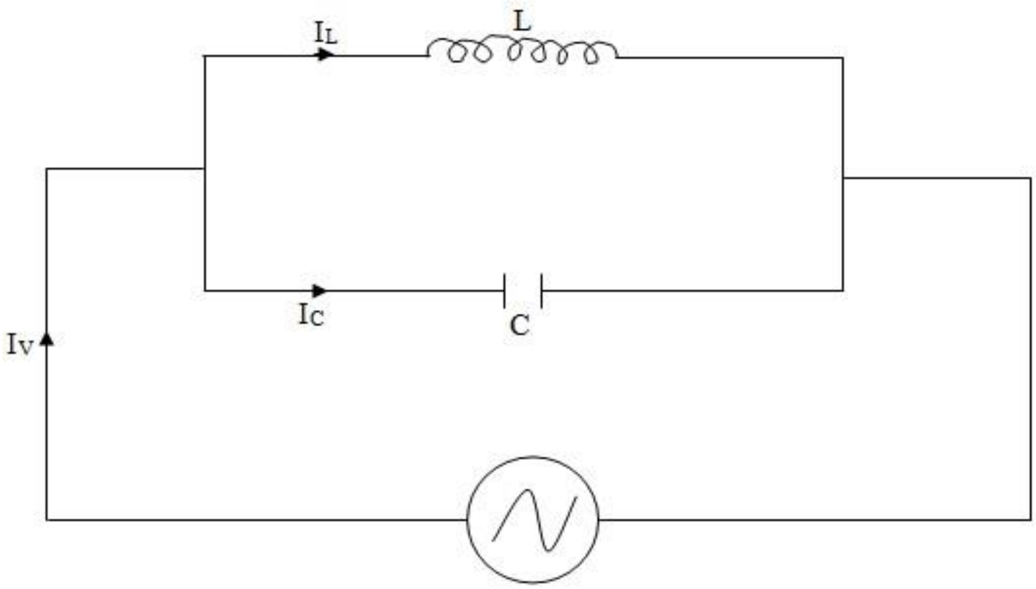
Is the ration of resonant frequency fo to the bandwidth of the circuit

$$Q = \frac{f_o}{BW}$$

RESONANCE IN PARALLEL A.C CIRCUIT

A parallel a.c circuit containing reactive element (L and C) is said to be in electrical resonance when the circuit power factor is unity. This is called parallel resonance

Consider a pure inductor of inductance L connected in parallel with a capacitor of capacitance across an a .c source of voltage E_V (r. m . s)



The circuit will be in resonance when the circuit power factor is unity.
This means that wattles component of the circuit current should be zero.

$I_L - I_C = 0$

The resonance in a parallel a .c circuit can be achieved by changing the supply frequency because X_L and X_C are frequency dependent .

At a certain frequency called resonant frequency f_0, I_C become equal to I_L and resonance occur.

At resonance

$I_C = I_L$

From

$$V_L = V_C$$

$$I_L = I_C$$

$$\frac{E_V}{X_L} = \frac{E_V}{X_C}$$

$$X_L = X_C$$

$$2\pi f_o L = \frac{1}{2\pi f_o C}$$

$$f_o = \frac{1}{2\pi\sqrt{LC}}$$

f_0 will be in Hz if L is in Henry and C is in farad

EFFECTS OF PARALLEL RESONANCE

1) The circuit power factor becomes unity. This implies that the circuit act as a resistor

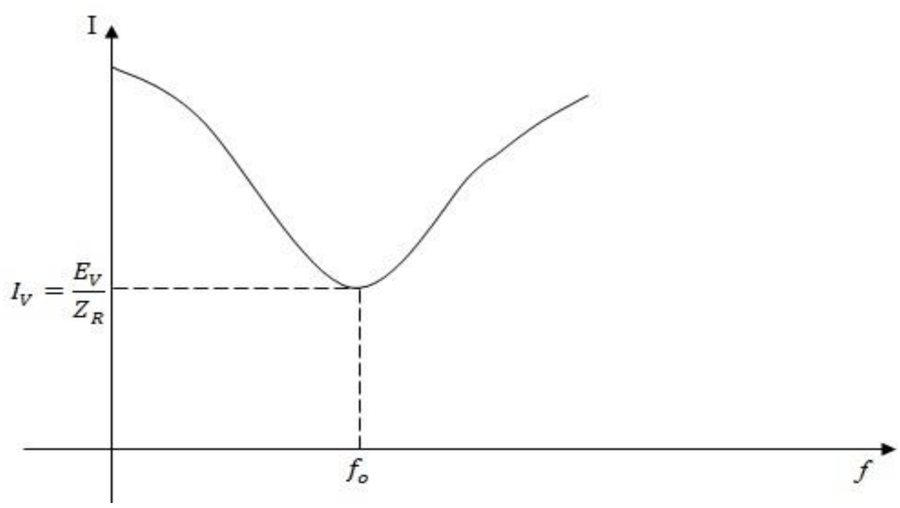
2) The impedance (resistive) of the circuit becomes maximum.

3) The circuit current is minimum the small current I_V flowing in the circuit is only the amount needed to supply the resistance losses.

RESONANCE CURVE

This is the curve between the circuit current and the supply frequency.

Figure below shows the resonance curve of a parallel a.c circuit.



The circuit current I_V is minimum at parallel resonance.

As the frequency changes from resonance, the circuit current increases rapidly.

This action can be explained as follows. For frequencies other than the resonance, the reactive currents (I_L and I_C) in the two branches of the circuit are not equal.
The resultant reactive current must be supplied by the a.c source .
As the difference of the reactive currents in the two branches increase with the amount of deviation from the resonant frequency, the circuits current will also increase.

ADVANTAGE OF A.C OVER D.C

(i)Alternating voltages can be stepped up or stepped down efficiently by a transformer.
This permits the transmission of electric power at high voltages to achieve economy and distribute the power at utilization voltages.

(ii)A.C motors are cheaper and simpler in construction than the d.c motors.

(iii) A. C can be easily converted into d.c by rectifiers.

(iv)Alternating current can be controlled with a choke coil without any appreciable loss of electrical energy.

(v) The switch gear (e.g. switches, circuit breakers) for a.c system is cheaper than the d.c system.

DISADVANTAGE OF A.C OVER D.C

(i) For the same voltage (same value of voltage), a.c is more dangerous than d.c.

(ii) The shock of a.c is attractive whereas that of d.c is repulsive.

(iii) A.C cannot be used for some processes e.g. electroplating, charging of batteries etc.

(iv) A.C is transmitted more over the surface of the conductor than from inside. This is called skin effect.

To avoid skin effect, a.c is transmitted over several fine insulated wires instead of a single thick wire.

WORKED EXAMPLES

1. A coil of resistance 100Ω and inductance 100μH is connected in series with a 100 pF capacitor. The circuit is connected to a 10V variable frequency source.

Calculate:

(i) Resonant Frequency

(ii) Current at resonance

(iii) Voltage across L and C at resonance.

Solution:

Capacitance C = 100pF = 100 x 10^{-12} F

Inductance L = 100μH = 100 x 10^{-6} H

(i) Resonant frequency f_0

$$f_o = \frac{1}{2\pi\sqrt{LC}} = \frac{1}{2\pi \times 10^{-4} \times 10^{-10}}$$

f_0 = 1 . 59 x 10^6 HZ

(ii) Current at resonance I_R

$$I_R = \frac{E_V}{R} = \frac{10}{100}$$

I_R = 0.1 A

(iii) Voltage across L

V_L = Ir X_L

$$V_L = I_r \times 2\pi f_o L$$

$$VL = 0.1 \times 2\pi \times 1.59 \times 10^6 \times 10^{-4}$$

V_L = 100V

(iv) Voltage across C

$$V_C = I_r \times C$$

$$V_C = 0.1 \times \frac{1}{2\pi f_o C}$$

$$V_c = 0.1 \times \frac{1}{2\pi \times 1.59 \times 10^6 \times 10^{-10}}$$

V_C = 100V

At series resonance , the voltage across L or C is much greater than the applied voltage.

2. A circuit, having a resistance of 4Ω and inductance of 0.5 H and a variable capacitance in series, is connected across a 100V, 50 Hz supply. Calculate;

(i) The capacitance to give resonance

(ii) The voltage across inductance and capacitance.

Solution

(i) At series resonance

$X_L = X_C$

$$2\pi f_o L = \frac{1}{2\pi f_o C}$$

$$C = \frac{1}{4\pi^2 f_0^2 L} = \frac{1}{4\pi^2 \times 50^2 \times 0.5}$$

$C = 20.26 \times 10^{-6}$F

(ii) Current at resonance Ir.

$$I_r = \frac{E_v}{R}$$

$$I_r = \frac{100}{4}$$

Ir = 25 A

P.d across L

$V_L = I_r X_L$

$$V_L = 25 \times 2\pi \times f_o L$$

$$V_L = 25 \times 2\pi \times 50^2 \times 0.5$$

$V_L = 3927V$

P.d across C

$V_C = Ir\ X_C$

$$V_C = I_r \times \left(\frac{1}{2\pi f_o C}\right) = 25 \times \frac{10^6}{2\pi \times 50 \times 20.26}$$

V_C = 3927 V

3. A series R – L- C circuit consists of a 100 Ω resistor, an inductor of 0.318 H and a capacitor of unknown value. When this circuit is energized by √2 x 230 sin 314t volts are supply the current is found to be √2 x2.3 sin 314t find.

(i) The value of capacitor in microfarad.

(ii) Voltage across indicator

(iii) Total power consumed.

Solution

From,

$E_O = \sqrt{2}\ E_V$

$E = \sqrt{2} \times 230 \text{ Sin } 314t$

$E = E_v \text{ Sin } \omega t$

$E_v = \sqrt{2} \times 230$

$\sqrt{2} \times 230 = \sqrt{2} \times E_V$

E_V = 230V.

Circuit current I_V

$$I_V = \frac{E_V}{R} = \frac{230}{100}$$

Iv = 2.3 A

Also further phase angle is zero

$\omega = 314\ S^{-1}$

f = 50 Hz

i) At resistance

$X_L = X_C$

$$2\pi f_o L = \frac{1}{2\pi f_o C}$$

$$C = \frac{1}{4\pi^2 f_o^2 X_L} = \frac{1}{(2\pi \times 50)^2 \times 0.318}$$

C = 31.4×10^{-6} F

C = 31.4 x μF

ii) Voltage across inductor V_L

$$\boldsymbol{V_L = I_V X_L}$$

$$V_L = 2.3 \times 2\pi \times 50 \times 0.318$$

V_L = 230 V

iii) Total power consumed

$$P = I_V^2 R$$

P= 2.3^2 x 100

P = 529 W

4. A coil of inductance 8 µH is connected to a capacitor of capacitor of capacitance 0.02 µF. To what wavelength is this circuit is tuned?

Solution

L = 8 x 10 $^{-6}$ H

C = 0.02 x 10 $^{-6}$F

Resonant frequency f_0

$$f_0 = \frac{1}{2\pi\sqrt{LC}} = \frac{1}{2\pi\sqrt{(8 \times 10^{-6} \times 0.02 \times 10^{-6})}}$$

f_0 = 3.98 x 10^5 Hz

If C = 3. 10^8 m /s is the velocity of the e.m wave, then wavelength λ

$$\lambda = \frac{c}{f_o} = \frac{3 \times 10^8}{3.98 \times 10^5}$$

λ= 7.54 X 10^2 m

5. a) A Sinusoidal voltage of peak value 283 V and frequency 50 Hz is applied to A series L.C. R circuit in which R = 3 Ω , L = 25.48 µH and C = 796 µF . Find

i) The impendence of the circuit.

ii) The phase difference between voltage across the source and the current

iii) The power dissipated in the circuit

iv) Power factor.

b) Suppose the frequency of the source in the previous example can be varied.

i) What is the frequency of the source at which resonance occurs?

ii) Calculate the impedance the current and the power dissipate at the resonant condition.

Solution

a) $E_0 = 283$ V

f= 50 Hz

R = 3Ω

L = 25.48 x 10^{-3} H

C = 796 x 10^{-6} F

$$X_L = 2\pi f l = 2\pi \times 50 \times 25.48 \times 10^{-3}$$

X_L = 8Ω

$$X_C = \frac{1}{2\pi f C} = \frac{1}{2\pi \times 50 \times 796 \times 10^{-6}}$$

X_C = 4Ω

i) Circuit impedance Z

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

$$\sqrt{3^2 + (X_L + X_C)^2}$$

Z = √25

Z = 5 Ω

ii) The phase difference between voltage and circuit current

$$\tan \emptyset = X_L - X_C$$

$$= \frac{8-4}{3}$$

$$\tan \emptyset = \frac{4}{3}$$

$$\emptyset = \tan^{-1}\frac{4}{3}$$

$$\emptyset = 53.1°$$

iii) Circuit current I_V

$$I_V = \frac{E_V}{Z}$$

$$I_o = \frac{\frac{E_o}{\sqrt{2}}}{Z}$$

$$I_V = \frac{283}{\sqrt{2} \times 5}$$

I_V = 40 A

iv) Power dissipated in the circuit

P = I_V R

P = 40^2 x 3

P = 4800 W

Power factor = Cos 53.1º

b) i) the frequency at which resonance occur

$$f_o = \frac{1}{2\pi\sqrt{LC}}$$

$$f_o = \frac{1}{2\pi \times \sqrt{25.48 \times 10^{-3} \times 769 \times 10^{-6}}}$$

f_0 = 35.4 HZ

ii) The impedance z at resonance is equal to R

Z = 3Ω

The circuit Current I_V

$$I_v = \frac{E_V}{Z} = \frac{\frac{E_o}{\sqrt{2}}}{Z}$$

$$I_V = \frac{283}{\sqrt{2} \times 3}$$

I_V = 66 .7 A

Power dissipate at resonance is

$$P = I_v^2 R$$

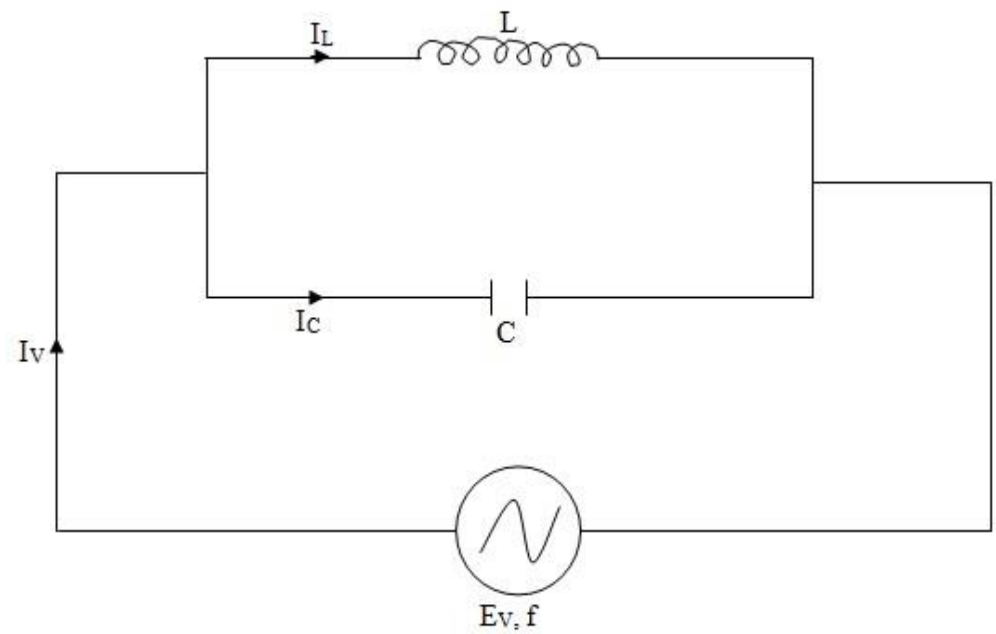
P = (66.7) x 3

P = 13350 W

L - C IN PARALLEL CIRCUIT.

Consider inductor of inductance L Henry connected in parallel with capacitor of capacitance C.

Figure below shows the two component of current.



I_L lags by 90º on E_V but I_C leads by 90º on E

CASE 1

If I_C is greater than I_L at the particular frequency

$$I = I_C - I_L$$

$$I = \frac{E_V}{X_C} - \frac{E_V}{X_L}$$

$$I = E_r\left[\frac{1}{X_c} - \frac{1}{X_L}\right]$$

Since I leads by 90º on E in this case we say that the circuit in net capacitive.

CASE 2

If I_L is greater than I_C

$I = I_L - I_C$

$$I = \frac{E_V}{X_L} - \frac{E_V}{X_C}$$

$$I = E_V\left[\frac{1}{X_L} - \frac{1}{X_C}\right]$$

Since I lags by 90º on E in this case, the circuit is Net inductive.

CASE 3

At the Resonance

$X_C = X_L$

$$\frac{1}{2\pi f_o C} = 2\pi f_o L$$

$$f_o = \frac{1}{2\pi\sqrt{LC}}$$

ELECTRONICS

1. Conductors

Posses free electrons

Metals are all good conductors due to having low resistance to the flow of current.

2. Insulators

They do not have free electrons for conduction. They have high resistance to the flow of current.

All non metals are bad conductors. Eg. dry wood, paper and air.

3. Semiconductors

These are class of materials whose conductivity is between that of good conductors and insulators

Silicon and Germanium are examples of semiconductors elements widely used in electronic industry.

INTRINSIC SEMICONDUCTORS

These are pure semiconductors.

EXTRINSIC SEMICONDUCTORS

These are impure semi conductors material.

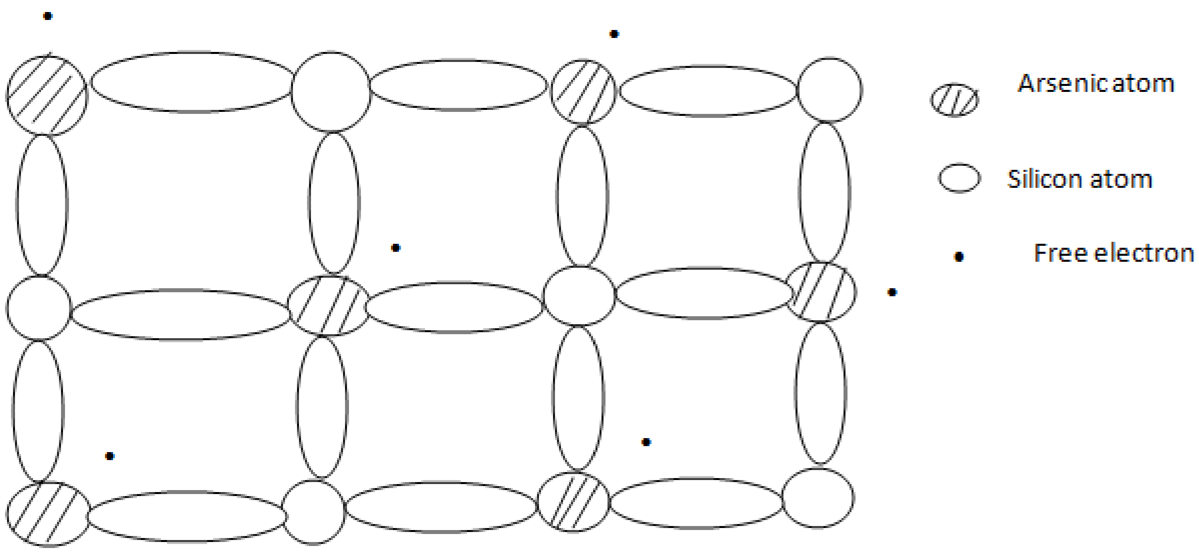
DOPING

Is a process of introducing a tiny amount of impurity into a semiconductors material to form extrinsic semiconductors shells.

N-SEMI CONDUCTORS

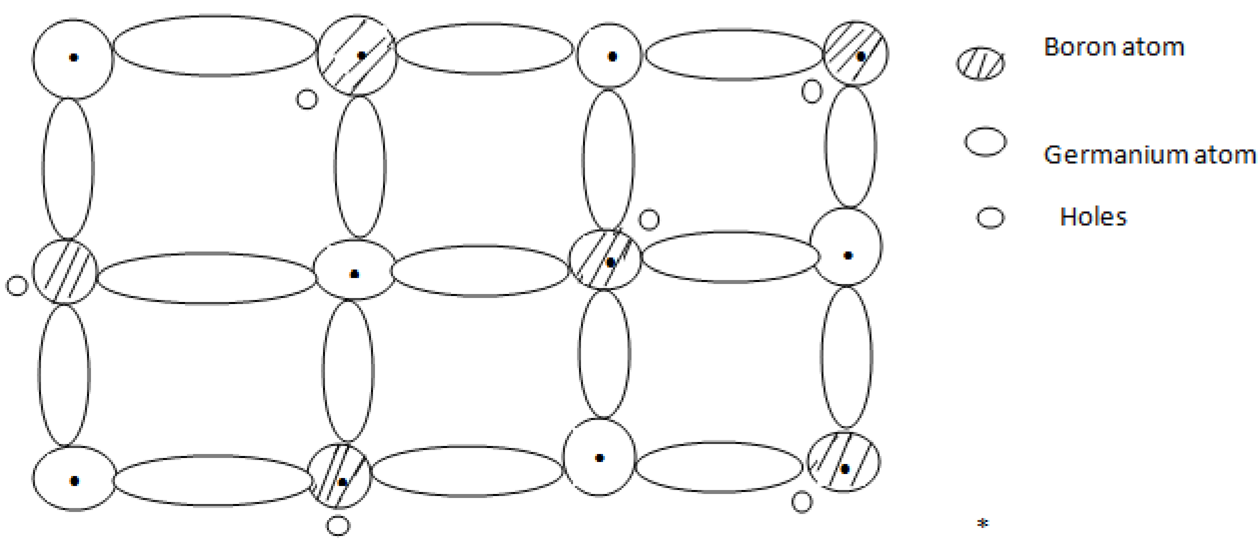
-Silicon and germanium atoms are tetravalent

-They have four electrons in their outermost shell.
-When a doner atom with fine electrons in its outer most shell (ie Arsenic) is added to a silicon crystal, the fifth electrons becomes a free change carriers since there is production of large number of negative charge carriers(electrons) the impure semiconductors is called N-Semiconductors



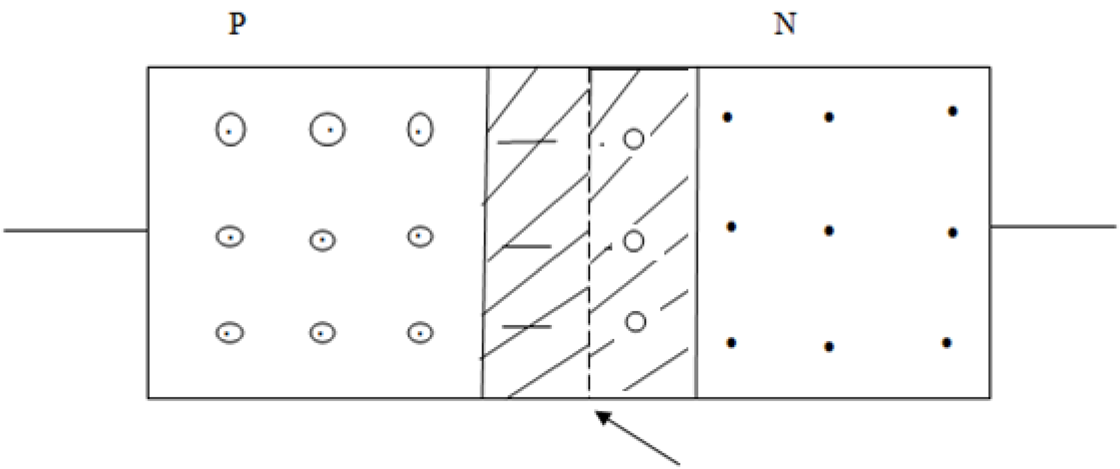
P-SEMICONDUCTORS

A-P- Semiconductor is made by adding a trivalent atom (an acceptor) such as B or on to pure semi conductor such as germanium. Since there is a production of large number of holes (positive charges) the impure semiconductor is called P- Semiconductor.



P-N=JUNCTION/DIODE:

This is formed when P and N semiconductors are melted to form a junction between them



The marrow region at the P-n junction which contains the negative and positive charge is called depletion layer.

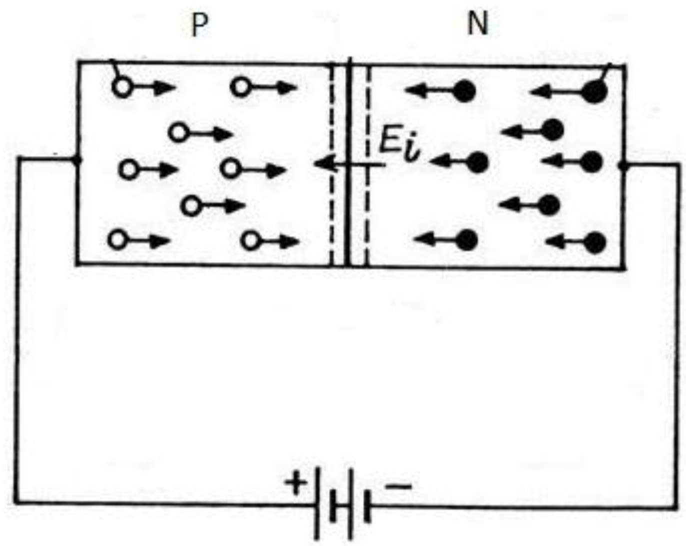
A barrier dip is a p.d which oppose more diffusion of charges across the junction.

This is produced when the flow of +ve and -ve Charges ceases

P – N JUNCTION AS A RECTIFIER:

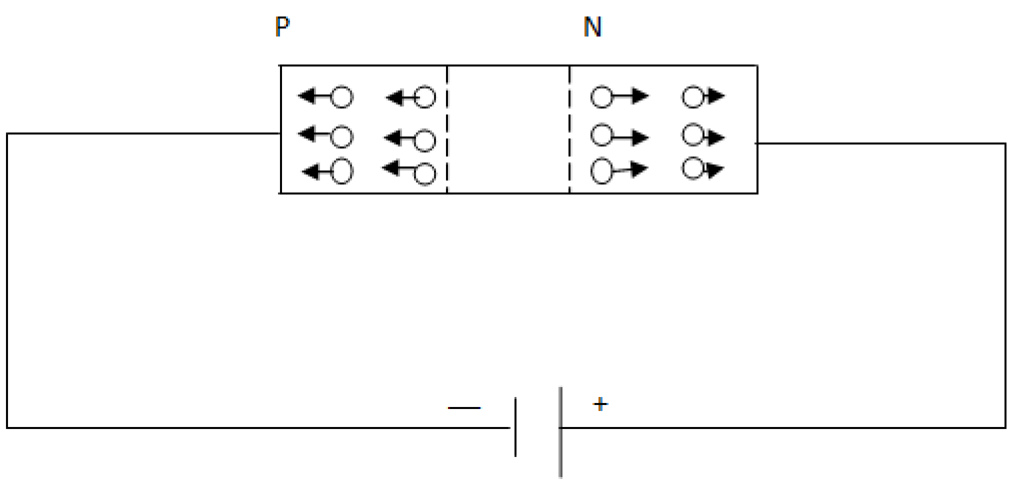
FORWARD BIAS.
Is said to be forward biased when its P- semiconductor is connected to the +ve terminals of the battery and its N- Semi conductors is connected to the -ve terminal at the battery.

In this case electrons and holes flow across the P-n junction. This happen because the +ve pole of the battery repel the +ve charge and –Ve pole rel the –ve charges.

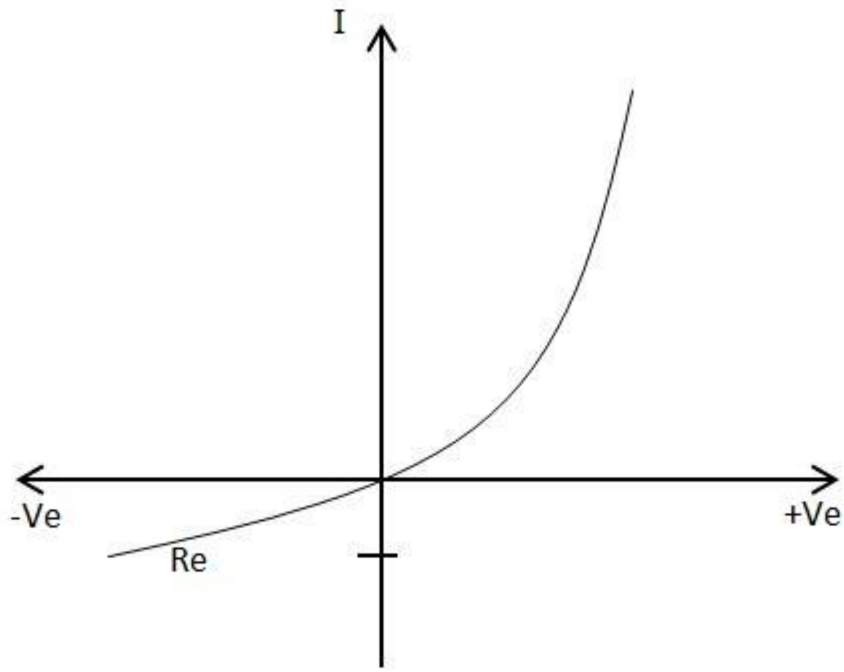


REVERSE BIAS

A-P-N junction is said to be reverse biased when its P. Semiconductor is connected to the negative pole junction of a battery and N. Semiconductor is connected to the +ve [p;e pf the battery in this case only a very small a current flows.



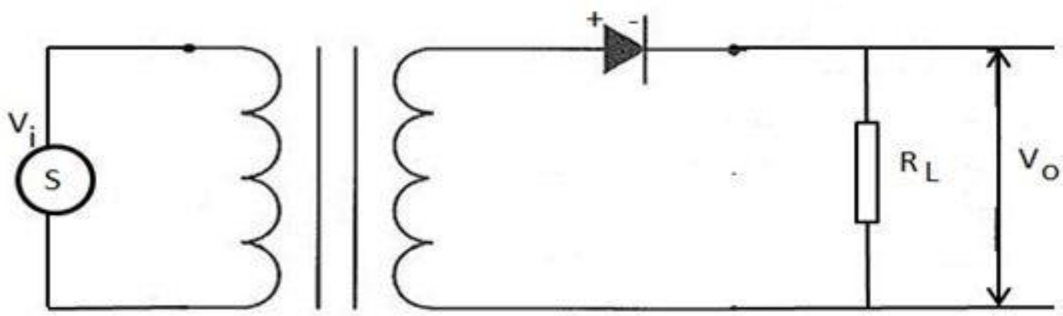
P-N JUNCTION AS RECTIFIERS

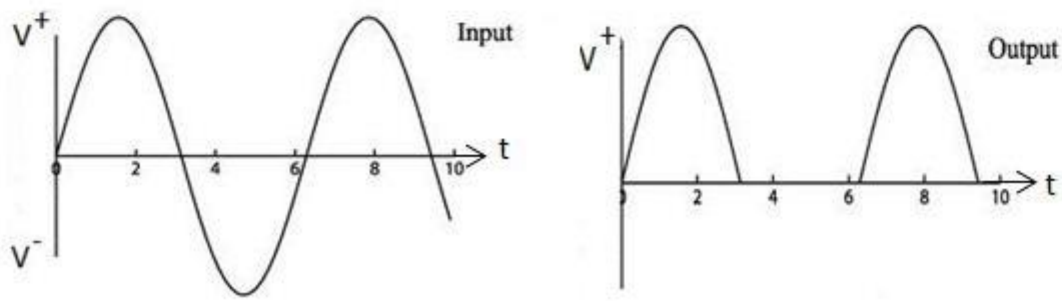


The graph shows that P-N junction acts as a rectifier, it has low resistance in one direction of P.d (ie + v) and higher resistance in the opposite direction of P.d (-v)

RECTIFIER CIRCUITS:

HALF-WAVE RECTIFIER CIRCUIT

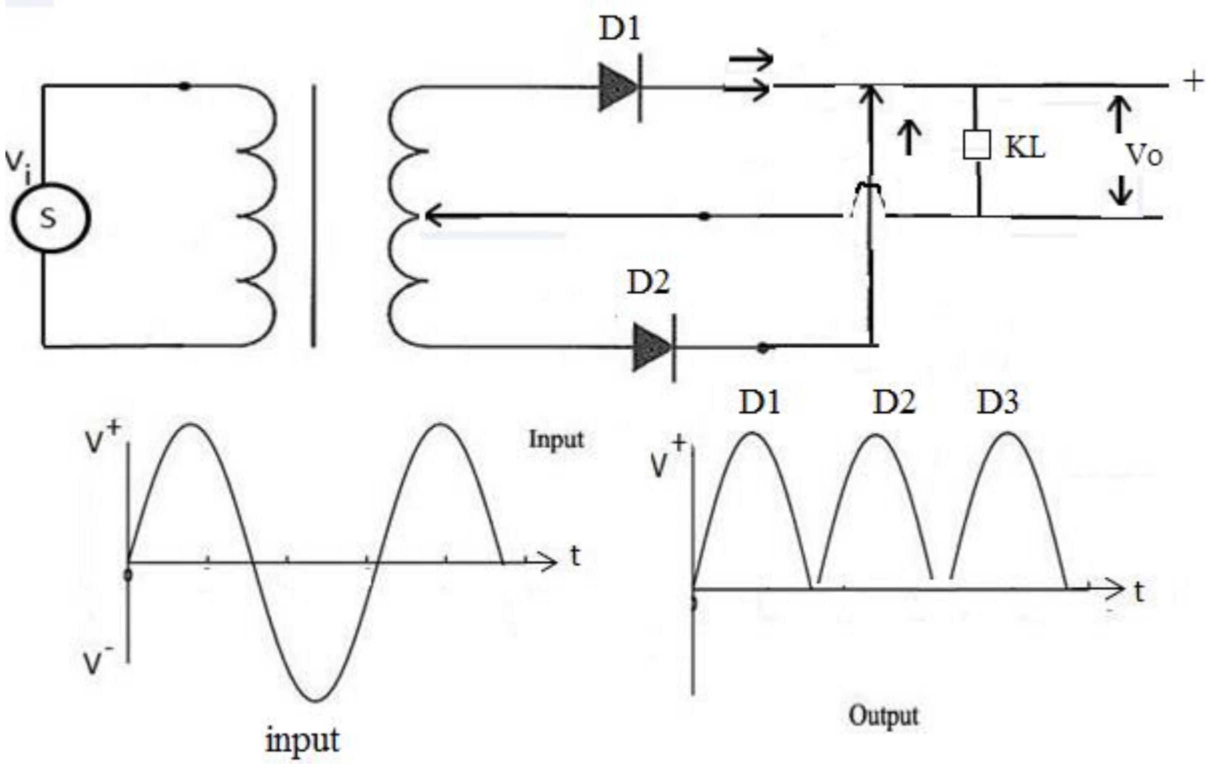




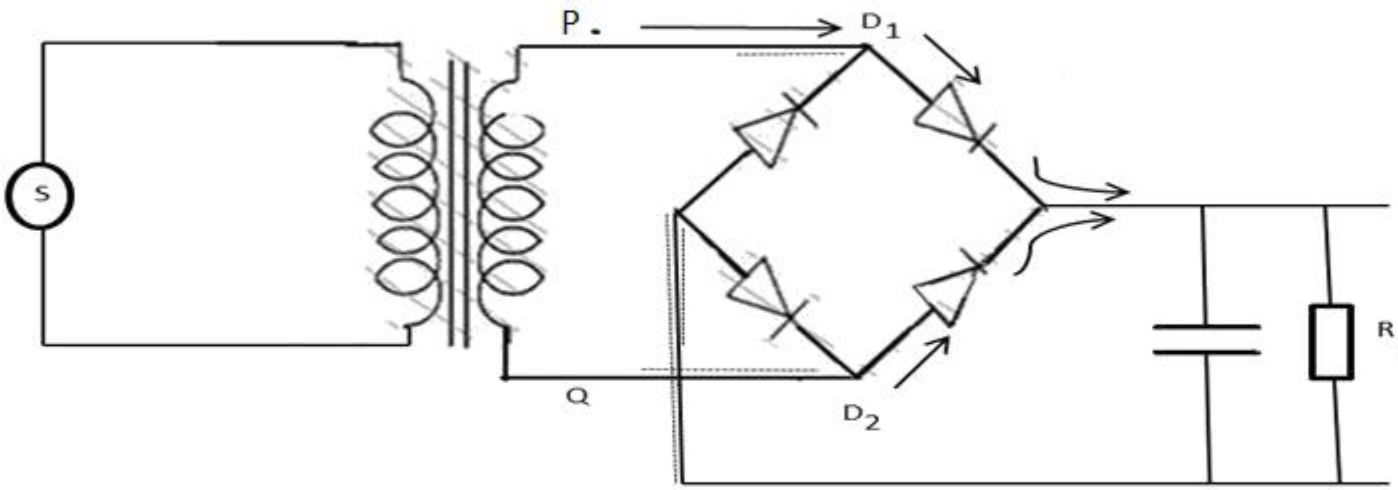
A rectifier is a circuit which allow the flow of current I P.d in one direction only:

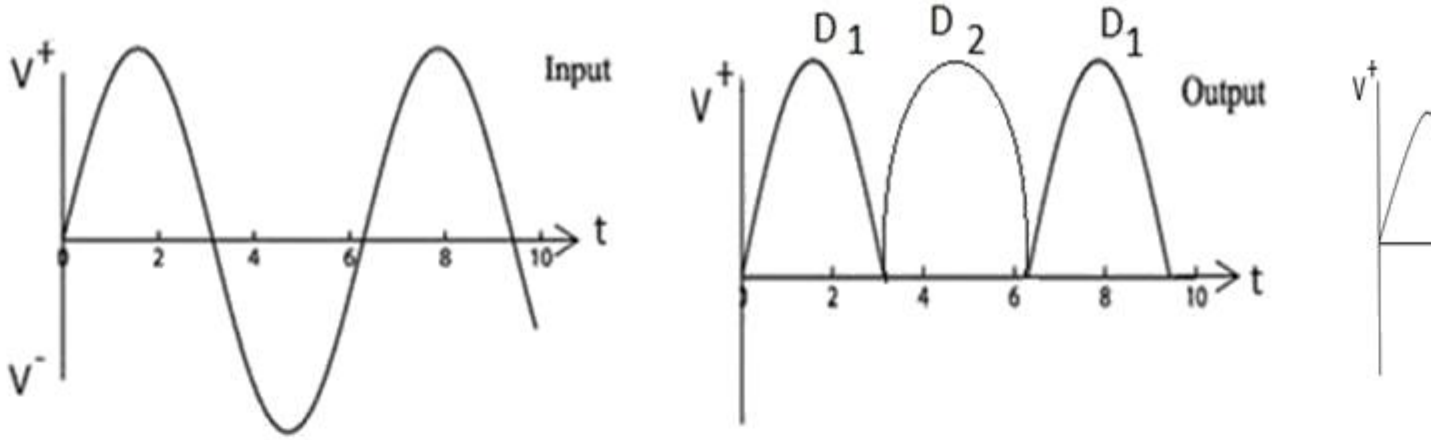
FULL –WAVE RECTIFIERS CIRCUITS:

a) Using centre –tapped transformer.



b) Using bridge circuits





On one half of a cycle when P is +v relative to Q only diode D_1 conducts

On one other half the same cycle only the diode D_2 conducts.

In both cases the current gees through resistor RL in the same direction.

The large capacitor C is used for stabilizing the marying d.c voltage.

B) TRANSISTORS

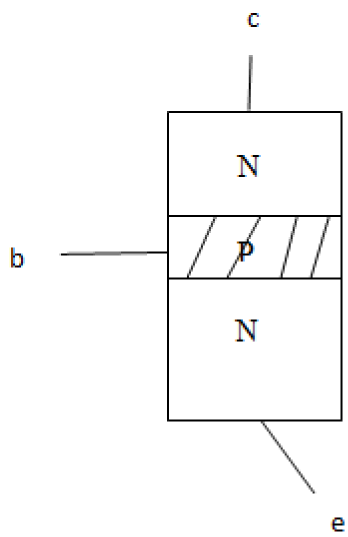
Transistor is a component which amplifies current. It is made from three layers of P and n. Semiconductors. The layers are called the emitter (E) base (B) an collector (C)

There are two types of transistors.

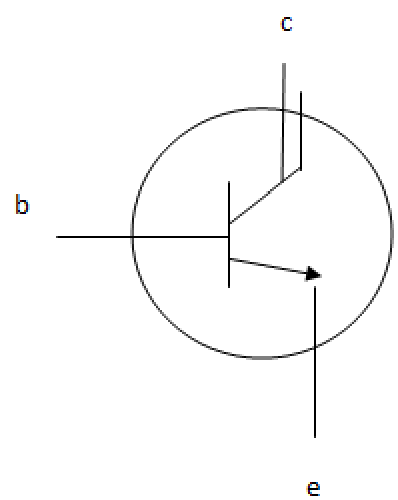
I. n.p.n transistor

II. p.n.p transistor

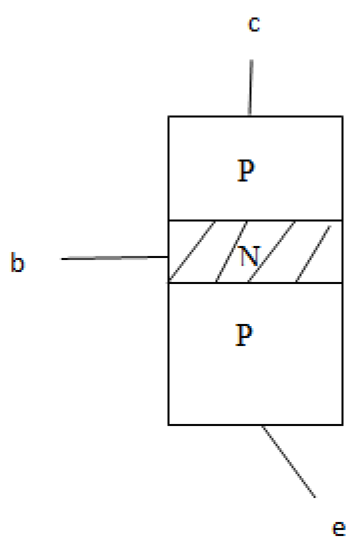
Construction



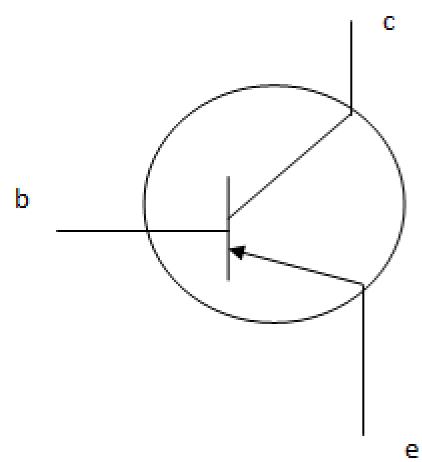
Symbol



Construction



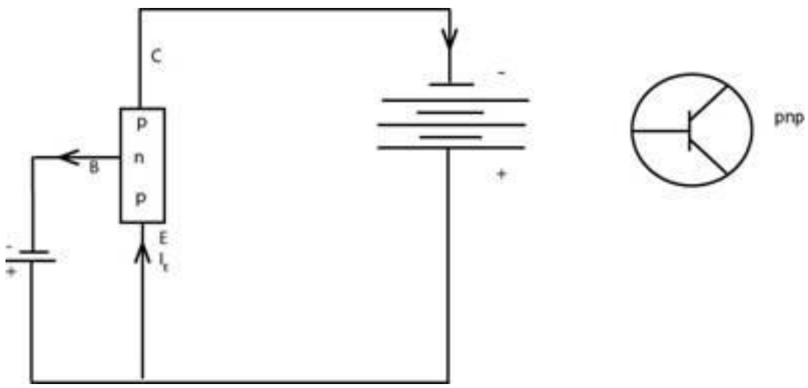
Symbol



Formation of a transistor

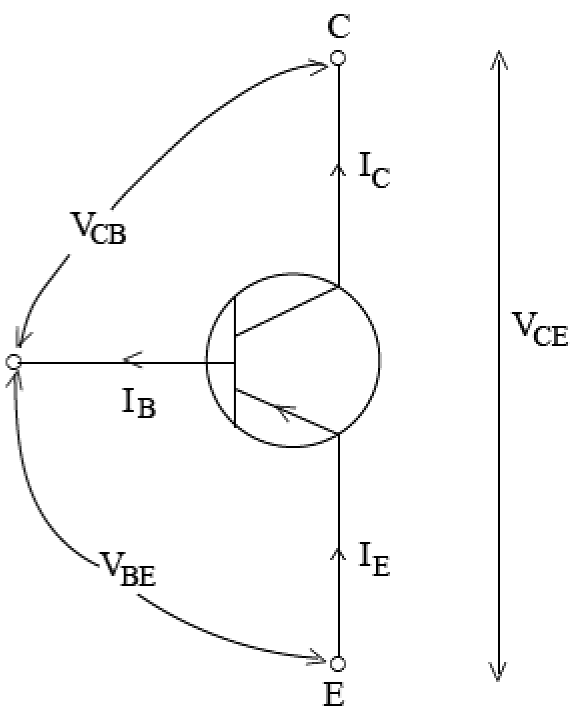
A transistor is formed by putting the doped semiconductors together in such a way that two junction are formed.

The pnp transistor (bipolar transistor).

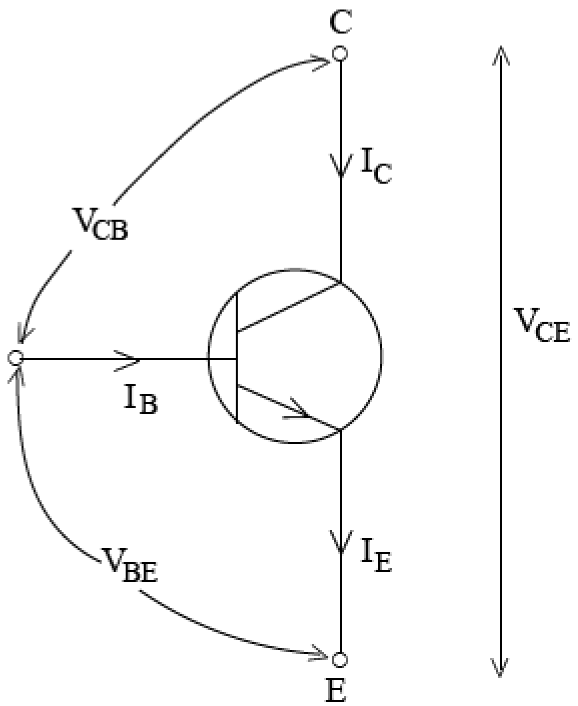


-Bipolar means n p n and p n p transistor as they have two opposite polarity of doped semiconductors and voltages across terminals

P n p



n p n



Transistor configuration

There are 3 basic configurations

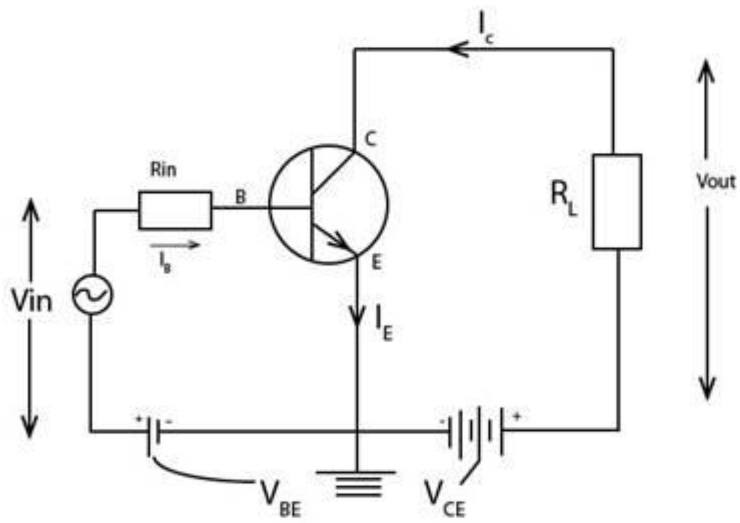
1. Common emitter configuration

2. Common base configuration

3. Common collector configuration

1. Common emitter configuration (n p n)

Under thus configuration the transistor has both voltage gain and current gain.



To get volt you need a resistance R_L

$V_{out} = {}_I R_L$

Ring is used for injecting only a small current for great amplification on E by C

Current gain

$\frac{I_C}{I_B}$ = very large $I_C \ggg I_B$

$$I_E = I_B + I_C$$

$$\frac{I_C}{I_E} = \alpha$$

$$\frac{I_C}{\beta} = I_B$$

$$I_E = \frac{I_C}{\beta} + I_C$$

$$I_E = I_C(\frac{1}{\beta} + 1)$$

$$\frac{I_E}{I_C} = \frac{1}{\beta} + 1$$

But

$$\frac{1}{\alpha} = \frac{I_E}{I_C}$$

$$\frac{1}{\alpha} = \frac{1}{\beta + 1}$$

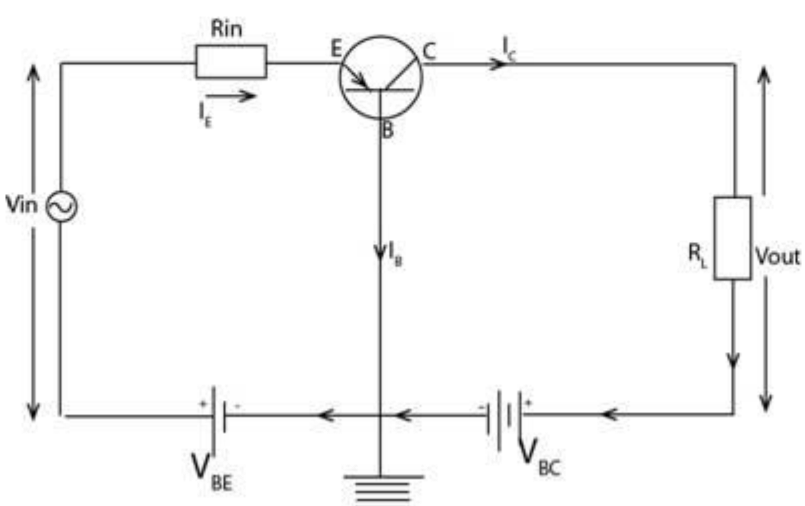
$$\frac{1}{\alpha} = \frac{\beta + 1}{\beta}$$

Also

$$\beta = \frac{\alpha}{1 - \alpha}$$

V_{out} is the reflection of V_{in}

Common base configuration (PnP)



Under this configuration the transistor has voltage gain but no current gain

Earthing puts the common line at p.d=0

$$V_{in} = R_{in} I_E$$

$A_V = Voltage\ gain_{=} \frac{voltage\ output}{voltage\ input}$

$\frac{R_L}{R_{in}} = Resistance\ gain$

$$A = \frac{I_C R_L}{I_E R_{in}}$$

Common collector configuration

Under this configuration the transistor has current gain but no voltage gain

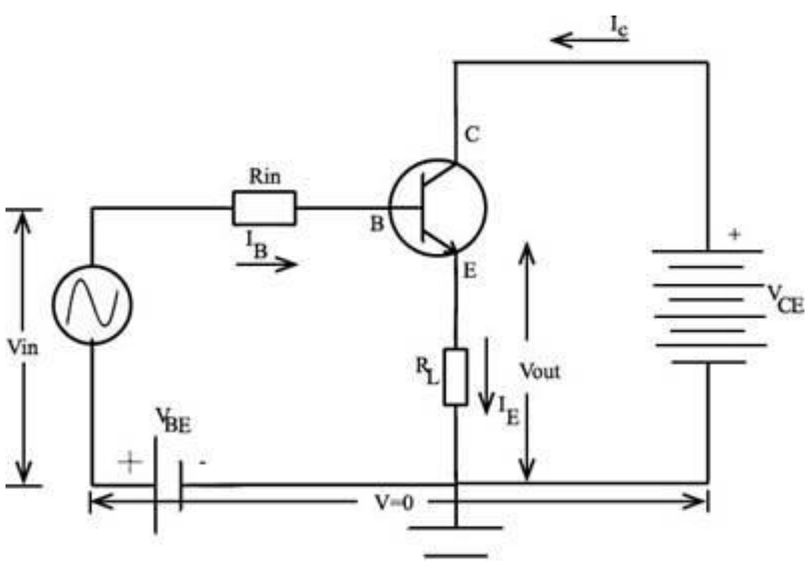
A_F=Amplification factor

$V_{o=}I_C R_L$

$I_{E=}I_{C+}I_B$

$$A_f = \frac{I_E}{I_B} = \frac{I_C + I_B}{I_B}$$

$$A_f = \frac{I_C}{I_B} + 1$$



$$A = \frac{V_{out}}{V_{in}} = 1$$

For common emitter

$V_{o} = I_C R_2$

$V_{i} = I_B R_B$

$$I_E = I_B + I_C$$

But

$$I_C \alpha I_B$$

$$I_C = \beta I_B$$

$$\beta = \frac{I_C}{I_B}$$

$$AV = \frac{V_o}{V_{in}} = \frac{I_E R_L}{I_B R_B} = \left(\frac{I_E}{I_B}\right) \times \left(\frac{R_L}{R_B}\right)$$

$$AV = \beta \left(\frac{R_L}{R_B}\right) \quad Voltage\ Amplification$$

$$From\ \frac{I_C}{I_E} = \alpha\ and\ \frac{I_C}{I_B} = \beta$$

From, $I_E = I_B + I_C$

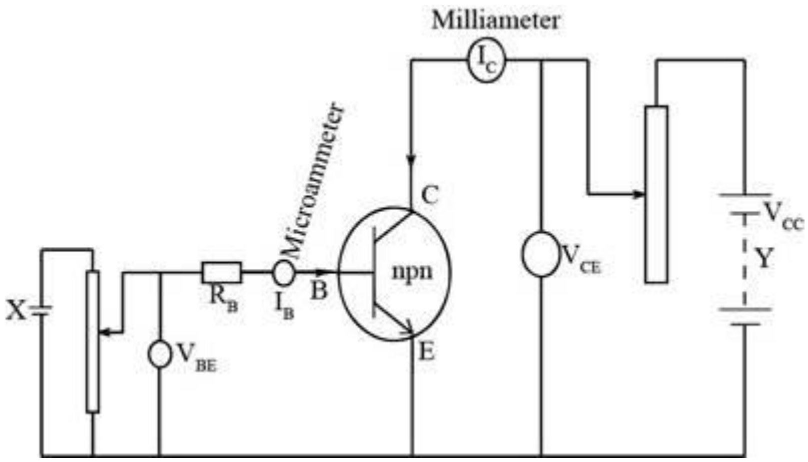
$$\frac{I_E}{\alpha} = \frac{I_C}{\beta + I_C}$$

$$\frac{1}{\alpha} = \frac{1}{\beta + 1}$$

$$\propto = \frac{\beta}{\beta + 1}$$

$$\beta = \frac{\propto}{1-\propto}$$

Common emitter characteristic curve

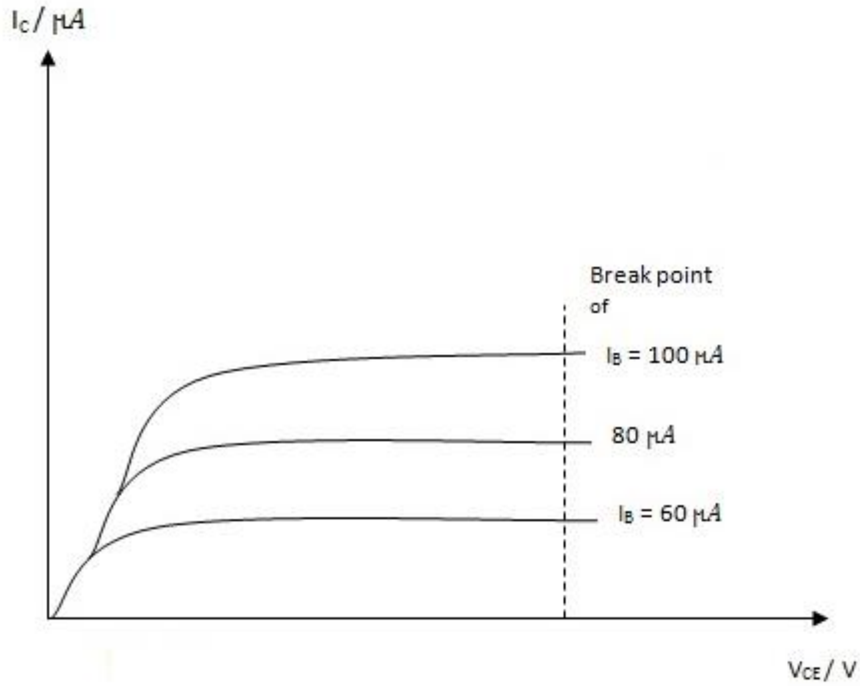


The circuit above is for investigating the variation of current with voltage in the input and output circuits.

OUTPUT CHARACTERISTICS
I_C-V_{CE} with I_B constant .

The results are plotted below.
The knee of the curves shown corresponds to a low P.d(0.2) the output for higher P.d the output I_C varies linearly with V_{CE} for a given value of base current I_B.

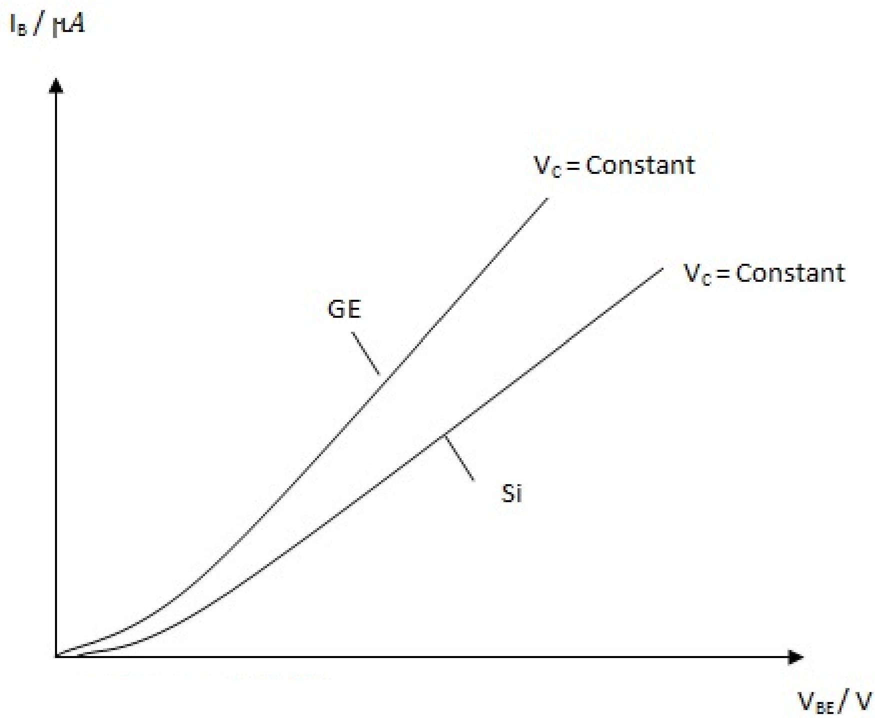
The linear part of the characteristic is the one used in the audio frequency (a.f) amplifiercircuits so that the output is undistorted.



INPUT CHARACTERISTICS

I_B-V_{BE} with V_{CE} constant

The results are as follows:-
The input characteristics is non-linear



TRANSFER CHARACTERISTICS
I_C-I_B with V_{CE} constant

The results are as shown below:-
The output current I_C varies linearly with the input current I_B. The current transfer ratio or current gain is given by

$$\frac{\Delta I_C}{\Delta I_B} \text{ for a.c signal conditions}$$

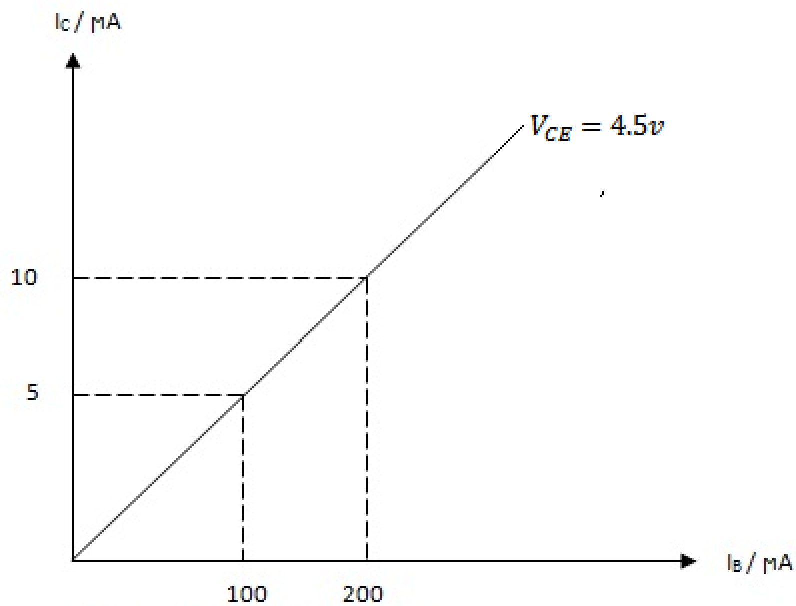
$$\text{for d.c signal conditions } \beta = \frac{\Delta I_C}{\Delta I_B}$$

In the figure below

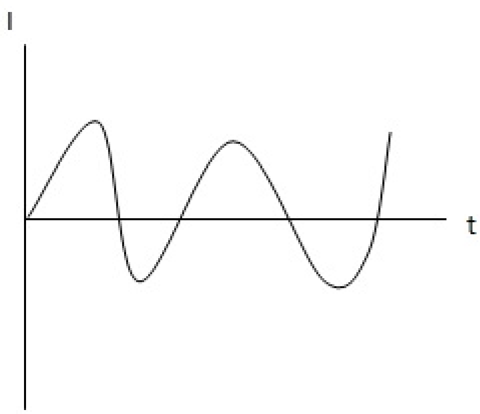
$$\beta = \frac{\Delta I_C}{\Delta I_B} = \frac{10 - 5}{200 - 100} = 50$$

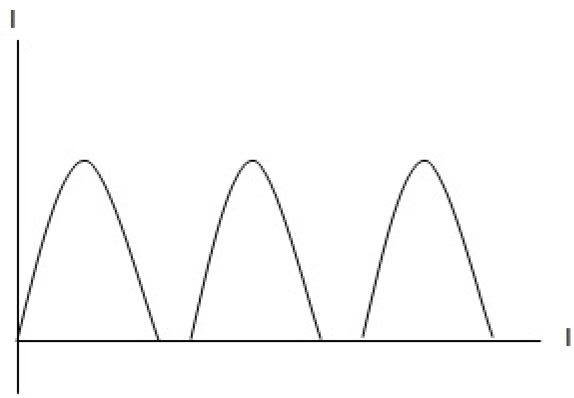
$$\beta = 50$$

$So\ the\ current\ has\ been\ amplified\ 50\ times$ β is also written as life



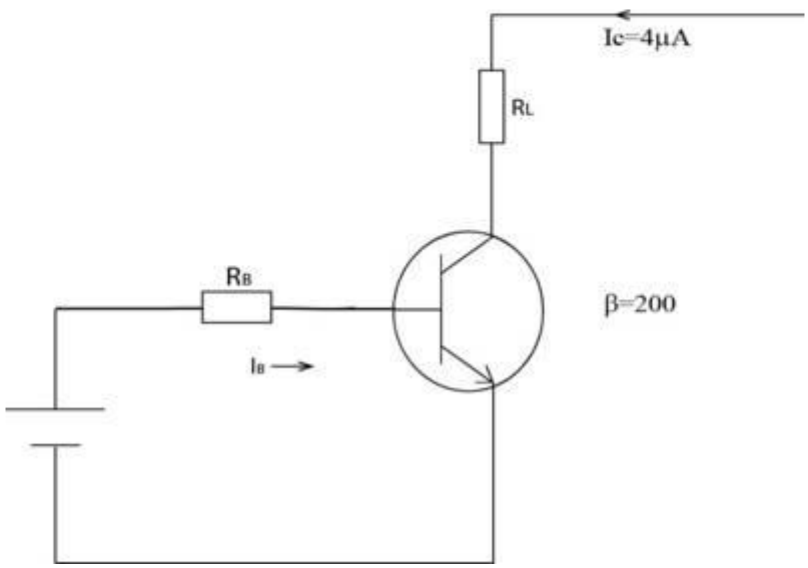
For common emitter (complification)





Questions

1. An npn transistor has a current gain (Beta) value of 200. Calculate the base current I_B required to switch a resister of 4μA.



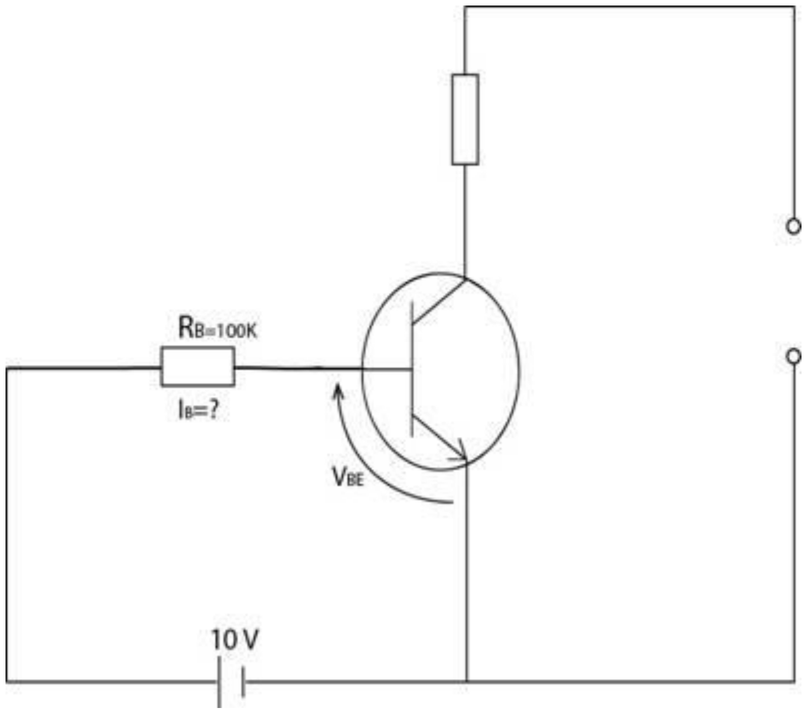
$$\beta = \frac{I_C}{I_B}$$

$$I_B = \frac{I_C}{\beta} = \frac{4 \times 10^{-3}}{200} = 2 \times 10^{-5} A$$

$\boldsymbol{I_{B=20\mu A}}$

2. An npn transistor has a dc base bias voltage V_B of and an input base resistor R_B of 100kΩ.What will be the of base current into the transistor?

(The transistor is a silicon type)



Data

From Kirchhoff$'s\ 2^{nd}$ law

For silicon

V_B =10v

R_B = 100kΩ

V_{BE} =0.6v (wasted voltage)

Solution

$V_B - I_B R_B - V_{BE} = 0$

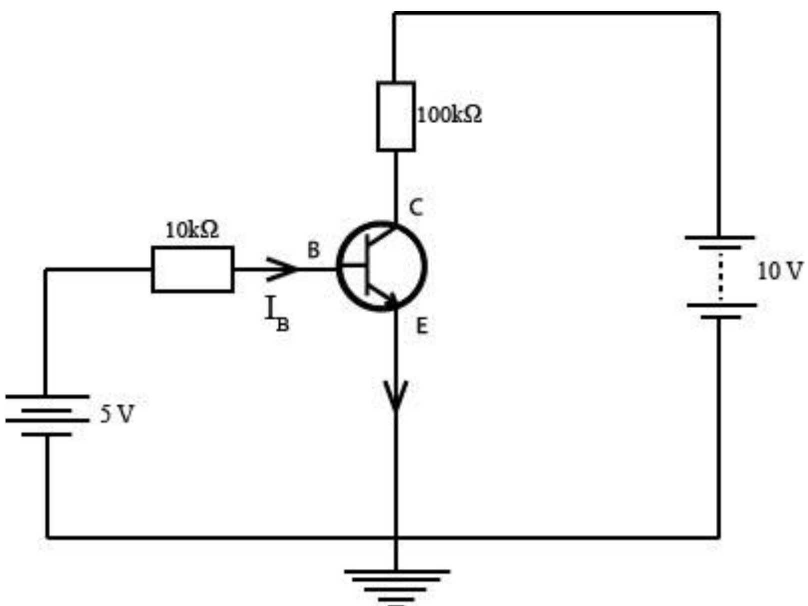
$V_B - V_{BE} = I_B R_B$

$$I_B = \frac{V_B - V_{BE}}{R_B}$$

$$I_B = \frac{10v - 0.6v}{100k\Omega} = 9.4 \times 10^{-5}\text{A}$$

Example

Given the circuit below, determine



I_B

I_C

V_{EB}

V_{CE}

V_{CB}

The transistor has β =150

The transistor of silicon type

Solution

Second Kirchhoff's law in the input circuit

$V_B - I_B \times 1000 - V_{BE} = 0$

$5V - 0.6V = I_B \times 1000$

$I_B = \frac{5V - 0.6V}{1000}$

$I_B = 4.4 \times 10^{-4}$ A

$I_B = 4.4\mu$A

$\frac{I_C}{I_B} = \beta$

$I_C = \beta \times I_B$

$I_C = 150 \times 4.4 \times 10^{-4}$

$I_C = 0.066A$

$\boldsymbol{I_C = 66mA}$

$V_{EB} = 0.6v$

$V_{CL} - I_C R_L - V_{CE} = 0$

$10 - (6.6 \times 10^{-2} A \times 100\Omega) - V_{CE} = 0$

V_{CE} = **3.4V**

$(10 - 5) - I_C R_L - V_{CB} + I_B R_B = 0$

V_{CB} =5+ ($I_B R_B \times I_C R_C$)

V_{CB} =5+ (10000×4.4×10^{-4}) $-$ (6.6)

V_B=**2.8v**

Example

1. A common emitter amplifier has R_L = 1.2kÎ• and supply Voltage of V=12v. Calculate the maximum collector current I_C following throughout resistor when switched fully on (saturation assume $V_{CE} = 0$.Also find $R_E$with a voltage drop of 1v across it, the transistor silicon.

Solution

$$V_{CC} = I_C R_L - V_{CE} + 1$$

$$V_{CC} = 12v$$

$$V_{CE} = 0v$$

$$\mathrm{I_C} = \frac{\mathrm{V_{CL}} - 1}{\mathrm{R_L}}$$

$$I_C = \frac{12-1}{1200}$$

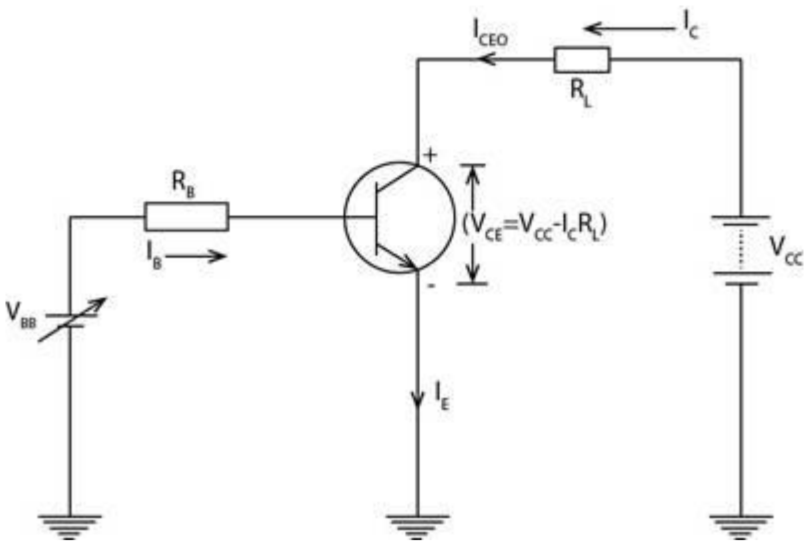
$$= \mathbf{9.167 \times 10^{-3}}\text{ A}$$

Quiescent point:

It's a point when the current flow is smooth i.e. not being clicked (excess) and transistor functions.

Saturation point:

If I_B=0 transistor is in the cutoff region, there is a small current collector leakage, CEO



Normally I_{CEO} is neglected so that V_{CE} = V_{CC}

In cutoff both the base emitter and base collector junction are reverse based.

When base emitter becomes forward based. I_B is increase, then IC also increases when V_{CE} decreases as a result.

When V_{CE} reaches its saturation value BC junction becomes forward based and I_C can increase no further even with continued increase in I_B

At the point of saturation ($I_C = \beta I_B$) not longer valid)

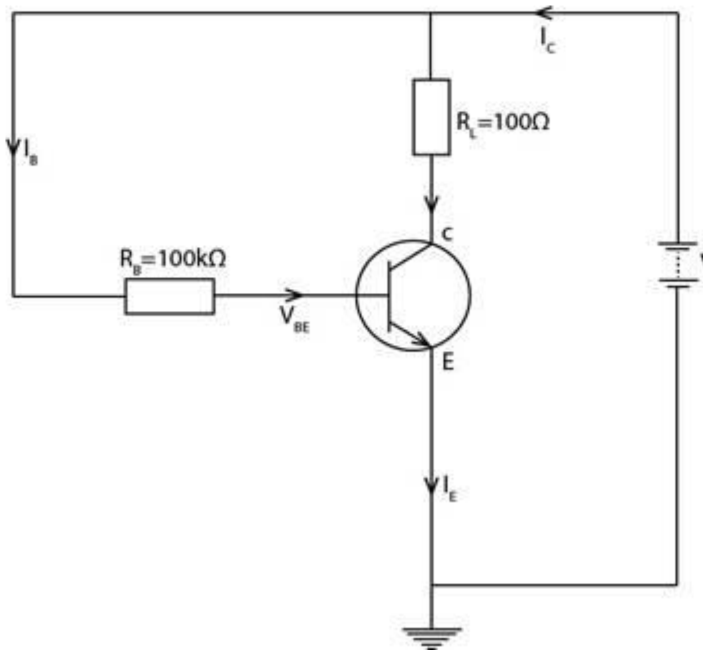
$V_{CE(Sat)}$ for a transistor occurs somewhere below the knees of the collector curve.

The saturation value for V_{CE} (Sat) is usually a few tenth of volt for silicon transistors.

The DC load line, the cutoff and saturation can be illustrated by the load line.

Between the cutoff point and the saturation point is where the transistor is active and as most active at the quiescent point.

Self biasing /fixed bias



I_BR_B-(V_{CB})-(-I_CR_C)=0

$V_{CC} = I_C R_L + V_{BE}$

$V_{CC} = I_C R_L + I_E R_E$

$V_{BE} = 0.06V$

Outer loop

$I_B R_B = V_{CC} + V_{BE}$

$I_B R_B = 10 - 0.6$

$I_B = \frac{9.4}{R_B} = 9.4 \times 10^{-5}$ A

$\beta = \frac{I_C}{I_B}$

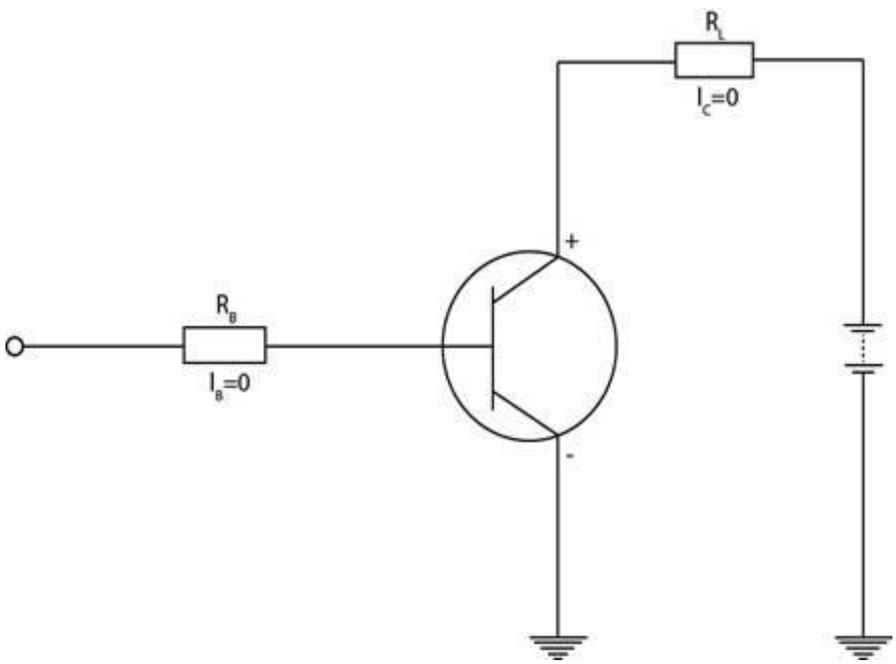
$100 = \frac{I_C}{9.4 \times 10^{-5}}$

$I_C = 9.4 \times 10^{-3}$ A

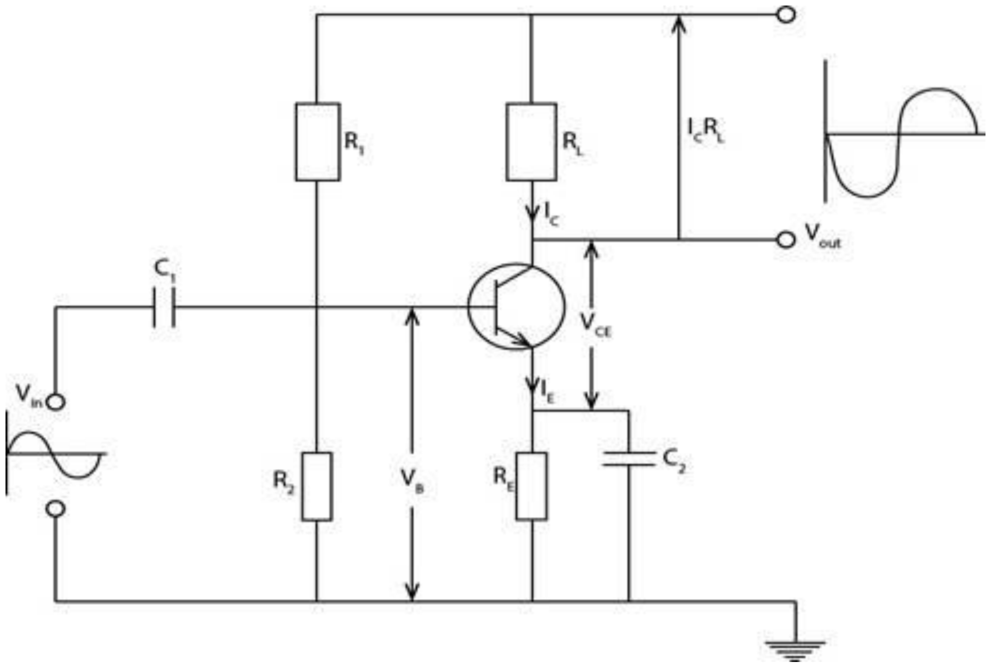
$V_{CC} = I_C R_L + V_{CE}$

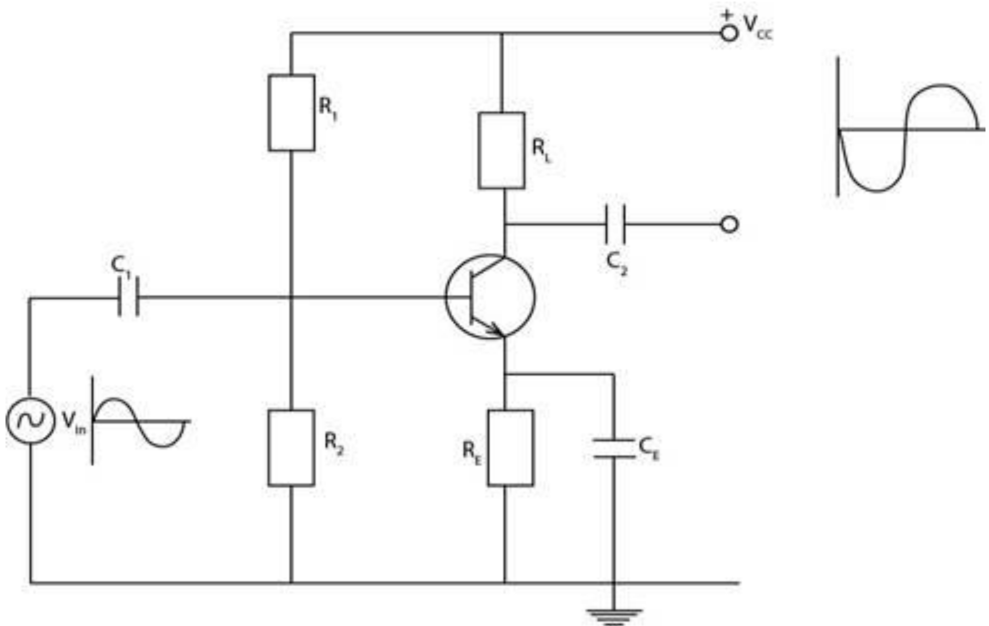
10 = $9.4\times10^{-3}\times100 + V_{CE}$

V_{CE} = **9.06V**



Common emitter amplifier circuit



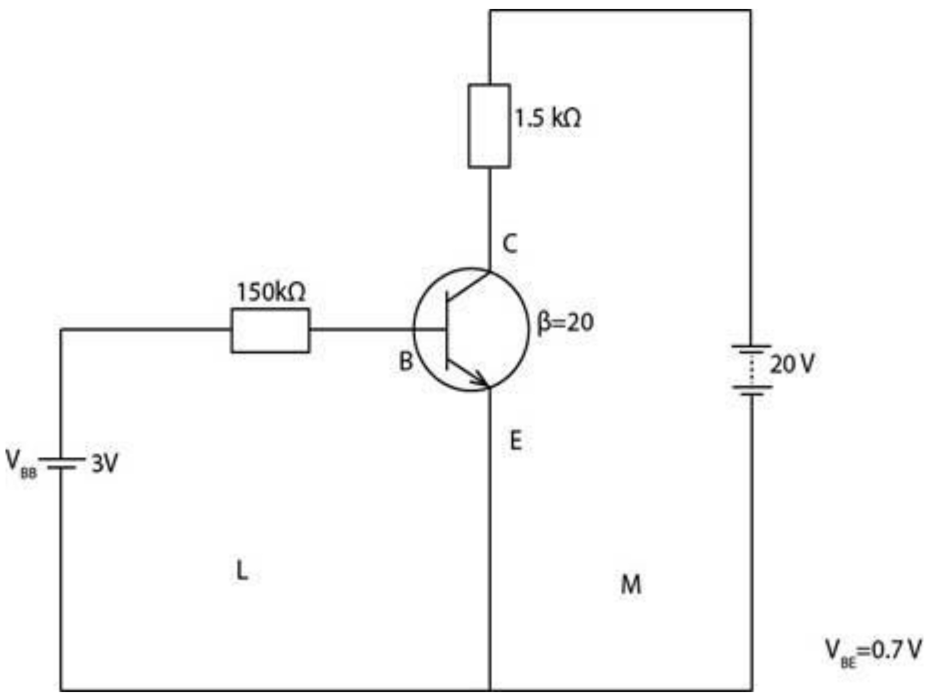


$$A_V = \frac{V_o}{V_{in}}$$

$$\beta = \frac{I_C}{I_B}$$

Faithful amplification- is the application or the output that is not distorted.

Question



a) Pd across base resistor

Consider loop (L), from Kirchhoff's law

3V$-V_{RB} - V_{BE}$=0 but V_{BE}=0.7

V_{RB}**= 2.3V**

b) From Ohms law

$V_{B=}I_B \times R_B$

$I_{B=}\frac{2.3}{150\times10^{-3}}$

$I_B = (\frac{2.3}{150\times10^{-3}})$**= 1.53**$\times \mathbf{10^{-2}\mu A}$

$c)V_{CC=}I_C R_{L+} V_{CB} - I_B R_B + V_{BB}$

V_{CB} = $(V_{CC} - V_{BB})$ + $(I_B R_B - I_C R_L$

V_{CB} = (20- 3) + (2.3-1.836)

V_{CB}= (17+0.463)

V_{CB} =**17.463v**

d) Find V_{CE}

From Kirchhoff's law

Given

V_{CC}-$I_C R_L$-V_{CE}=0

β=20

V_{CC}=20v

So

20v-(1.5$\times 10^3$×1.224$\times 10^{-3}$)-V_{CE}=0

V_{CE}=**18.164v**

2. (a)Q (V_{CE}, I_C)= $(7.8,47)$=

(b) V_{CC} =$I_C R_L$ +V_{CE}

V_{CC} = 25V

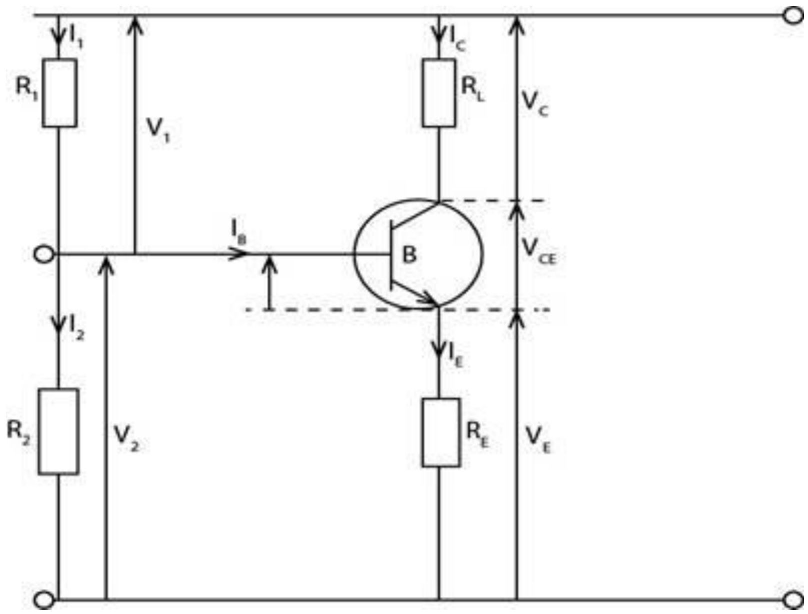
I_C = 47mA = 4.7×10^{-2}A

R_L = $\frac{V_{CC}-V_{CE}}{I_C}$

= $\frac{25-7.8}{47}$

R_L =**0.3659Ω**

Question



For the circuit above the transistor has a current gain β=80 the collector supply voltage V_{CC} = 40 . The required biased conditions are V_{BE}= 0.7V and I_C= 1mA. Determine the suitable values for resistors R_1, R_2, R_E & R_L,

$R_2 = 10R_E$, $V_E = 1$

$V_E = I_V$.

Given $R_2 = 10R_E,\ V_E = 1_V$

I_C= 1×10 $^{-3}$A

β=80

I_B =0.0000125

I_E=$I_B + I_C$ = 0.0010A

$R_E = \frac{1V}{1\times10^{-3}}$=1kΩ

R_2=1×10

$= \mathbf{10K\Omega}$

$V_{BE} = V_B - V_E$

$V_B = V_{BE} + V_E = 0.7 + 1$

V_B = 1.7 but

$V_{CE} + V_C = 9$

$V_{CE} + V_{C} + V_{E}$=10

But

$V_{CE}=V_C - V_E$

Therefore

V_C -V_E+V_C -V_E =10

$2V_C = 10$

$V_C = \mathbf{5v}$

Since $V_C=R_L \times I_C$

$R_L= \mathbf{5k\Omega}$

$I_1=I_B + I_2$

From ohms law

$I_2=\frac{V_2}{R_2}=\frac{1.7}{10\times10^3}=1.7\times10^{-4}$

$V_B = V_2$

$I_1= 1.7\times10^{-4}\text{A} + 1.25 \times 10^{-5}\text{A}$

$I_1=\mathbf{1.825}\times 10^{-4}$

Then

$$0 = I_C R_L + V_{CB} - I_1 R_1$$

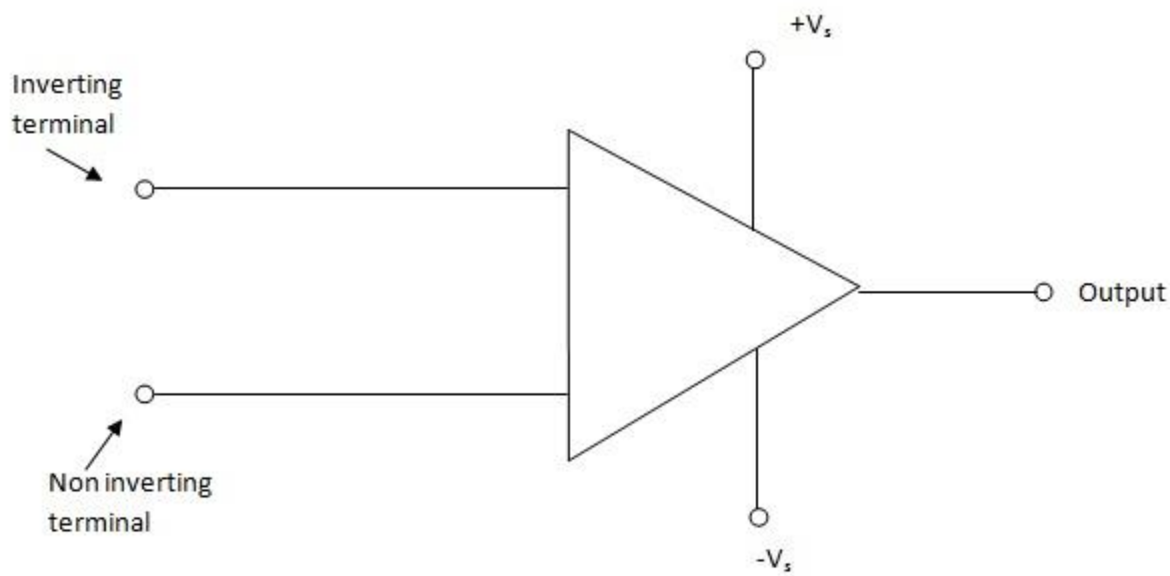
$$R_1 = \frac{V_{CB+I_{C}R_L}}{I_1} \text{ but } V_{CB} = V_C + V_B = 1.7+5=6.7$$

$$R_1 = \frac{6.7+(5\times10^3\times1\times10^3)}{1.825\times10^{-4}}$$

Operational Amplifier (Op amp)

An operational amplifier (op amp) is an electronic device consist of a large number i.e. twenty and above.

It has 3 terminals two input terminals and one output terminal.



The op amp can perform electronically mathematically of such as additional, subtraction, multiplication, differentiation, integration

Properties of an op- amp

i) It has got a very high voltage gain called the open loop gain which typically is 10^5 for dc and low frequency but decrease with frequency.

ii) It has a very high input resistance typically 10, it draw a minute current from the signal source.

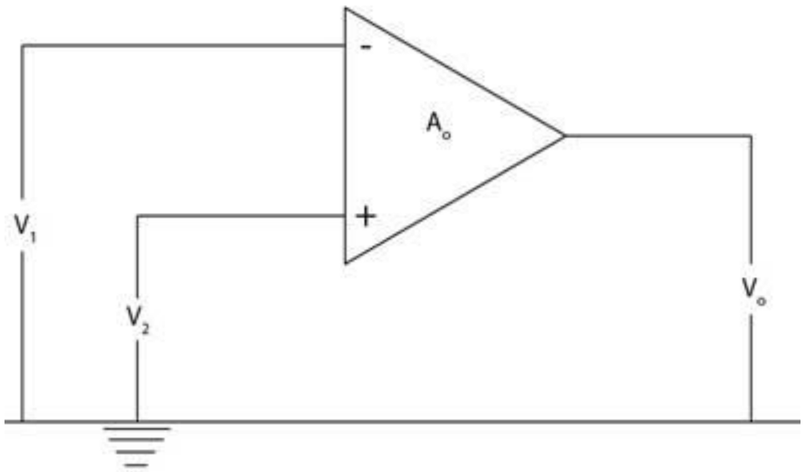
iii) It has a very low output resistanceR_0, typically 100Î• .

Description

It has one output and two inputs and one non inverting (+) and one inverting (-).

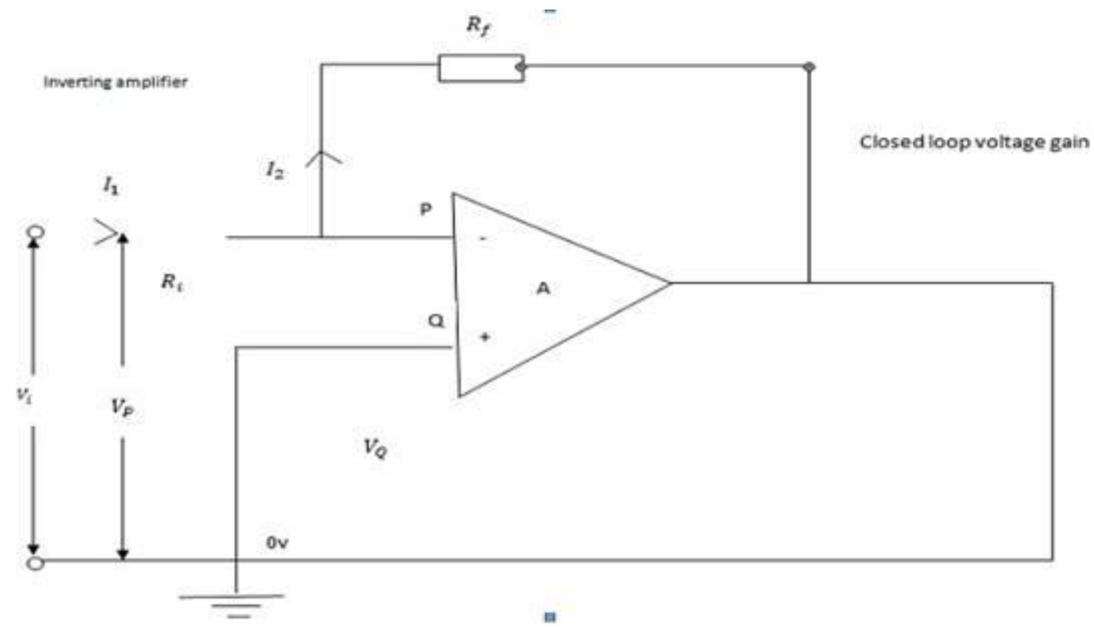
Its operation must convenient from a dual balanced power supply giving its equal +ve and –ve voltage (+Vs, or,-Vs)

Inverting amplifier



$V_o = A_O(V_2 - V_1)$

$A_o = 10^5$



Some of the output goes back to the input .This red called the amplification from A_0 to A

$$I_1 = \frac{V_i - 0}{R_i}$$

$$I_2 = \frac{0 - V_0}{R_f}$$

But

$I_1, = I_2$

$$\frac{V_i - 0}{R_i} = \frac{0 - V_0}{R_f}$$

$$\frac{V_i}{R_i} = \frac{-V_0}{R_f}$$

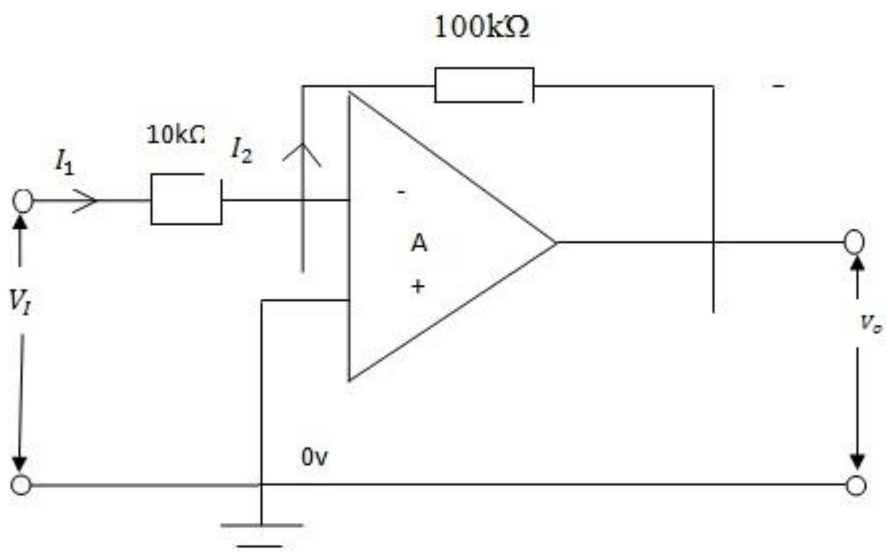
$$\frac{R_f}{R_i} = \frac{-V_0}{V_i}$$

$$\frac{V_0}{V_i} = \frac{-R_f}{R_i}$$

$$But\ A = \frac{V_0}{V_i}$$

$$Voltage\ gain, A = \frac{-R_f}{R_i}$$

Example



(i)Find the closed loop gain of the inverting amplifier

From

$$I_1 = \frac{V_o - 0}{R_i}$$

$$I_2 = \frac{0 - V_i}{R_f}$$

$$\frac{V_o - 0}{R_i} = \frac{-V_f}{R_f}$$

$$\frac{V_o}{R_i} = \frac{-V_f}{R_f}$$

$$\frac{V_o}{V_f} = \frac{-R_f}{R_i}$$

$$But\ A = \frac{V_0}{V_i}$$

$$A = \frac{-R_f}{R_i}$$

$$\mathrm{A} = \frac{-100\mathrm{k}\,\Omega}{10\,\mathrm{k}\,\Omega}$$

= -10

ii) Supposed the voltage gain is to be increased to 40 and the current of R_i remains the same .What are the values of the resistors required to gain this

$$A = 40$$

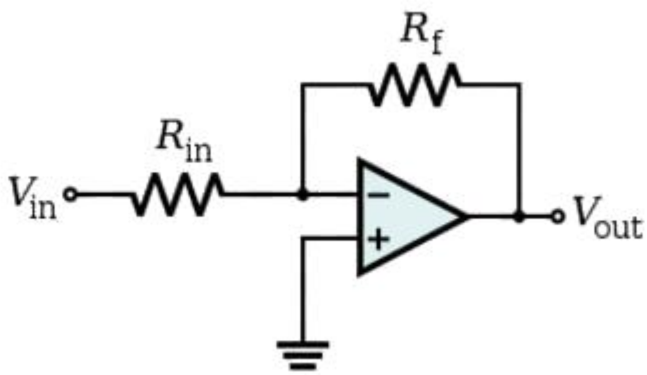
$$A = \frac{-R_f}{R_i}$$

$$R_f = -AR_i$$

$$R_f = 40 \times 10k\Omega$$

$$R_f = 400\,k\Omega$$

<u>NON INVERTING AMPLIFIER</u>

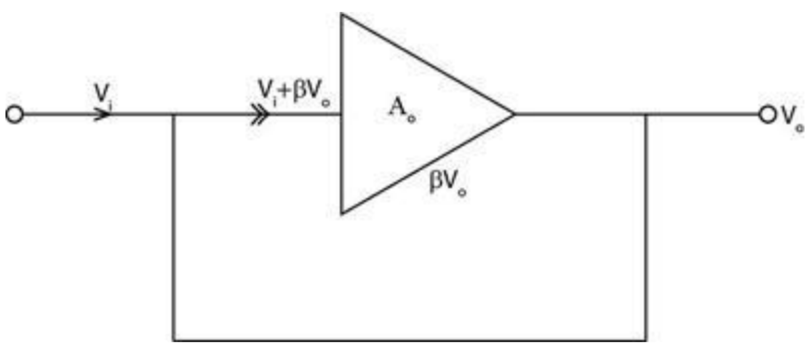


The fraction β fed back via *Rf*

$$\beta = \frac{R_i}{R_i + R_f}$$

$$A = \frac{R_i}{R_i + R_f} = \frac{R_i}{R_f} + \frac{R_f}{R_i} = \frac{R_f}{1 + R_i}$$

LOGIC GATES (Non inverting)



$$V_O = A_O(V_i + \beta V_O)$$

$$V_O = A_O V_{i\,+} A_O \beta V_O$$

$$A_O V_i = -A_O V_O \beta + V_O$$

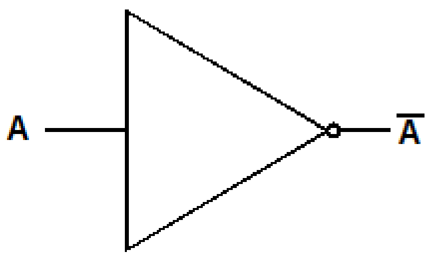
$$A_O V_i = V_O(1 - \beta A_O)$$

$$\frac{V_0}{V_i} = \frac{A_0}{1 - \beta A_0}$$

LOGIC GATES

SYMBOL

i. NOT GATE(INVERTER)



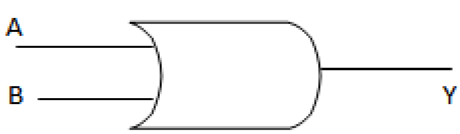
he

It has only one input and one output.

$Input(A)$	$Input\ (\bar{A})$
0	1
1	0

ii. OR GATE

This can have many number of inputs but only one input. It gives high output if either of the inputs is high or all inputs are high.

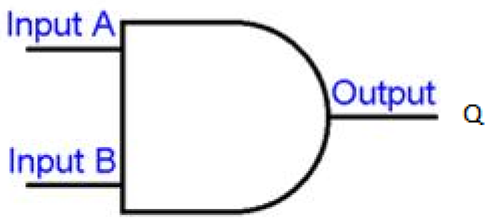


TRUTH TABLE FOR OR GATE

Inputs		Output
A	B	Y
0	0	0
0	1	1
1	0	1
1	1	1

iii. AND GATE

It can have many number of inputs but only only one output. It gives high output when both input are high.

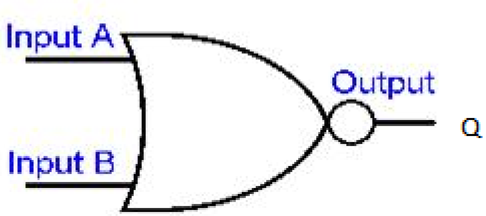


TRUTH TABLE FOR AND GATE

Inputs		Output
A	B	*Q*
0	0	0
0	1	0
1	0	0
1	1	1

iv. NOR GATE

This is equivalent to OR gate followed by NOT gate.All outputs of OR gate are inverted

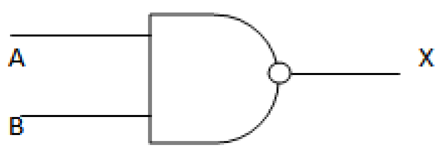


TRUTH TABLE FOR NOR GATE

Inputs		Output
A	B	R
0	0	1
0	1	0
1	0	0
1	1	0

v. NAND GATE

This is the AND gate followed by NOT gate . This is widely used gate . In this case the outputs of AND gate re inverted.



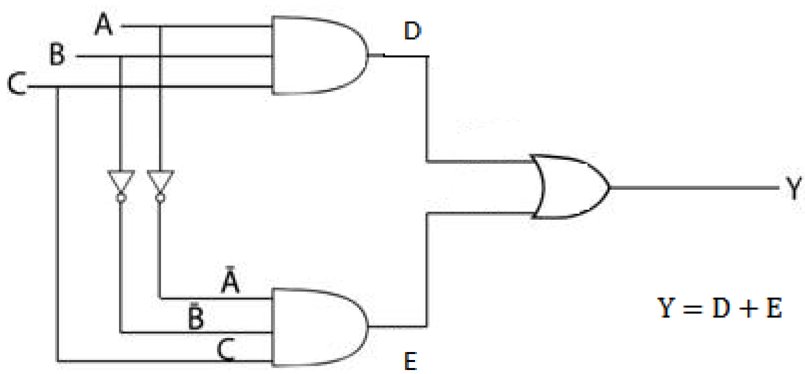
TRUTH TABLE FOR NAND GATE

Inputs		Output
A	B	X
0	0	1
0	1	1
1	0	1
1	1	0

All logic gates described can be connected together to form different function
(i)They are used to control traffic light
(ii)They are used in communication system
(iii)They are used in arithmetic and data processing

Questions

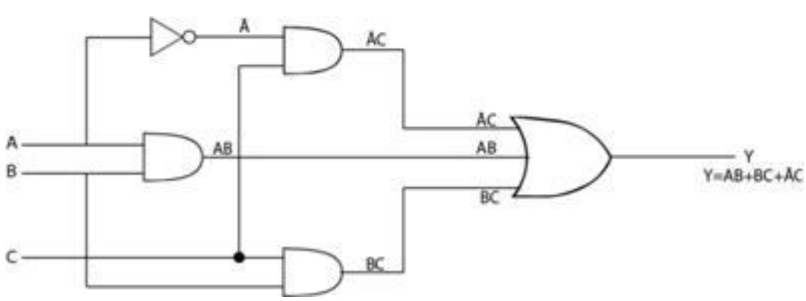
1. Find the expression for Y and form the truth table of the following diagram.



Solution

A	B	C	$\bar{A}$	$\bar{B}$	D	E	Y
1	1	1	0	0	1	0	1
1	1	0	0	0	0	0	0
1	0	1	0	1	0	0	0
1	0	0	0	1	0	0	0
0	1	1	1	0	0	0	0
0	1	0	1	0	0	0	0
0	0	1	1	1	0	1	1
0	0	0	1	1	0	0	0

2. From the logic circuit below form the Boolean expression and draw the truth table

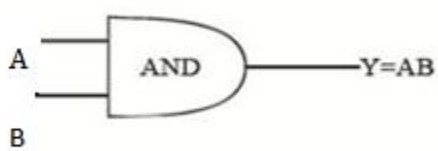


Solution

Truth Table

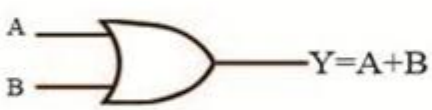
A	B	C	Y
1	1	1	1
1	1	0	1
1	0	1	0
1	0	0	0
0	0	0	0

A	Y
0	1
1	0



A	B	Y
0	0	0
0	1	0
1	0	0
1	1	1

A	B	Y
0	0	1
0	1	1
1	0	1
1	1	0



A	B	Y
1	0	1
0	1	1
0	0	0
1	1	1

Laws of Boolean algebra

T_1: Commutative law

a) A+B=B+A

b) AB=BA

T_2: Associative law

a) (A+B) +C=A+ (B+C)

b) (AB) C=A (BC)

T_3: Distributive law

a) A (B+C) =AB+AC

b) A+BC= (A+B) (A+C)

T_4: Identity law

a) A+A+=A

b) AA=A

$T_5: a) AB + \overline{AB} = A$

$b) (A + B)(A + \bar{B}) = A$

T_6: Redundancy law

a) A+AB=A

b) A (A+B) =A

T_7: a) 0+A=A

b) 0A=0

T_8: a) I+A=I

b) 1A=A

T_9: a) $\bar{A} + A = 1$

b) $\bar{A}A = 0$

T_{10}: a) $A + \bar{A}B = A + B$

b) $A(\bar{A} + B) = AB$

T_{11}: De Morgan's theorem

Example

1) Prove that

$A + \bar{A}B = AI + \bar{A}B$

Algebraically

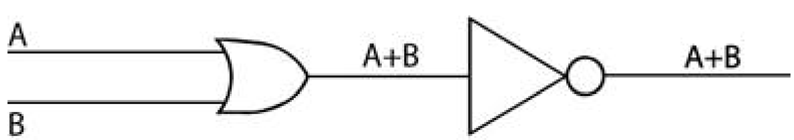
$A + \bar{A}B = AI + \bar{A}B$

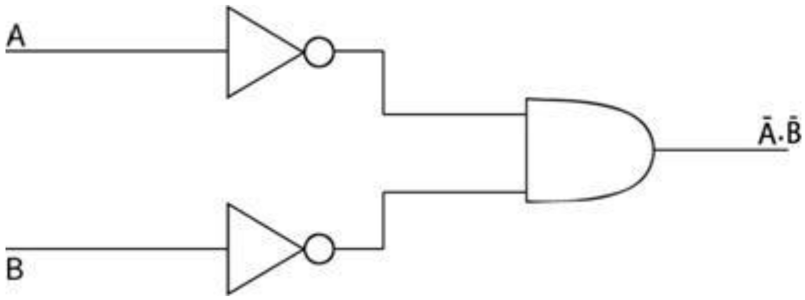
$=A\ (I+B) + \bar{A}B$

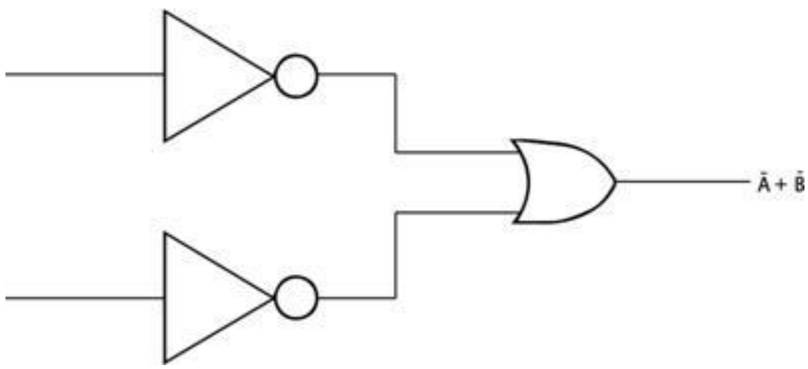
$= A(I + B) + \bar{A}B$

$= A + B(A + \bar{A})$

$= A + B$







Question

For a lift (L), these are the conditions:

i) The lift door must be closed giving d=1

ii) The appropriate floor button (B) must be pressed B

L= Bd but not L=B+ d

A boiler shut down solenoid (s) will operate if the temperature T reaches 50 and the circulating pump P ise turned off or if the pilot light L goes out.

$S = (\bar{L} + \bar{P}T)$

Or

$S = (\bar{P}T + L)$

CURRENT ELECTRICITY

An electric current is set up when a net charge Q passes through any section of the conductor in time t.

The charge assumed to be constant: i.e $I = \frac{Q}{t} \dots\dots\dots (i)$
If the value of the rate flow of charge is not constant the current varies with time and is given by the equation. $I = \frac{dQ}{dt}$.

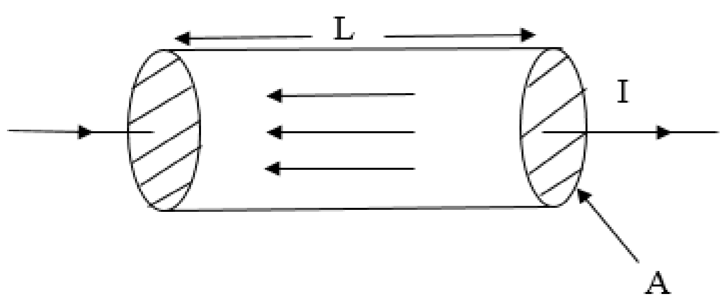
VELOCITY OF CURRENT CARRIES OR DRIFT VELOCITY (V).

The conduction of electricity in metal is due to free electrons. Free electron have the thermal energy and wonder randomly through the metal from atom to atom and hence collide.

The amplitude of vibration are then increased and the temperature of the metal rise.

On the average the electron drift in the direction of the drift in the mean velocity called drift velocity

Consider a metal wire of cross sectional area A and of length L.



Let “n” be the number of electrons per unit volume and “e” is the charge of an electron

The volume occupied by electron is AL

The total number of electrons is nAL

The total charge (Q) in the wire is nALe........ (1)

The time taken by the electron to travel the distance L is $t = \frac{L}{v} \dots\dots\dots\dots\dots\dots .(2)$

If I is the current flowing through this wire then,
I = Q/t ---------------------(3)

But $Q = nALe$

$I = \frac{nALe}{t}$

$I = \frac{nALe}{L/V}$

Therefore: $I = neVA \ldots\ldots\ldots\ldots\ldots.(4)$

Or $V = \frac{I}{neA} \ldots\ldots\ldots\ldots\ldots..(5)$

From equation 5

$$V = \frac{I}{A}.\frac{1}{ne}$$

The ratio $\frac{I}{A} = J \ldots\ldots\ldots\ldots\ldots\ldots\ldots(6)$

Where $J =$ current density. Then equation 6 becomes

$V = \frac{J}{ne} \ldots\ldots\ldots\ldots\ldots\ldots..(7)$

Examples

1. 10^{20} Electrons each having a charge passing through point x towards a point y in 0.1 sec. What is the current and its direction?

1. Solution

Given

T= 0.1

Q=$1.6x10^{-19}x20^{10}$

I.

Current charge $\underline{= 1.6x10^2 \text{A}}$

<u>The direction of current is from point Y to X</u>

The current is given by

$I = \frac{Q}{t} = \frac{1.6x10^1}{0.1}$

I = $1.6x10^2$A

2. A conductor with a cross section of $10^{-4}m^2$ and electric current of 1.2A. If the number of electrons be $5x10^{28}/m^3$ calculate the electron velocity.(Take charge of an electron=$1.6x10^1Q$

2. Solution   Formula

A= $10^{-4}m^2$

I= 1.2 A $V = \frac{I}{nAe} = \frac{1.2}{10^{-4}\ x\ \ 5\ x\ 10^{28}\ x\ 1.6\ x\ 10^{-19}}$

n= 5 x $5^{28}/m^2$ $V = \frac{1.2}{8} x\ 10^{-5} = 0.15\ x\ 10^{-5}$

e= 1.6 x 10^{-19}

Drift Velocity = $1.5\ x\ 10^{-6}\ {}^{m}/_{s}$.

Q3 A uniform copper wire of diameter 0.051mm carries a current 10A. what is the drift velocity of electrons in the wire, assuming the charge of electrons is

1.6 X 10^{-19} C

Solution   Formula

A= 3.14 x $(2.5\ x\ 10^{-4})^2$ $V = \frac{I}{nAe} = \frac{10}{1.6\ x\ 10^{-19}\ x\ 3.14\ x\ (2.5\ x\ 10^{-4})^2\ x\ 8.4\ x\ 10^{28}}$

I= 10 A

n= $8.4\ x\ 10^{28}$ V= 3.79 x $10^{-3}\ {}^{m}/_{s}$:

e= $1.6\ x\ 10^{-19}C$

The drift velocity of the electron = V= 3.79 x $10^{-3}\ {}^{m}/_{s}$:

RESISTANCE AND RESISTIVITY:

Ohm had shown that by using wires of different lengths and diameters.

Resistance of a wire R is proportional to its length

R α L

and inversely proportional to its cross section area. A i.e

$\mathrm{R}\ \alpha\ \frac{l}{A}$

From (1) and (2) we have

$\mathrm{R} = \rho\ \frac{l}{A}$

Where is the constant for the material of the wire and it is called. Resistivity.

TEMPERSTURE COEFFICIENT OF RESISTANCE ($\propto$)

The resistance of pure metal increases with temperature.

The temperature coefficient of resistance of a substance is increase in resistance of substance per degree centigrade rise of temperature per resistance at 0°C

$\propto = \frac{R_\theta - R_o}{\theta R_o}$ where $R_\theta = R_o\,(1+\propto\theta)$

If $R_1\ and\ R_2$ are resistances at θ_1 respectively

Then $R_1 = R_o\,(1+\propto\theta_1)$ and $R_2 = R_o\,(1+\propto\theta_2)$

Hence $\frac{R_1}{R_2} = \frac{R_o\,(1+\propto\theta_1)}{R_o(1+\propto\theta_2)}$

$\frac{R_1}{R_2} = \frac{1+\propto\theta_1}{1+\propto\theta_2}$

Example

A copper coil has a resistance of 50.0Ω at $20°c$ what its resistance at $80°c$?

Take the temperature coefficient of copper is $4.28 \times 10^{-3} C^{-1}$

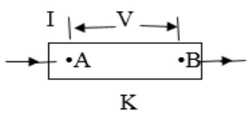
Solution

$$\frac{R_{80}}{R_{20}} = \frac{R_0\,(1 + 80\alpha)}{R_0(1 + 20\alpha)} = \frac{1 + 80\;x\;4.28x\;10^{-3}}{1 + 20\;x\;4.28x\;10^{-3}}\,x\;50$$

$$\underline{\underline{R_{80}\;= 61.8\Omega}}$$

HEAT AND ELECTRICAL POWER

Consider an electrical device k and let potential difference between A and B be V.



If Q is the charge passing through the ends in a time t then Q = It........... (1)

Total energy given out by the conductor is W = VQ ....(2)

Substitute eg (1) into eq (2) we get

W= VIt ...... (3)

but workdone = electrical heat developed (H)

hence W= H =VIt ....(4)

Now V= IR so the eqn (4) becomes

$$W = H = I^2Rt \ldots \ldots \ldots \ldots \ldots \ldots .\,(5)$$

Again I = $\frac{V}{R}$ thus eqn (5) become

$$W = H = \frac{V^2t}{R} \ldots\ldots\ldots\ (6)$$

Equation (4), (5) and (6) gives heat development

Equation (4), (5) and (6) gives

i. Power = VI
ii. Power= I^2R
iii. Power= $\frac{V^2}{R}$

KIRCHHOFF'S LAWS

Some circuits cannot be broken down into set of series and parallel combination of conductors. It is then necessary to make use of generalized form of circuits laws already considered. These laws are known as Kirchhoff's law

Kirchhoff's first law:

The algebraic sum of current at any junction in the net work is zero

i.e. $\sum I = 0$

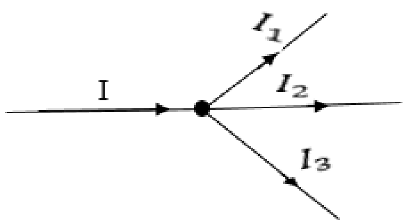


Fig .I

A current is given a positive sign if it flows towards a junction and negative sign if it flows away from the junction.

From figure I above we have

$$I + (-I_1) + (-I_2) + (-I_3) = 0 \text{ or } I - I_1 - I_2 - I_3 = 0$$

Kirchhoff's second law:

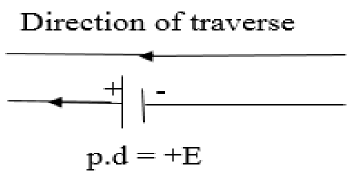
The algebraic sum of emf's around the loop is equal to the algebraic sum of all potential differences in that circuit i.e

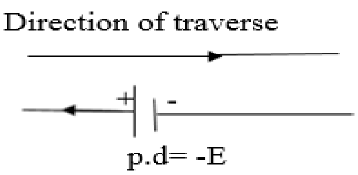
$$\sum E - \sum IR = 0 \quad \text{or } \sum E = \sum IR:$$

RULES FOR FINDING POTENTIAL DIFFERENCE(pd) IN A LOOP

If the resistance is traversed in the direction of the current,the chage in potential is -IR in the opposite direction is +IR

If a set of emf is traversed in the direction of emf the change in potential is +ve(+E) in the opposite direction it is(-E)





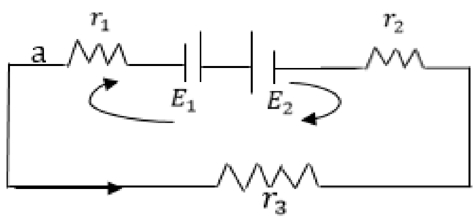
The choice of the direction of the traverse depend on yourself

Qn

In fig given below E=2.0 volts $E_2 = 4.0$ volts

r_1= 1.0Ω r_2= 2.0Ω r_3= 5Ω.

What is the current I flowing in the circuit



Solution

E_1 and E_2 are opposing each other but since $E_2 > E_1$ then E_2 must be controlling the direction of current.

Applying Kirchhoff's laws in the clockwise direction we have

$$Ir_1 + E_1 - E_2 + Ir_2 + Ir_3 = 0$$

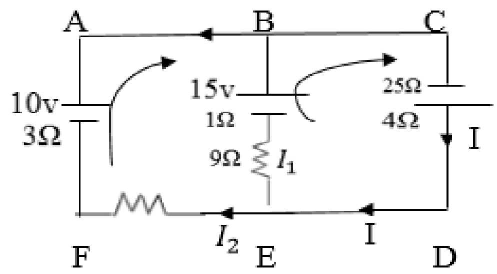
$$Ir_1 + Ir_2 + Ir_3 = E_2 - E_1$$

$$I(r_1 + r_2 + r_3) = E_2 - E_1$$

$$I = \frac{E_2 - E_1}{(r_1 + r_2 + r_3)}$$

Qn

Find the current in each resistor in the circuit shown below:



Take the loop A B E F in the clockwise direction

$-15+I_1+9I_1 - 2I_1 + 10 + (-3I_2) = 0$

$10I_1 - 5I_2 = 5 = 2I_1 - I_2 = 1 \ldots\ldots\ldots\ldots..(1)$

Take the loop BCDE in clockwise direction.

$25 - 4I - 9I_1 + 15 - I_1 = 0$

But $I = I_1 + I_2$

$4I = 4(I_1 + I_2)$

$\therefore 25 - 4I_1 - 4I_2 - 9I_1 + 15 - I_1 = 0$

$-14I_1 - 4I_2 = -40$ divide by -2 both side

$7I_1 + 2I_2 = 20 \ldots\ldots\ldots\ldots\ldots(2)$

To solve the equations simultaneously.

$7I_1 + 2I_2 = 20$

$2I_1 - I_2 = 1$

$I_1 = \frac{1+I_2}{2}$

$$7\left(\frac{1+I_2}{2}\right) + 2I_2 = 20$$

$$7 + 7I_2 + 2I_2 = 40$$

$$11I_2 = 33$$

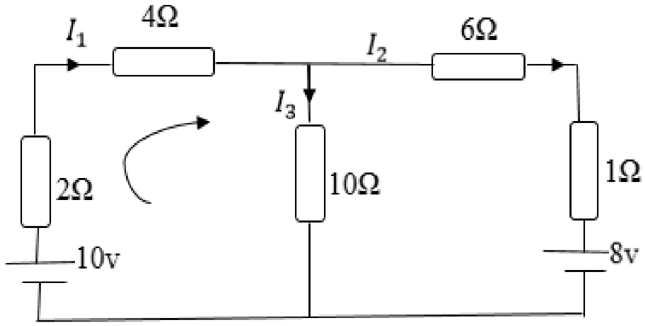
$$I_2 = \frac{33}{11} = 3$$

$$I_1 = \frac{1+3}{2} = 2$$

$$I_1 = 2A \qquad and\ I_2 = 3A$$

Qn.

Determine the magnitude of $I_1\ and\ I_3$



$$-4I_1 + 10I_3 + 10 - 2I_1 = 0$$

$$-6I_1 + 10I_3 = -10$$

$$6I_1 - 10I_3 = 10 \ldots\ldots\ldots\ldots (1)$$

$$-6I_2 - I_2 - 8 - 10I_3$$

$$-7I_2 - 10I_3 = 8 \ldots\ldots\ldots\ldots. (2)$$

But $I_2 = I_1 + I_3$

$$-7I_2 = -7(I_1 + I_3) = -7I_1 - 7I_3$$

$= -7I_1 - 7I_3 - 10I_3 = 8 \dots\dots\dots(3)$

Solve simultaneous eqn(1) and (3)

$6I_1 - 10I_3 = 10$

$-7I_1 - 17I_3 = 8$

$I_1 = \left(\frac{10 + 10I_3}{6}\right)$

$-7\left(\frac{10 + 10I_3}{6}\right) - 17I_3 = 8$

$-70 - 70I_3 - 102I_3 = 48$

$-172I_3 = 118$

$I_3 = -\frac{118}{172} = 0.686A$

$I_1 = \frac{10 + 10(0.686)}{6} = 0.523$

$I_2 = I_1 + I_3$

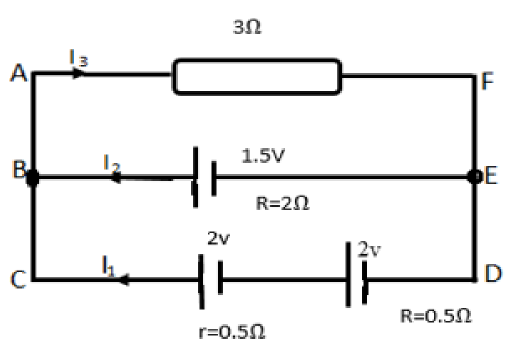
$= 0.523 + 0.686$

$= 0.163$

The minus sign tells that the current is in of the site direction (from the one assumed)

Qn.
On the basis of Kirchhoff's laws determine the numerical values for the current

I_1, I_2 and I_3



From the loop A, B, E, F the traverse in clockwise direction

$$-3I_3 + 1.5v - 2I_2 = 0$$

$$I_3 = \frac{-1.5}{-3} = 0.5.$$

From the loop B E C D the traverse in clockwise direction

$$-15 - 2I_2 + 2 - 0.5I_2 + 2 - 0.5I_1 = 0$$

$$-2.5I_2 - 0.5I_1 = -2.5 \ldots \ldots \ldots \ldots (3)$$

But $I_1 + I_2 = I_3$.

$$\therefore I_1 + I_2 = 0.5$$

$$I_1 = 0.5 - I_2$$

$$-2.5I_2 - 0.5[0.5 - I_2] = -2.5$$

$$-2.5I_2 - 0.25 + 0.5I_2 = -2.5$$

$$-2I_2 = \frac{-2.25}{-2}$$

$$I_2 = 1.125A$$

$$I_1 = 0.5 - 1.125$$

$$= -0.625A$$

$$\therefore\ I_1 = -0.625 \qquad I_2 = 1.125 \qquad I_3 = 0.5$$

Take the loop B E D C in clockwise direction.

$$2I_2 - 1.5 - 0.5I_1 + 2 + 2 - 0.5I_1 = 0$$

$$2I_2 - I_1 = -2.5 \ldots\ldots\ldots\ldots\ldots.(2)$$

But $I_3 = I_1 + I_2$.

$$-3[I_1 + I_2] - 2I_2 = -1.5$$

$$= -3I_1 - 5I_2 = -1.5 \ldots\ldots\ldots\ldots\ldots.(3)$$

To solve eqn (2) and (3)

$$2I_2 - I_1 = -2.5$$

$$5I_2 - 3I_1 = 1.5$$

$$\therefore I_2 = \frac{-2.5 + I_1}{2}$$

$$5\left(\frac{-2.5+I_1}{2}\right) - 3I_1 = -1.5$$

$$-12.5 + 5I_1 - 6I_1 = -2.5$$

$$I_1 = \frac{-9.5}{11}$$

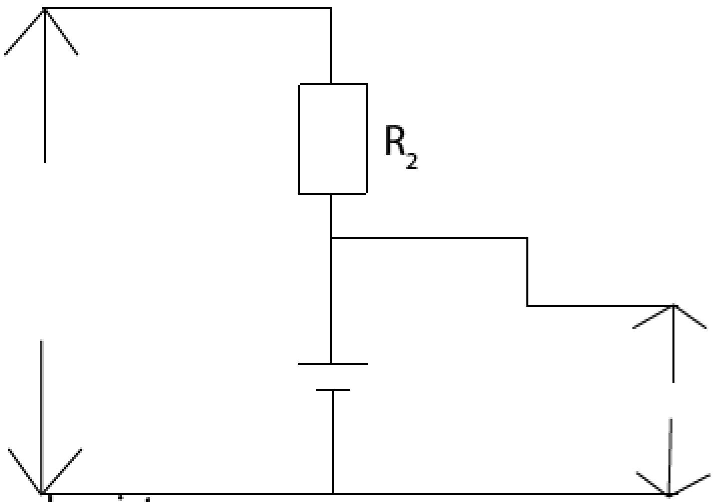
$$I_2 = -2.5 + \frac{18}{11}$$

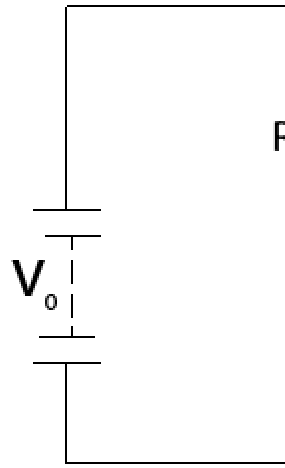
$$= \frac{91}{22}A$$

$$I_2 = \frac{91}{22}\text{A}$$

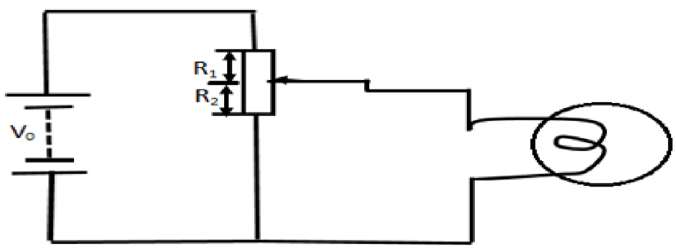
POTENTIAL DIVIDER

The resistance box in series are often used to provide known fraction of a given potential difference





Fixed resistor



(iii) With load.

The current I flowing is given by

$$I = \frac{V_0}{R_1 + R_2} \dots \dots \dots \dots \dots \dots (1)$$

$$V_1 = IR_1 \dots \dots \dots \dots \dots \dots (2)$$

$$V_1 = IR_1 \dots \dots \dots \dots \dots \dots (2)$$

Substitute equation (1) in (2) we get

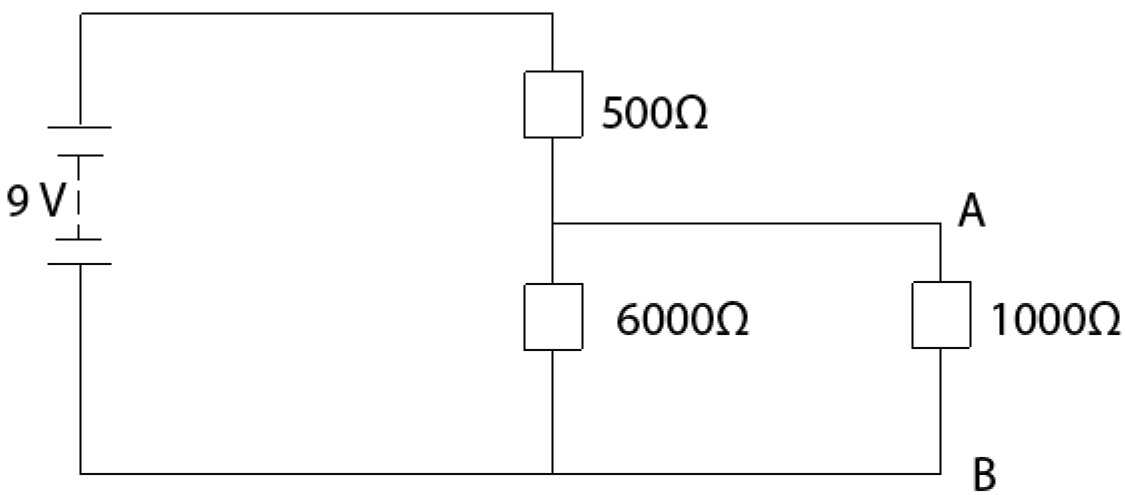
$$V_1 = \frac{R_1}{R_1 + R_2} V_0 \dots \dots \dots \dots \dots \dots (3)$$

A resistor with a sliding contact can similarly be used as shown in figure (ii) to provide continuous variable potential difference from zero to a full supply by value $V_{0.}$

This is a convenient way of controlling the voltage applied to a load such as lump. The resistor of the load R_3 however acts in parallel with resistor R_1 hence equation (1) is no longer true.

Qn.

What is the final potential difference between A and B in the direction circuit below (I) in the circuit as shown (II) if an additional 5000Ω resistor where connected from A to B



Solution

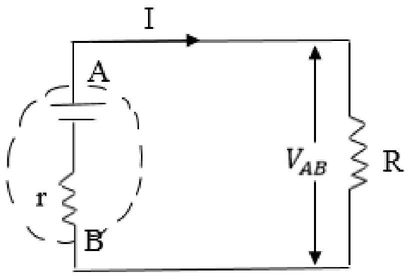
(i) Final p.d between AB= $\frac{1000}{6000} x\ 9 = 1.5V$

(ii) The $P.d$ between 500Ω

$$V = \frac{500\times 9\times((6000)+500)}{6000\times 500} = \frac{58500}{6000}$$

$$V = 9.75\ volts$$

OHM'S LAW FOR COMPLETE CIRCUIT



There exist certain device such as battery and electronic generators which are able to maintain a potential balance between two points to which they are attached to devices are called a set of electromotive force put the normally have some internal resistances. When the current flow through the battery setsup a p.d **Ir** across the internal resistance. The resistor of resistance R which is connected to the battery is called the load.

OUT PUT AND AN EFFICIENCY

V_{AB} = E $-I_r$ .... (1)

It is called the voltage drop.

- Power generated by the source EI ... ... (2)
- Heat developed per second = $I^2R + I^2r$ ... ... (3)

But power generated = Heat developed per second

$= EI = I^2(R + r) = E = I(R + r)$

Or I=$\frac{E}{R+r}$ .... (4)

Equation (4) is called ohms law for complete circuit

But from eqn (1) we have

V_{AB}=IR+Ir−Ir

V_{AB}= IR

$V_{AB} = \frac{R}{R+r}E$

The power developed for the load is called the out put power (P_{out})

P_{out}= IV_{AB} ... ... (5)

power generated by the source the current

Efficiency $(\eta) = \frac{P_{out}}{P_{gen}}$

$\eta = \frac{IV_{AB}}{EI} = \frac{V_{AB}}{E}$ .... (7) put * **into (7) we get**

$\eta = \frac{R}{R+r} \cdot \frac{E}{E}$

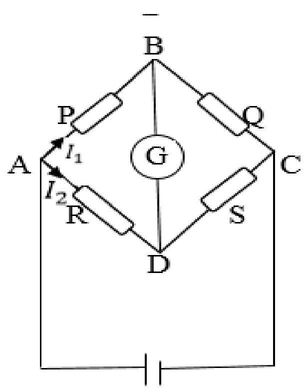
$\boxed{\eta = \frac{R}{R+r}}$ i.e As R. becomes very large r will be negligible and so

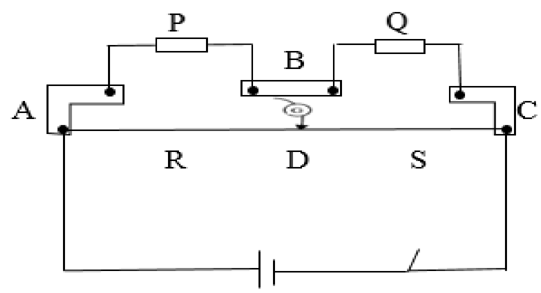
$\eta = \frac{R}{R} = 1$ or 100% efficiency

$P_{gen} = EI \ldots\ldots\ldots\ldots (6)$

WHEATSTONE BRIDGE:

- Wheat stone bridge is a circuit that enable resistance to be measured accurately:





Four resistors P, Q , R and S are joined as shown in the figure above

If P is the known resistor Q must be known as well as R and S or their ratio one or more of Q, R and S are adjusted until there is no deflection on G, the bridge is then said to be balanced.

At balance point no current through G so p.d across BD=0 and p.d across AB=p.d across AD. Also current through R = current through $S_1 = I_2$. Current through p=current through Q= I_1

Therefore $I_1 P = I_2 R_1 \ldots \ldots (1)$ similarly IQ $= I_2 S \ldots \ldots \ldots \ldots (2)$

Dividing (1) by (2) gives

$$\frac{IQ = I_2 S}{I_1 P = I_2 R_2}$$

$$\frac{P}{Q} = \frac{R}{S}$$

Now since R α C then we have

P α AD and Q α DC so

$$\frac{P}{Q} = \frac{AD}{DC}$$

Qn.

A wire of length 1.1 meter and Radius $7x10^{-5}m$ is connected across the right gap of the metre bridge when a resistance of 45 Ω is introduced in the left gap from a resistance box connected across it. The balance point is obtained 0.6m from left side. Calculate the specific resistance (resistivity) of the material of the wire

Solution

The value of resistance

$$\frac{45}{0.6} = \frac{x}{0.4}$$

$$X = \frac{45 \, x \, 0.4}{0.6} = 30\Omega$$

Formula of resistivity

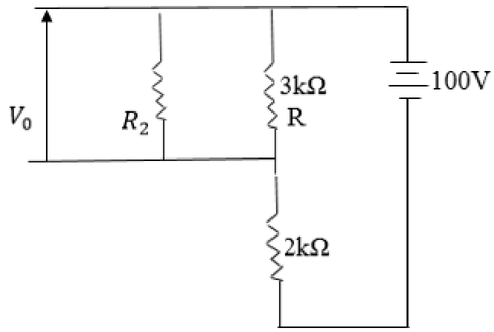
$$\rho = \frac{RA}{L}$$

$$\therefore = \frac{30 \, x \, 3.14 \, x(7.0 \, x \, 10^{-5})^2}{1.1}$$

$$\rho = 4.2x \, 10^{-7}\Omega m$$

Qn.

A $2\text{k}\Omega$ and $3\text{k}\Omega$ resistor are connected in series combination is connected across a 100v supply of negligible internal resistance as shown in the figure below:



(i) What is the output voltage V_{out} of the load resistance R_1 =30kΩ

(ii) What current resistance through the load resistor R_2 in (i) above.

I=21.15.

Solution

The resistance $= 2k\Omega + \frac{6\times30}{33} = 52$

Current $= \frac{V}{R} = \frac{100\times11}{52} = 21.15$

Voltage $V = \frac{30}{11} \times 21.5$

$= 57.6V$