

Memory Based

Maths Memory Based Questions

Question 1. General solution of differential equation

Ans. The general solution of a differential equation refers to the set of all possible solutions that satisfy the equation. The specific form of the general solution depends on the type and order of the differential equation.

Different types of differential equations have different methods for finding their general solutions. Here are a few common types of differential equations and their general solutions:

Ordinary Differential Equations (ODEs):

First-Order Linear ODE: $y = C \cdot e^{-\int P(x)dx} + \int [Q(x)e^{-\int P(x)dx}]dx$

Second-Order Homogeneous ODE: $y = C_1e^{r_1x} + C_2e^{r_2x}$

Second-Order Non-Homogeneous ODE (with constant coefficients): $y = y_h + y_p$, where y_h is the general solution of the associated homogeneous equation and y_p is a particular solution.

Partial Differential Equations (PDEs):

Separation of Variables: The general solution is found by separating variables and solving each resulting ordinary differential equation independently. Fourier Series: The general solution is represented as an infinite series involving trigonometric functions (sine and cosine) or other orthogonal functions. It's important to note that the general solution may contain arbitrary constants (C_1, C_2 , etc.) that are determined by applying initial or boundary conditions to obtain a particular solution that satisfies specific constraints.

The specific method for finding the general solution of a differential equation depends on the equation's type, order, and particular characteristics. More advanced techniques, such as Laplace transforms, Fourier transforms, or numerical methods, may be required for more complex differential equations.

To obtain the general solution for a specific differential equation, it is necessary to determine the equation's type and apply the appropriate methods and techniques accordingly.

Question 2. Dist. of point from line

Ans. To find the distance between a point and a line in a two-dimensional space, you can use the formula for the distance between a point and a line.

Consider a line defined by the equation $Ax + By + C = 0$, where A, B , and C are constants representing the coefficients of the line equation. Let (x_0, y_0) be the coordinates of the point.

The formula to calculate the distance between the point and the line is given by:

$$d = |Ax_0 + By_0 + C| / \sqrt{A^2 + B^2}$$

where $|Ax_0 + By_0 + C|$ denotes the absolute value of the expression $Ax_0 + By_0 + C$.

By substituting the appropriate values for A , B , C , x_0 , and y_0 , you can calculate the distance (d) between the point and the line.

It's worth noting that this formula assumes a straight line and Euclidean distance in a two-dimensional Cartesian coordinate system. If you are working in a different coordinate system or considering a non-linear curve, the calculation method may vary.

Question 3. Lpp area of solution region

Question 4. Fog function

Ans. The fog function, also known as a composite function, is a mathematical operation that combines two functions together. It is denoted by " $f \circ g$ " and is read as "f composed with g."

Given two functions, $f(x)$ and $g(x)$, the fog function is defined as:

$$(f \circ g)(x) = f(g(x))$$

In other words, to evaluate the fog function at a specific value of x , you first apply the function $g(x)$ to x , and then apply the resulting value to the function $f(x)$.

The order of the composition is important. The function $g(x)$ is applied first, followed by the function $f(x)$. The output of $g(x)$ serves as the input to $f(x)$.

The fog function allows you to combine functions to create more complex functions and perform calculations in a step-by-step manner. By applying one function to the output of another, you can create a chain of operations.

Question 5. Definite integration with Wall's formula

Ans. I apologize, but I am not familiar with a specific mathematical formula called "Wall's formula" related to definite integration. It's possible that you may be referring to a formula or method that I am not aware of or that is not widely recognized by that name.

Definite integration typically involves finding the area under a curve between two given limits. It is often denoted by the integral symbol (\int) and has the form:

$$\int [a, b] f(x) dx$$

where " a " and " b " represent the lower and upper limits of integration, " $f(x)$ " represents the integrand (the function being integrated), and " dx " represents the differential element.

If you can provide more specific details or clarify the context or formula you are referring to as "Wall's formula," I would be happy to assist you further.

Question 5. Conditional probability with given $P(A)$ and $P(A \cup B)$ and asked to find $P(B)$ two questions of this type

Ans. If we are given the probability of event A, $P(A)$, and the probability of the union of events A and B, $P(A \cup B)$, and we are asked to find the probability of event B, $P(B)$, we can use the formula for conditional probability.

The conditional probability of event B given event A is denoted as $P(B|A)$ and can be calculated using the following formula:

$$P(B|A) = P(A \cap B) / P(A)$$

Where $P(A \cap B)$ represents the probability of the intersection of events A and B.

Now, let's rearrange the formula to solve for $P(B)$

$$P(B|A) * P(A) = P(A \cap B)$$

$$P(B) = P(A \cap B) / P(A|B)$$

Given that $P(A)$ and $P(A \cup B)$ are provided, we can use these probabilities to calculate $P(B)$ as follows:

$$P(A \cap B) = P(A \cup B) - P(A \cup B')$$
 (Using the formula for the probability of the intersection of two events)

$P(A|B) = P(A \cap B) / P(B)$ (Using the formula for conditional probability) Simplifying, we can express $P(B)$ in terms of the given probabilities:

$$P(B) = P(A \cap B) / P(A|B) = (P(A \cup B) - P(A \cup B')) / P(A|B)$$

Therefore, to find $P(B)$, we need to know the conditional probability $P(A|B)$ or additional information about the relationship between events A and B.

Question 6. $(b-a) / 2$ thing of Definite integrals (integrate from a to b)

Ans. The expression $(b - a) / 2$ represents half the difference between two values, b and a. In the context of definite integrals, the expression $(b - a) / 2$ often arises when using certain numerical integration methods, such as the midpoint rule.

The midpoint rule is a simple numerical approximation method for evaluating definite integrals. It divides the interval $[a, b]$ into subintervals and approximates the area under the curve by treating each subinterval as a rectangle with a height determined by the function's value at the midpoint of the subinterval.

To apply the midpoint rule, you first divide the interval $[a, b]$ into n equal subintervals. The width of each subinterval is given by $\Delta x = (b - a) / n$. Then, you evaluate the function at the midpoint of each subinterval and multiply it by the width of the subinterval. Finally, you sum up all these approximations to get an estimate of the definite integral.

Mathematically, the approximation of the definite integral using the midpoint rule can be written as:

$$\int [a, b] f(x) dx \approx \Delta x * (f(a + \Delta x/2) + f(a + 3\Delta x/2) + f(a + 5\Delta x/2) + \dots + f(b - \Delta x/2))$$

where $\Delta x = (b - a) / n$ and n is the number of subintervals.

So, $(b - a) / 2$ appears as half the width of each subinterval in the midpoint rule, as it calculates the midpoint for each subinterval.

Question 7. INTEGRATION SIN⁴X.COS²X INTEGRATION LIMIT $\pi/2$

Ans. To evaluate the definite integral of $\sin^4(x) * \cos(2x)$ with integration limits from $\pi/2$, we can use the power-reducing identities for sine and double angle identity for cosine. Let's break down the steps:

Apply the power-reducing identity for sine:

$$\sin^4(x) = (1/2)(1 - \cos(2x))^2$$

Apply the double angle identity for cosine:

$$\cos(2x) = 2\cos^2(x) - 1$$

Now we have:

$$\int(\pi/2) \sin^4(x) * \cos(2x) dx$$

$$= \int(\pi/2) (1/2)(1 - \cos(2x))^2 * (2\cos^2(x) - 1) dx$$

Expanding and simplifying:

$$= (1/2) * \int(\pi/2) (1 - 2\cos(2x) + \cos^2(2x)) * (2\cos^2(x) - 1) dx$$

Distribute and simplify further:

$$= (1/2) * \int(\pi/2) (2\cos^2(x) - 4\cos^3(2x) + 2\cos^4(2x) - 2\cos^2(x) + \cos^2(2x)) dx$$

Combining like terms:

$$= (1/2) * \int(\pi/2) (\cos^2(x) - 4\cos^3(2x) + 2\cos^4(2x) + \cos^2(2x)) dx$$

Now we have separate terms that we can integrate individually:

Term 1: $\int(\pi/2) \cos^2(x) dx$

Term 2: $-4 * \int(\pi/2) \cos^3(2x) dx$

Term 3: $2 * \int(\pi/2) \cos^4(2x) dx$

Term 4: $\int(\pi/2) \cos^2(2x) dx$

Each of these terms can be integrated using basic integration techniques. Once integrated, you can substitute the upper limit $\pi/2$ and subtract the result with the lower limit to obtain the definite integral.

Question 8. Binomial standard formula $C_0 + C_1 + \dots + C_n$

Ans. The binomial standard formula, also known as the binomial theorem, expresses the expansion of a binomial raised to a positive integer power. It can be stated as follows:

$$(x + y)^n = C(n, 0) * x^n * y^0 + C(n, 1) * x^{(n-1)} * y^1 + C(n, 2) * x^{(n-2)} * y^2 + \dots + C(n, n) * x^0 * y^n$$

In this formula, $(x + y)$ represents a binomial expression with x and y as variables, n is a positive integer exponent, and $C(n, k)$ represents the binomial coefficient. The binomial coefficient $C(n, k)$ represents the number of ways to choose k items from a set of n items and is given by the formula:

$$C(n, k) = n! / (k! * (n-k)!)$$

Now, let's rewrite the formula using the notation you provided:

$$C_0 + C_1 + \dots + C_n$$

This notation represents the sum of binomial coefficients from C_0 to C_n . It can be written using the binomial standard formula as follows:

$$C_0 + C_1 + \dots + C_n = (x + y)^n$$

In other words, the sum of all binomial coefficients from C_0 to C_n is equal to the binomial expansion of $(x + y)^n$.

Question 9. how many cases are possible if a batsman can hit 0, 1,2,3,4,6 on 6 balls to make total score 30

Ans. To determine the number of cases in which a batsman can score a total of 30 runs by hitting 0, 1, 2, 3, 4, or 6 on each of the 6 balls, we can use combinatorics. We need to find the number of combinations of scores that sum up to 30.

Let's break it down:

We have 6 balls to consider, and each ball can have 6 possible scores (0, 1, 2, 3, 4, or 6).

We need to find the number of combinations of these scores that sum up to 30.

We can solve this problem using a technique called "stars and bars" or "balls and urns."

Consider the 6 balls as 6 "stars." We need to distribute these stars among the scores (0, 1, 2, 3, 4, and 6) to get a total of 30 runs. To represent the distribution, we can use "bars" or "urns" between the scores.

For example, if we have the distribution: 6 | 4 | 6 | 6 | 4 | 4, it means the batsman hit a 6 on the first ball, a 4 on the second ball, a 6 on the third ball, a 6 on the fourth ball, a 4 on the fifth ball, and a 4 on the sixth ball.

To find the number of cases, we need to find the number of ways to arrange the 6 stars and 5 bars. The bars divide the stars into different scores. We can calculate this using the concept of combinations.

The number of cases is given by:

$$\text{Number of cases} = C(\text{stars} + \text{bars}, \text{bars})$$

In this case, we have 6 stars and 5 bars, so:

$$\text{Number of cases} = C(6 + 5, 5) = C(11, 5)$$

Using the binomial coefficient formula:

$$C(n, k) = n! / (k! * (n-k)!)$$

we can calculate the number of cases as:

$$C(11, 5) = 11! / (5! * (11-5)!) = 11! / (5! * 6!)$$

Calculating this expression:

Number of cases = 462

Therefore, there are 462 possible cases in which a batsman can score a total of 30 runs by hitting 0, 1, 2, 3, 4, or 6 on each of the 6 balls.

Question 10. $3\mathbf{j} + 5\mathbf{z}$, $[\mathbf{1}-4]\mathbf{+z}$, $2\mathbf{i} + 5\mathbf{j} - k\mathbf{z}$ they are collinear find k

Ans. To determine if the vectors $3\mathbf{j} + 5\mathbf{z}$, $[\mathbf{1}-4]\mathbf{+z}$, and $2\mathbf{i} + 5\mathbf{j} - k\mathbf{z}$ are collinear, we need to check if they are scalar multiples of each other. This means that one vector can be obtained by multiplying another vector by a scalar constant.

Let's compare the components of the given vectors:

Vector 1: $3\mathbf{j} + 5\mathbf{z}$

Vector 2: $[\mathbf{1}-4]\mathbf{+z}$

Vector 3: $2\mathbf{i} + 5\mathbf{j} - k\mathbf{z}$

Comparing the x-components, we have 0 in Vector 1 and 1 in Vector 3. They are not equal, so the vectors cannot be scalar multiples of each other.

Therefore, the vectors $3\mathbf{j} + 5\mathbf{z}$, $[\mathbf{1}-4]\mathbf{+z}$, and $2\mathbf{i} + 5\mathbf{j} - k\mathbf{z}$ are not collinear.

Since the vectors are not collinear, we do not need to find the value of k.

Chemistry Memory Based Questions

Question 1. pH order of 0.1M OF sodium acetate, NH_4Cl and NaCl

Ans. To determine the pH order of three solutions: 0.1 M sodium acetate, NH_4Cl , and NaCl , we need to consider the nature of the dissolved compounds and their effect on the pH of the solution.

Sodium Acetate (CH_3COONa):

Sodium acetate is a salt of a weak acid (acetic acid) and a strong base (sodium hydroxide). When sodium acetate dissolves in water, it undergoes hydrolysis, which results in the formation of acetate ions (CH_3COO^-) and a small amount of hydroxide ions (OH^-). Acetate ions being the conjugate base of a weak acid, are relatively basic.

As a result, the presence of sodium acetate in water will slightly increase the pH of the solution. Therefore, the pH of the 0.1 M sodium acetate solution will be slightly basic, but the extent of the increase will depend on the concentration.

NH_4Cl :

NH_4Cl is a salt of a weak base (ammonia) and a strong acid (hydrochloric acid).

When NH_4Cl dissolves in water, it dissociates into ammonium ions (NH_4^+) and chloride ions (Cl^-). Ammonium ions are the conjugate acid of a weak base and are relatively acidic. The presence of NH_4Cl in water will slightly decrease the pH of the solution.

Therefore, the pH of the NH_4Cl solution will be slightly acidic, but again, the extent of the decrease will depend on the concentration.

NaCl:

NaCl is a salt of a strong acid (hydrochloric acid) and a strong base (sodium hydroxide). When NaCl dissolves in water, it dissociates into sodium ions (Na^+) and chloride ions (Cl^-). Neither sodium nor chloride ions significantly affect the pH of the solution.

Therefore, the pH of a NaCl solution will be neutral, around 7.

In summary, the pH order of the given solutions from most acidic to most basic would be:
 $\text{NH}_4\text{Cl} < \text{NaCl} < 0.1 \text{ M sodium acetate}$.

Question 2. which has least basic conjugate base options hcl,hbr,oxalic acid and acetic acid

Ans. To determine which compound has the least basic conjugate base among HCl, HBr, oxalic acid, and acetic acid, we need to consider the acidity of their respective conjugate bases.

The conjugate base of an acid is formed when the acid donates a proton (H^+). The stronger the acid, the weaker its conjugate base.

Let's analyze each compound:

HCl (hydrochloric acid):

The conjugate base of HCl is chloride ion (Cl^-). Chloride ion is very weakly basic.

HBr (hydrobromic acid):

The conjugate base of HBr is bromide ion (Br^-). Bromide ion is also very weakly basic.

Oxalic acid:

Oxalic acid has two acidic protons, and its conjugate base is oxalate ion ($\text{C}_2\text{O}_4^{2-}$). Oxalate ion is a weak base but stronger than chloride or bromide ions.

Acetic acid:

Acetic acid has a weak acidity, and its conjugate base is acetate ion (CH_3COO^-). Acetate ion is a relatively weak base but stronger than chloride, bromide, or oxalate ions.

Therefore, among the given compounds, oxalic acid has the least basic conjugate base.

In summary, the order of least basic conjugate bases is: $\text{HCl} < \text{HBr} < \text{acetic acid} < \text{oxalic acid}$

Question 3. molecular orbital config of N_2

Ans. The molecular orbital configuration of N_2 (nitrogen gas) can be determined by combining the atomic orbitals of two nitrogen atoms. Nitrogen has atomic number 7, so each nitrogen atom has 7 electrons.

The molecular orbital diagram for N_2 can be represented as follows:

σ_{2s} (bonding)

σ_{2s}^* (anti-bonding)

σ_{2p} (bonding)

σ_{2p}^* (anti-bonding)
 π_{2p} (bonding)
 π_{2p}^* (anti-bonding)

In this diagram, the σ_{2s} and σ_{2p} orbitals are lower in energy and bonding orbitals, while the σ_{2s}^* and σ_{2p}^* are higher in energy and anti-bonding orbitals. The π_{2p} and π_{2p}^* orbitals are degenerate and form bonding and anti-bonding orbitals.

For N_2 , we need to fill these molecular orbitals with a total of 14 electrons (7 electrons from each nitrogen atom). We start by filling the orbitals from lowest to highest energy, following the Pauli exclusion principle and Hund's rule.

The electron configuration for N_2 is as follows:

σ_{2s} : $\uparrow\downarrow$
 σ_{2s}^* : (empty)
 σ_{2p} : $\uparrow\downarrow$
 σ_{2p}^* : (empty)
 π_{2p} : $\uparrow\uparrow\downarrow\downarrow$
 π_{2p}^* : (empty)

The molecular orbital configuration shows that N_2 has a triple bond. The π_{2p} bonding orbitals are fully occupied with 4 electrons, resulting in a strong covalent bond between the nitrogen atoms. This triple bond is responsible for the stability and strength of the nitrogen molecule.

It's important to note that the diagram and configuration provided here are simplified representations of the molecular orbitals in N_2 . The actual molecular orbital energies and electron distributions can be more complex, but this serves as a basic understanding of the molecular orbital configuration of N_2 .

Question 4. is NaOH a standard titrant

Ans. Yes, NaOH (sodium hydroxide) is commonly used as a standard titrant in acid-base titrations. In acid-base titrations, a solution of known concentration (the titrant) is used to determine the concentration of an unknown solution (the analyte) by measuring the volume of titrant required to reach a specific endpoint.

NaOH is considered a standard titrant because it can be prepared with high accuracy and precision, and its concentration can be determined through primary standardization methods. A primary standard is a highly pure compound that can be accurately weighed and dissolved to prepare a solution of known concentration.

In the case of NaOH, its concentration can be determined by titration against a standardized acid solution (such as hydrochloric acid) using a suitable indicator or pH meter to determine the endpoint of the reaction.

Once the concentration of the NaOH solution is accurately determined, it can be used as a standard titrant in various acid-base titrations to determine the concentration of acidic analytes.

Question 5. miller indices (solid state)

Ans. Miller indices are a notation system used in solid-state physics and crystallography to describe

the orientation and planes of crystal lattice structures. They are denoted by three integers (hkl), known as Miller indices, which represent the reciprocals of the fractional intercepts made by a plane on the crystallographic axes.

The steps to determine the Miller indices of a plane in a crystal lattice are as follows:

Identify the intercepts: Determine the points at which the plane intersects the crystallographic axes (x, y, z).

Take the reciprocals: Take the reciprocals of the fractional intercepts to obtain integers. If there is no intercept along a particular axis, assign a value of infinity (∞) or zero (0).

Simplify the indices: If the indices have any common factors, divide all of them by the greatest common factor to simplify the Miller indices. **Enclose the indices in parentheses:** Write the simplified indices as (hkl), where h, k, and l represent the integers obtained in steps 2 and 3.

For example, if a plane intersects the x-axis at 1/2, the y-axis at 1/3, and does not intersect the z-axis, the Miller indices would be:

$h = 2$ (reciprocal of 1/2)

$k = 3$ (reciprocal of 1/3)

$l = 0$ (no intercept along the z-axis)

After simplifying, the Miller indices would be written as (2 3 0).

Miller indices are used to describe crystallographic planes, directions, and other lattice features in solid-state materials. They provide a compact and standardized way of representing the orientation and symmetry of crystal structures.

Question 6. when calcium is titrated with EDTA what is the pH of buffer

Ans. The pH of the buffer used in the titration of calcium with EDTA (ethylenediaminetetraacetic acid) depends on the specific composition of the buffer solution.

In general, when titrating calcium with EDTA, a common choice for the buffer is ammonia-ammonium chloride buffer, commonly known as "NH₄Cl/NH₃ buffer" or "ammonium buffer." This buffer system consists of a mixture of ammonium chloride (NH₄Cl) and ammonium hydroxide (NH₃) in water.

The pH of the ammonium buffer can be adjusted by controlling the relative concentrations of NH₄Cl and NH₃. The buffer capacity is typically achieved around pH 9-10, which is suitable for complexation reactions involving calcium and EDTA.

The actual pH of the buffer solution used in the titration would depend on the specific concentrations of NH₄Cl and NH₃, as well as the temperature and ionic strength of the solution. Therefore, to determine the precise pH value of the buffer, the specific composition and concentrations of the buffer components need to be known or specified.

It's important to note that maintaining a consistent and appropriate pH during the titration is crucial for accurate results, as the complex formation between calcium and EDTA is pH-dependent. Therefore, careful consideration and adjustment of the buffer pH are necessary to ensure the success of the calcium titration with EDTA.

Physics Memory Based Questions

Question 1. If a ring and disc are allowed to come down through inclined plane, then which will come down first

Ans. The ring and the disc will have different rates of descent when allowed to roll down an inclined plane due to their different moments of inertia and distribution of mass.

When a ring rolls down an inclined plane, its moment of inertia is higher compared to a disc of the same mass and radius. This is because the mass of the ring is distributed farther from its rotational axis, resulting in a larger moment of inertia. Due to its larger moment of inertia, the ring tends to resist rotational motion more than the disc.

As a result, the disc will come down the inclined plane first, followed by the ring. The disc's lower moment of inertia allows it to gain rotational motion and roll down the incline more easily than the ring.

Therefore, the disc will descend first, while the ring will take longer to roll down the inclined plane.

Question 2. Ek work nikalna tha f diya tha vector form mai s diya tha ek point sai duse point tak

Ans. To calculate the work done (W) given a force vector F and displacement vector s, you can use the dot product between the force vector and the displacement vector. The formula for calculating work in vector form is:

$$W = \mathbf{F} \cdot \mathbf{s}$$

where \cdot represents the dot product.

The dot product of two vectors is calculated by multiplying the corresponding components of the vectors and taking their sum. If $\mathbf{F} = (F_x, F_y, F_z)$ and $\mathbf{s} = (s_x, s_y, s_z)$, then the dot product $\mathbf{F} \cdot \mathbf{s}$ is given by:

$$\mathbf{F} \cdot \mathbf{s} = F_x * s_x + F_y * s_y + F_z * s_z$$

So, to calculate the work done, you need the components of the force vector F and the displacement vector s. Multiply the corresponding components and sum them up to obtain the scalar value of the work done.

Question 3. Field due to infinite sheet(gave 2 sheets with same charge density and based net field outside the region common

Ans. When considering two infinite sheets with the same charge density and opposite charges, the net electric field outside the region common to both sheets will be zero.

This can be understood by considering the symmetry of the system. Since the charge densities on both sheets are the same but have opposite signs, the electric fields produced by each sheet will cancel each other out in the region outside where they overlap.

However, it's important to note that the electric field due to each individual sheet will still exist and

contribute to the net field within their respective regions. The cancellation of the electric fields only occurs in the overlapping region outside both sheets.

Inside the region common to both sheets, the net electric field will depend on the distance from each sheet and their respective charge densities. To calculate the exact electric field within that region, the geometry and specific values of the charge densities would need to be known.

Question 4. Gave ratio of volume submerged in 2 diff liquids and they asked ratio of densities of liquid,

Ans. If the ratio of volumes submerged in two different liquids is given, we can use Archimedes' principle to determine the ratio of their densities.

According to Archimedes' principle, the buoyant force experienced by an object submerged in a fluid is equal to the weight of the fluid displaced by the object. The buoyant force can be expressed as:

$$F_{\text{buoyant}} = \rho_{\text{fluid}} * V_{\text{submerged}} * g$$

where ρ_{fluid} is the density of the fluid, $V_{\text{submerged}}$ is the volume of the object submerged in the fluid, and g is the acceleration due to gravity.

If we have two different liquids, let's say Liquid 1 and Liquid 2, and their respective volumes submerged are $V_{\text{submerged}_1}$ and $V_{\text{submerged}_2}$, we can set up the following equation:

$$\rho_1 * V_{\text{submerged}_1} * g = \rho_2 * V_{\text{submerged}_2} * g$$

Canceling out the common factors, we find:

$$\rho_1 / \rho_2 = V_{\text{submerged}_2} / V_{\text{submerged}_1}$$

Therefore, the ratio of the densities of the two liquids (ρ_1 / ρ_2) is equal to the ratio of the volumes submerged ($V_{\text{submerged}_2} / V_{\text{submerged}_1}$).

By knowing the ratio of volumes submerged, you can directly determine the ratio of densities of the two liquids.

Question 5. Block is thrown on inclined plane with friction and asked if velocity with which it will come back be treated or less

Ans. When a block is thrown on an inclined plane with friction, the velocity with which it will come back (assuming it returns along the same path) will be less than the velocity at which it was thrown.

Several factors contribute to this decrease in velocity:

Friction: The presence of friction between the block and the inclined plane will oppose its motion. Friction converts some of the kinetic energy into heat energy, causing a loss of mechanical energy. Consequently, the block will lose speed as it moves up the incline.

Work done against gravity: As the block moves up the incline, it works against gravity. This work done by the block results in a loss of kinetic energy and a decrease in velocity.

Losses due to air resistance: In real-world scenarios, air resistance can also play a role in reducing the velocity of the block. Air resistance acts opposite to the direction of motion and can further diminish the velocity of the block as it moves against it.

Overall, due to the combined effects of friction, work done against gravity, and air resistance (if applicable), the velocity with which the block will come back up the inclined plane will be less than the velocity at which it was thrown.

Question 6. work done by tension in circular motion

Ans. In circular motion, the tension force does work on an object moving in a circular path. The work done by tension can be determined by considering the change in kinetic energy of the object.

When an object is moving in a circular path at a constant speed, its velocity is constantly changing direction. This change in direction requires a force to act towards the center of the circular path, which is provided by the tension force in this case.

Since the tension force is perpendicular to the displacement of the object, the work done by tension is zero. This is because the dot product of the tension force and displacement vector is zero when the angle between them is 90 degrees. Mathematically, the work done by tension (W_{tension}) can be expressed as:

$$W_{\text{tension}} = T * s * \cos\theta$$

where T is the tension force, s is the displacement of the object, and θ is the angle between the tension force and displacement vector. Since $\cos(90^\circ) = 0$, the work done by tension becomes zero.

Therefore, in circular motion, the tension force does not do any work on the object because the displacement is perpendicular to the tension force. The work done by tension is zero.

Other Memory Based Questions

Question 1. -3,7,31,2,145; pattern completion

Ans. Looking at the given sequence of numbers: -3, 7, 31, 2, 145, it is not immediately evident what the pattern or rule is for completing the sequence. However, let's analyze the numbers further to see if any patterns emerge.

If we calculate the differences between consecutive terms, we get:

$$7 - (-3) = 10$$

$$31 - 7 = 24$$

$$2 - 31 = -29$$

$$145 - 2 = 143$$

The differences between the terms are not following a consistent pattern. It appears that there might not be a simple mathematical pattern or rule governing the sequence. It's also possible that there may be a pattern or rule that is not immediately obvious or that additional terms are required to discern the pattern accurately.

Without further information or context, it is difficult to determine the exact pattern or complete the sequence confidently.

