

---

# Diffusion

Presented By: Massa Baali

# Discriminative vs. Generative Models

---

- **Discriminative models learn to discriminate**

- Determine the class given the input
    - Compute  $P(y|x)$

- **Generative models can generate**

- Produce more instances like the training data
    - Compute and/or draw from  $P(x,y)$

# Discriminative vs. Generative Models

---

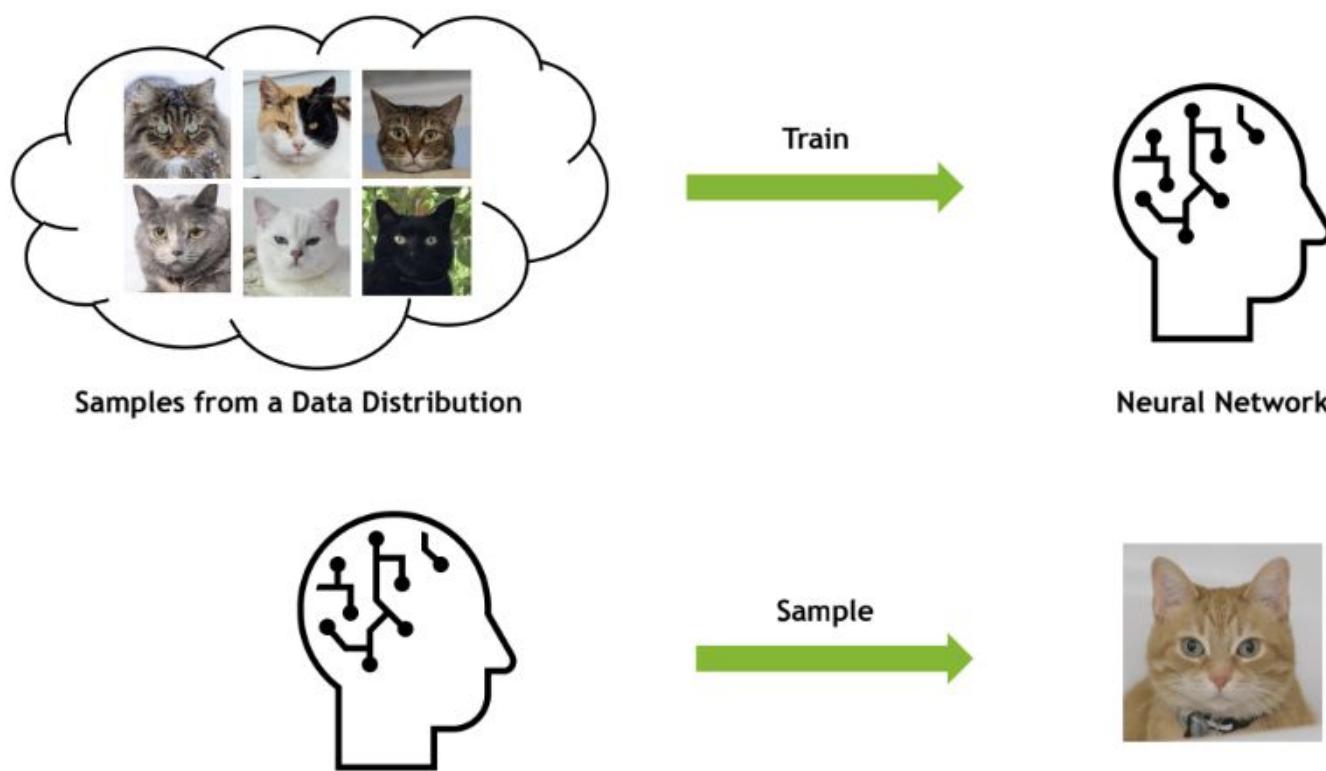
**Given a distribution of inputs  $X$  and labels  $Y$**

Discriminative models	Generative models
Discriminative models learn conditional distribution $P(Y   X)$	Generative models learn the joint distribution $P(Y, X)$
Learns decision boundary between classes.	Learns actual probability distribution of data.
Limited scope. Can only be used for classification tasks.	Can do both generative and discriminative tasks.

Harder problem, requires a deeper understanding of the distribution than discriminative models.

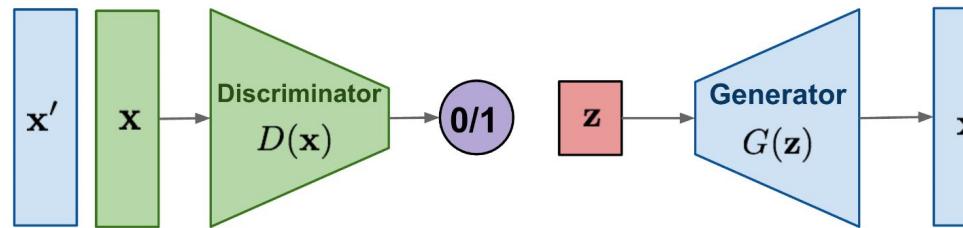
# Deep Generative learning

## Learning to generate data

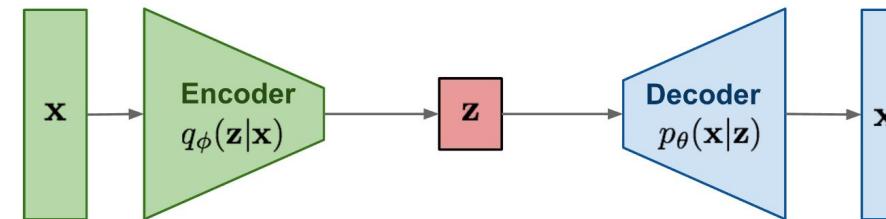


# Generative Models

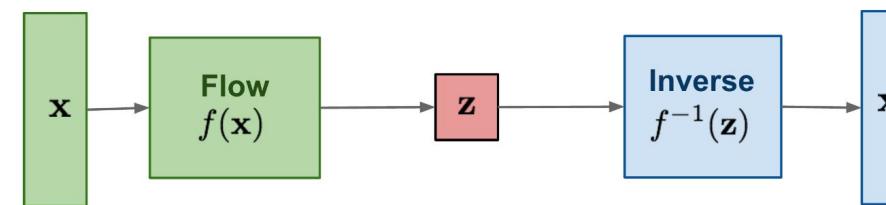
**GAN:** Adversarial training



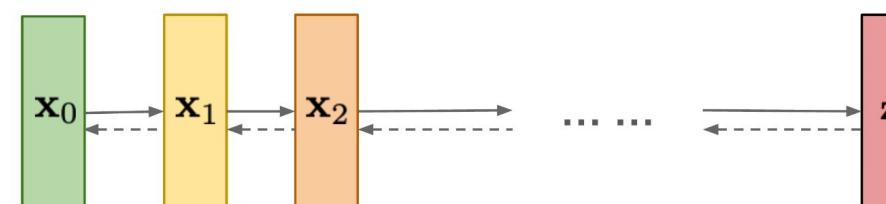
**VAE:** maximize variational lower bound



**Flow-based models:**  
Invertible transform of distributions

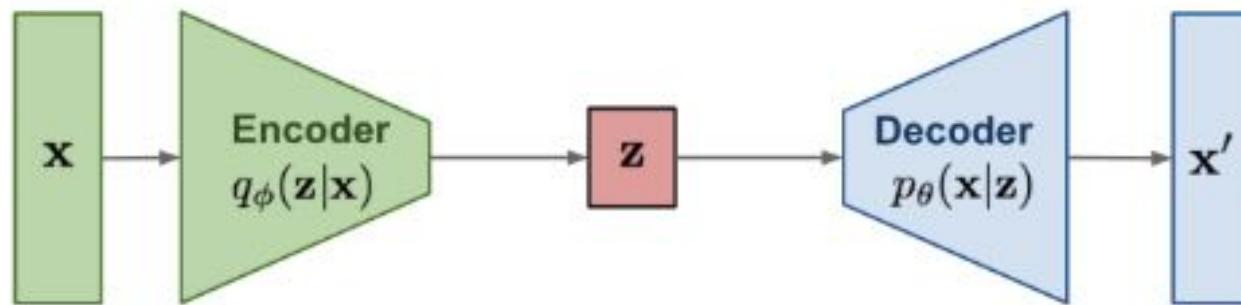


**Diffusion models:**  
Gradually add Gaussian noise and then reverse



# VAE (Recap)

---

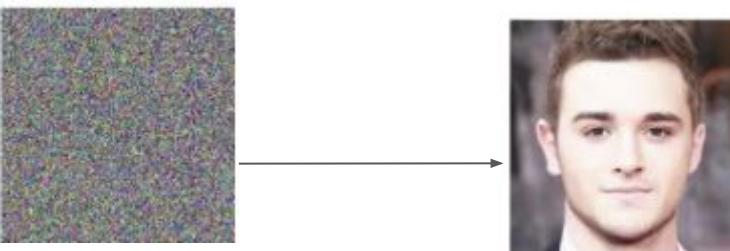


A generative model that learns to create new samples by:

- Compressing samples into a small code (*latent space*)
- Reconstructing samples from these codes

# Limitations of VAEs

---



The decoder must transform a standard Gaussian *all the way* to the target distribution in **one step**.

- Too large a gap to bridge in one step
- The decoder has to do ALL the work at once

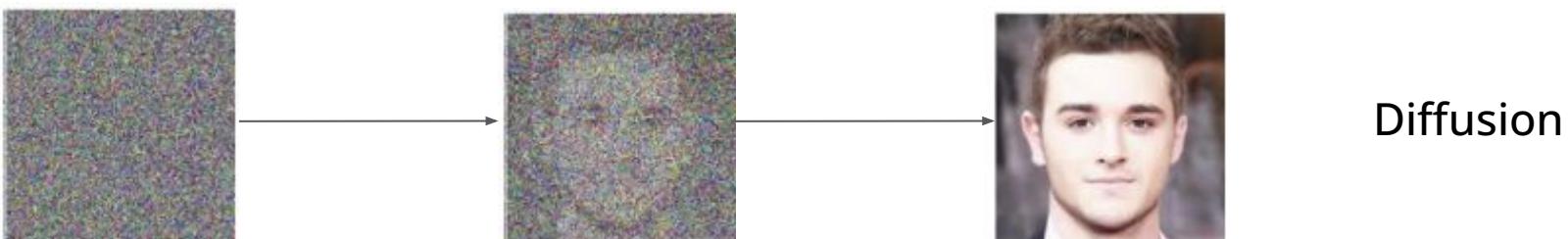
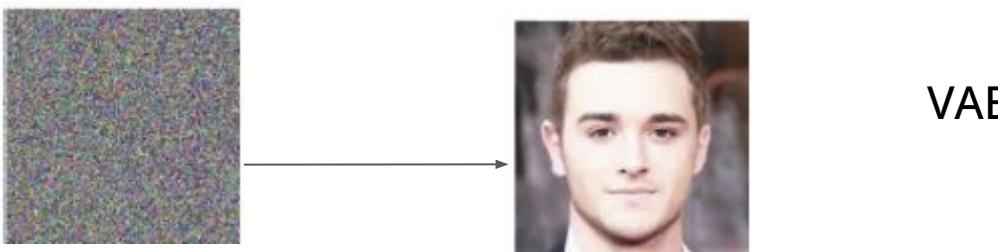
# Limitations of VAEs



## Result: Blurry, Low-Quality Images

# From Single-Step to Multi-Step Generation

---



# Hierarchicals VAEs

---

- Stack multiple VAEs with intermediate latent variables
- Each level:  $x_0 \rightarrow x_1 \rightarrow x_2 \rightarrow \dots \rightarrow x_t$
- Each step is a small, learnable transformation

## Diffusion Models: A Special Case

- Can be viewed as a hierarchical VAE with specific choices:
  - All latent variables have the same dimension as the original image
  - Forward process (encoder) is fixed (just adds Gaussian noise)
  - Only learn the reverse process (decoder/denoising)
  - Many steps (e.g.,  $T = 1000$ )

# Poll 1

---

Why do VAEs produce blurry images?

- A) Not enough training data
- B) Bad optimizer choice
- C) One-step jump from noise to image is too hard
- D) Network architecture is too small

# Poll 1

---

Why do VAEs produce blurry images?

- A) Not enough training data
- B) Bad optimizer choice
- C) One-step jump from noise to image is too hard
- D) Network architecture is too small

# Outline

---

- Definition of Diffusion
- Importance of Diffusion
- Explaining the process of diffusion
- Denoising Diffusion Implicit Models (DDIM)
- Diffusion Models from Stochastic Differential Equations and Score Matching Perspective
- Classifier-free Guidance (CFG)
- Performance Metrics: FID, IS, Precision and Recall
- Applications of Diffusion Models

# Definition

---



Adapted from: Ho, J., Jain, A. and Abbeel, P., 2020. Denoising diffusion probabilistic models. Advances in neural information processing systems, 33, pp.6840-6851.

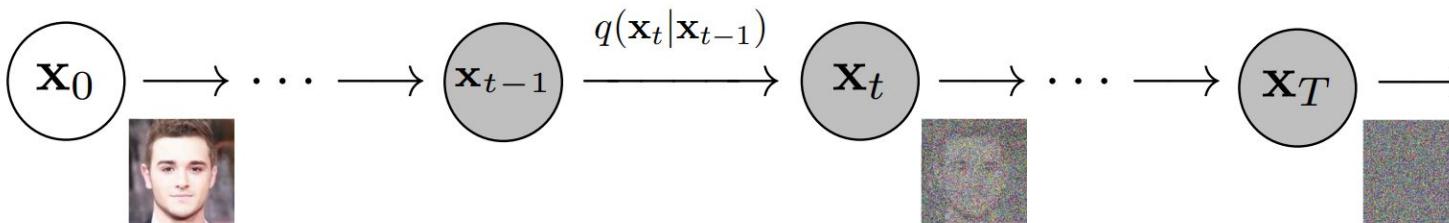
# Importance of Diffusion Models

---

- **High-Quality Outputs:** Exceptional generation of images, audio, and more with high fidelity.
- **Stable Training:** More consistent training outcomes compared to other models like GANs.

# Forward Process

---

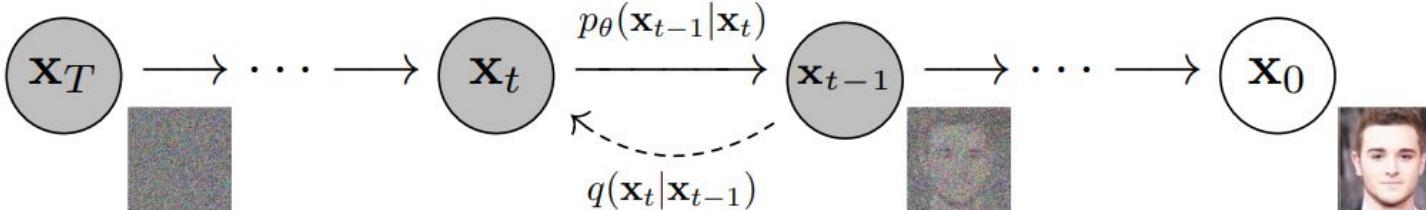


- **Gradual Noise Addition:** Over  $t$  time steps, noise is progressively added to an image.
- **Markov Chain Model:** The process, denoted by  $q$  takes the form of Markov Chain where the distribution at a particular time steps only depends on the sample from the immediate previous step.

Adapted from:  
<https://www.assemblyai.com/blog/diffusion-models-for-machine-learning-introduction/>

# Reverse Process

---



- **Denoise:** The model learns to gradually remove noise from the data at each step.
- **Reconstruction:** Aims to reconstruct the original image or data from its noisy version.
- The main goal of training a diffusion model is learning the reverse process specifically training  $p_\theta(x_{t-1} | x_t)$ .

Adapted from:  
<https://www.assemblyai.com/blog/diffusion-models-for-machine-learning-introduction/>

# How do we add noise?

---

- **Gaussian Noise:** The noise added in diffusion models follows a Gaussian (normal) distribution. This means at each diffusion step, we inject some random noise that has the familiar "bell curve" distribution.
- **Standard Normal Sampling:** We sample the noise from a standard normal distribution  $\mathcal{N}(0, \mathbf{I})$ , which has a mean of 0 and a variance of 1.

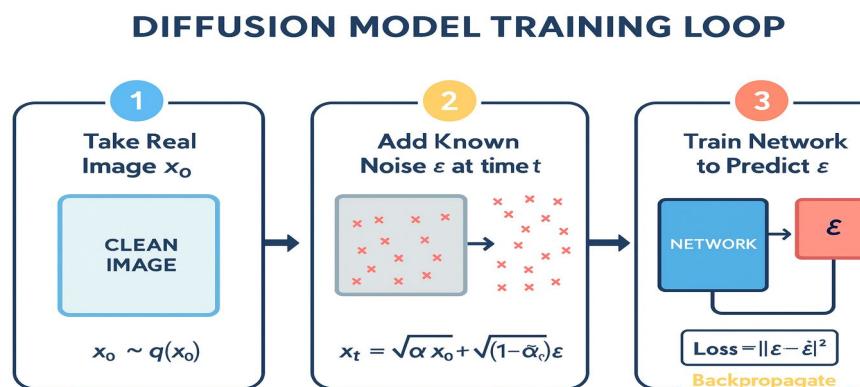
# What Does the Model Actually Learn?

## The Task:

- Input: Noisy image  $x_t$ , timestep  $t$
- Output: Predict the noise  $\epsilon$  that was added

## Why This Works:

- If we know the noise, we can subtract it
- Subtracting noise = denoising = moving toward clean data



# Training Objective

---

- **Minimize Error:** The goal is to minimize the Mean Squared Error (MSE) between the model's predicted noise and the actual noise added during the forward process.
- **High-Quality Reconstruction:** Effective training enables the model to accurately reconstruct original data from noise, enhancing its capability to generate high-quality samples.

$$L_{\text{simple}}(\theta) := \mathbb{E}_{t, \mathbf{x}_0, \boldsymbol{\epsilon}} \left[ \left\| \boldsymbol{\epsilon} - \boldsymbol{\epsilon}_\theta \left( \sqrt{\bar{\alpha}_t} \mathbf{x}_0 + \sqrt{1 - \bar{\alpha}_t} \boldsymbol{\epsilon}, t \right) \right\|^2 \right]$$

# Training

---

---

## Algorithm 1 Training

---

1: **repeat**

2:  $\mathbf{x}_0 \sim q(\mathbf{x}_0)$

3:  $t \sim \text{Uniform}(\{1, \dots, T\})$

4:  $\boldsymbol{\epsilon} \sim \mathcal{N}(\mathbf{0}, \mathbf{I})$

5: Take gradient descent step on

$$\nabla_{\theta} \|\boldsymbol{\epsilon} - \boldsymbol{\epsilon}_{\theta}(\sqrt{\bar{\alpha}_t} \mathbf{x}_0 + \sqrt{1 - \bar{\alpha}_t} \boldsymbol{\epsilon}, t)\|^2$$

6: **until** converged

---

# Sampling

---

---

## Algorithm 2 Sampling

---

```
1:  $\mathbf{x}_T \sim \mathcal{N}(\mathbf{0}, \mathbf{I})$ 
2: for  $t = T, \dots, 1$  do
3:    $\mathbf{z} \sim \mathcal{N}(\mathbf{0}, \mathbf{I})$  if  $t > 1$ , else  $\mathbf{z} = \mathbf{0}$ 
4:    $\mathbf{x}_{t-1} = \frac{1}{\sqrt{\alpha_t}} \left( \mathbf{x}_t - \frac{1-\alpha_t}{\sqrt{1-\bar{\alpha}_t}} \boldsymbol{\epsilon}_\theta(\mathbf{x}_t, t) \right) + \sigma_t \mathbf{z}$ 
5: end for
6: return  $\mathbf{x}_0$ 
```

---

## Poll 2

---

What does the neural network actually predict during training?

- A) The clean image  $x_0$
- B) The noisy image  $x_t$
- C) The noise  $\epsilon$  that was added
- D) The next timestep  $t+1$

## Poll 2

---

What does the neural network actually predict during training?

- A) The clean image  $x_0$
- B) The noisy image  $x_t$
- C) The noise  $\epsilon$  that was added
- D) The next timestep  $t+1$

# Outline

---

- Definition of Diffusion
- Importance of Diffusion
- Explaining the process of diffusion
- **Denoising Diffusion Implicit Models (DDIM)**
- Diffusion Models from Stochastic Differential Equations and Score Matching Perspective
- Classifier-free Guidance (CFG)
- Performance Metrics: FID, IS, Precision and Recall
- Applications of Diffusion Models

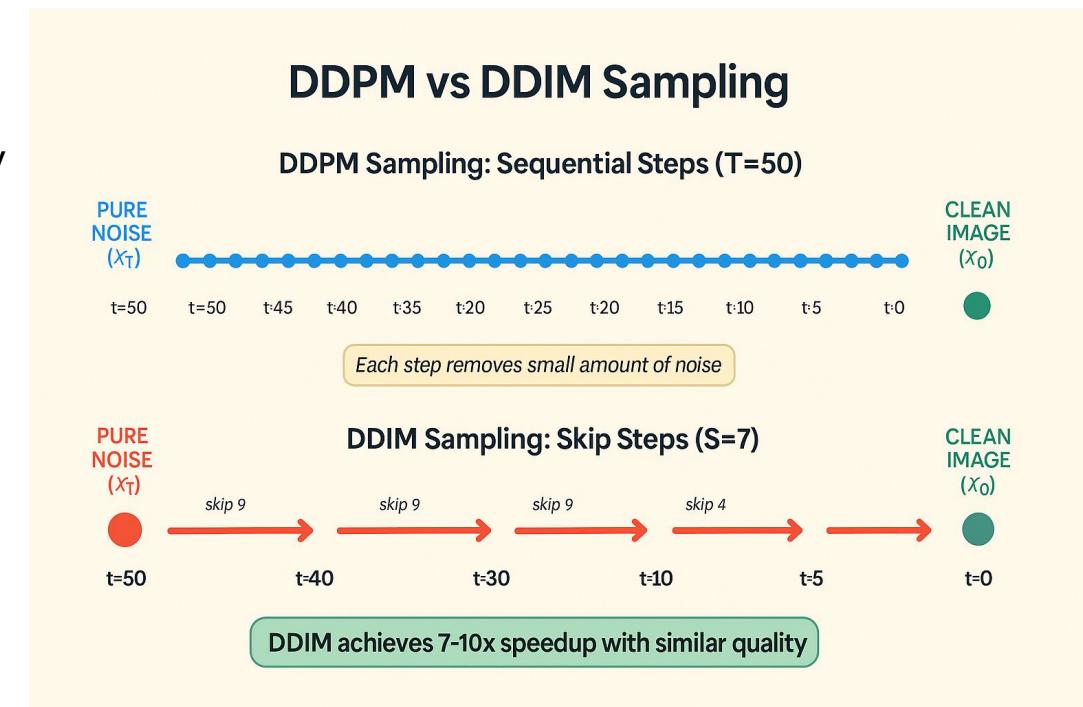
# The Speed Problem

## DDPM Challenge:

- Must run model  $T=1000$  times sequentially
- Each step removes tiny amount of noise
- Generation takes minutes per image

## DDIM Solution:

- Skip timesteps intelligently
- Use 10-50 steps instead of 1000
- 10-100x faster with similar quality



# Denoising Diffusion Implicit Models (DDIM)

---

## Limitations of DDPM Inference

**Sequential Denoising:** Must process each timestep in reverse to remove Gaussian noise. This happens due to the markov chain structure of the reverse process.

---

### Algorithm 2 Sampling

---

```
1:  $x_T \sim \mathcal{N}(0, I)$                                      ▷Initial isotropic gaussian noise sampling
2: for  $t = T, \dots, 1$  do
3:    $z \sim \mathcal{N}(0, I)$  if  $t > 1$  else  $z = 0$            ▷Sample random noise (if not last step)
4:    $\tilde{\epsilon} = \epsilon_\theta(x_t, t)$                    ▷Estimated noise in current noisy data
5:    $\tilde{x}_0 = \frac{1}{\sqrt{\bar{\alpha}_t}}(x_t - \sqrt{1 - \bar{\alpha}_t}\tilde{\epsilon})$     ▷Estimated  $x_0$  from estimated noise
6:    $\tilde{\mu} = \mu_t(x_t, \tilde{x}_0) \left( = \frac{1}{\sqrt{\bar{\alpha}_t}} \left( x_t - \frac{\beta_t}{\sqrt{1 - \bar{\alpha}_t}} \epsilon_\theta(x_t, t) \right) \right)$  ▷Mean for previous step sampling
7:    $x_{t-1} = \tilde{\mu} + \sigma_t z$                       ▷Previous step sampling
8: end for
9: return  $x_0$ 
```

---

# DDIM Overview

---

**Random Sampling:** In DDPM the inference step involves stochastic (random) sampling to reverse the diffusion process.

$$\tilde{x}_{t-1} = \mu_\theta(\tilde{x}_t, t) + \sigma_t(z_t - \epsilon_\theta(\tilde{x}_t, t))$$

denoising function      noise schedule      estimated noise

# DDIM Overview

---

**Deterministic Sampling:** DDIM uses a non-Markovian process allowing for fewer timesteps. It modifies the inference step by modifying the reverse process, making the process deterministic.

$$\mathbf{x}_{t-1} = \underbrace{\sqrt{\alpha_{t-1}} \left( \frac{\mathbf{x}_t - \sqrt{1 - \alpha_t} \epsilon_{\theta}^{(t)}(\mathbf{x}_t)}{\sqrt{\alpha_t}} \right)}_{\text{“predicted } \mathbf{x}_0\text{”}} + \underbrace{\sqrt{1 - \alpha_{t-1} - \sigma_t^2} \cdot \epsilon_{\theta}^{(t)}(\mathbf{x}_t)}_{\text{“direction pointing to } \mathbf{x}_t\text{”}} + \underbrace{\sigma_t \epsilon_t}_{\text{random noise}}$$

## Poll 3

---

DDIM is 10-50× faster than DDPM because:

- A) It uses a smaller neural network
- B) It skips timesteps intelligently
- C) It trains faster
- D) It uses a better GPU

## Poll 3

---

DDIM is 10-50× faster than DDPM because:

- A) It uses a smaller neural network
- B) It skips timesteps intelligently
- C) It trains faster
- D) It uses a better GPU

# Outline

---

- Definition of Diffusion
- Importance of Diffusion
- Explaining the process of diffusion
- Denoising Diffusion Implicit Models (DDIM)
- Diffusion Models from Stochastic Differential Equations and Score Matching Perspective
- Classifier-free Guidance (CFG)
- Performance Metrics: FID, IS, Precision and Recall
- Applications of Diffusion Models

# Why Move to Continuous Time?

---

## **Discrete View (DDPM):**

- Fixed timesteps:  $t = 1, 2, 3, \dots, 1000$
- Fixed noise schedule:  $\beta_1, \beta_2, \dots, \beta_{1000}$
- Limited flexibility

## **Continuous View (SDE):**

- Continuous time:  $t \in [0, T]$
- Smooth noise function:  $\beta(t)$
- Can discretize flexibly at sampling time

# From ODEs to SDEs - Modeling Diffusion

---

- Diffusion Models (Discrete View)
  - Forward: Add noise in  $T$  discrete steps (e.g.,  $T=1000$ )
  - Backward: Remove noise in  $T$  discrete steps

## **Problem:**

- Can't flexibly adjust ***steps*** during generation
- Limited flexibility in how we add/remove noise
- Can't use powerful continuous-time solvers

# From ODEs to SDEs - Modeling Diffusion

---

## Why SDEs for Diffusion Models?

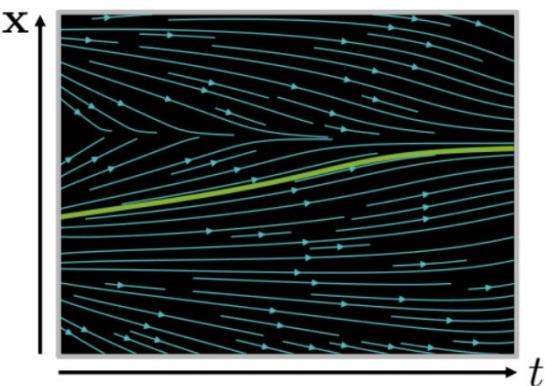
- Deterministic Sampling via Probability Flow ODEs
  - Derive equivalent ODE from any diffusion SDE (no randomness in generation)
  - Enable exact reconstruction, smooth interpolation, and image editing
- Unified Framework
  - DDPM, score-based, and flow models as special cases of one SDE
  - Transfer insights: train with one method, sample with another

# Ordinary Differential Equations

---

**Ordinary Differential Equation (ODE):**

$$\frac{dx}{dt} = f(x, t) \text{ or } dx = f(x, t)dt$$



Analytical Solution: 
$$x(t) = x(0) + \int_0^t f(x, \tau) d\tau$$

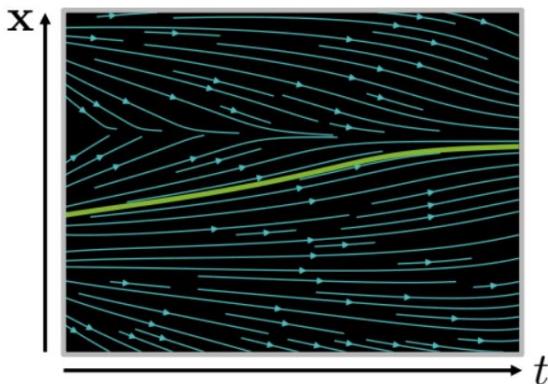
Iterative Numerical Solution: 
$$x(t + \Delta t) \approx x(t) + f(x(t), t) \Delta t$$

Slide credit to: <https://cvpr2022-tutorial-diffusion-models.github.io/>

# Stochastic Differential Equations

Ordinary Differential Equation (ODE):

$$\frac{dx}{dt} = f(x, t) \text{ or } dx = f(x, t)dt$$



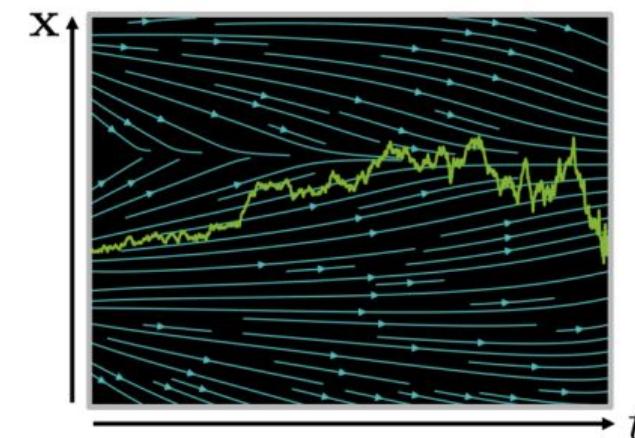
Analytical Solution:  $x(t) = x(0) + \int_0^t f(x, \tau)d\tau$

Iterative Numerical Solution:  $x(t + \Delta t) \approx x(t) + f(x(t), t)\Delta t$

Slide credit to: <https://cvpr2022-tutorial-diffusion-models.github.io/>

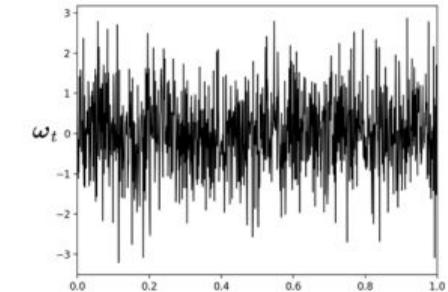
Stochastic Differential Equation (SDE):

$$\frac{dx}{dt} = \underbrace{f(x, t)}_{\text{drift coefficient}} + \underbrace{\sigma(x, t)\omega_t}_{\text{diffusion coefficient}}$$
$$(dx = f(x, t)dt + \sigma(x, t)d\omega_t)$$

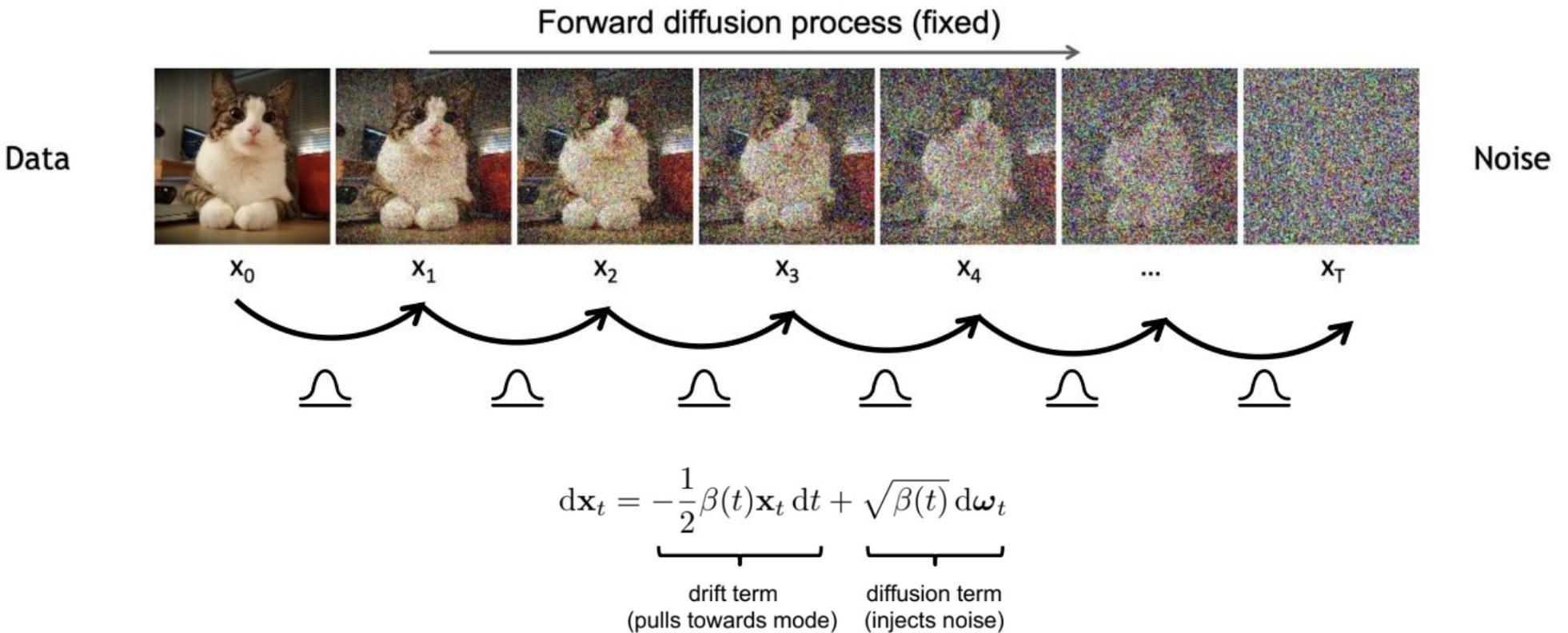


$$x(t + \Delta t) \approx x(t) + f(x(t), t)\Delta t + \sigma(x(t), t)\sqrt{\Delta t} \mathcal{N}(\mathbf{0}, \mathbf{I})$$

Wiener Process  
(Gaussian White Noise)

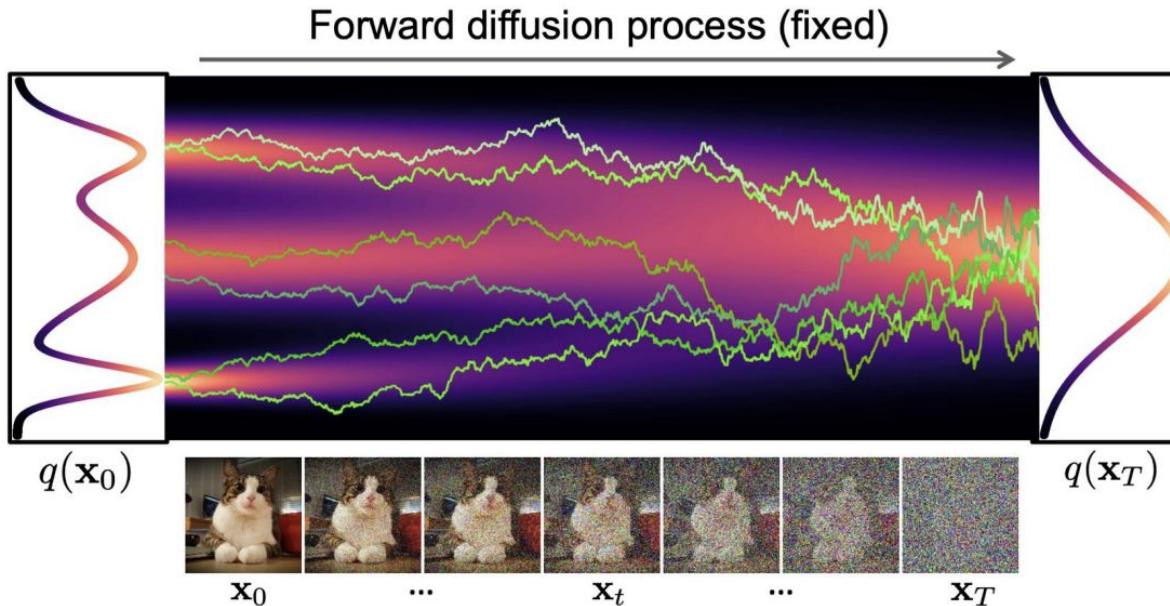


# Forward Diffusion Process as SDEs



Slide credit to: <https://cvpr2022-tutorial-diffusion-models.github.io/>

# Forward Diffusion Process as SDEs



$$d\mathbf{x}_t = -\frac{1}{2}\beta(t)\mathbf{x}_t dt + \sqrt{\beta(t)} d\omega_t$$

drift term (pulls towards mode)      diffusion term (injects noise)

Slide credit to: <https://cvpr2022-tutorial-diffusion-models.github.io/>

# Forward Diffusion Process as SDEs

---

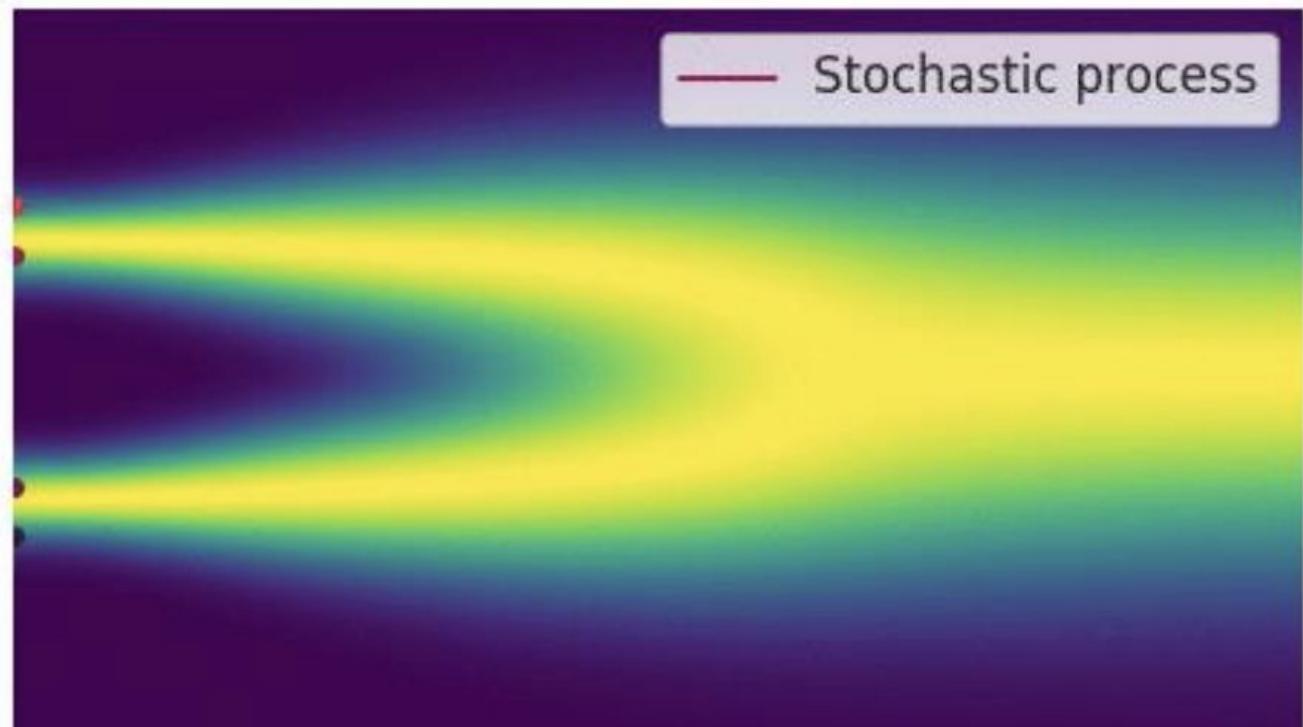
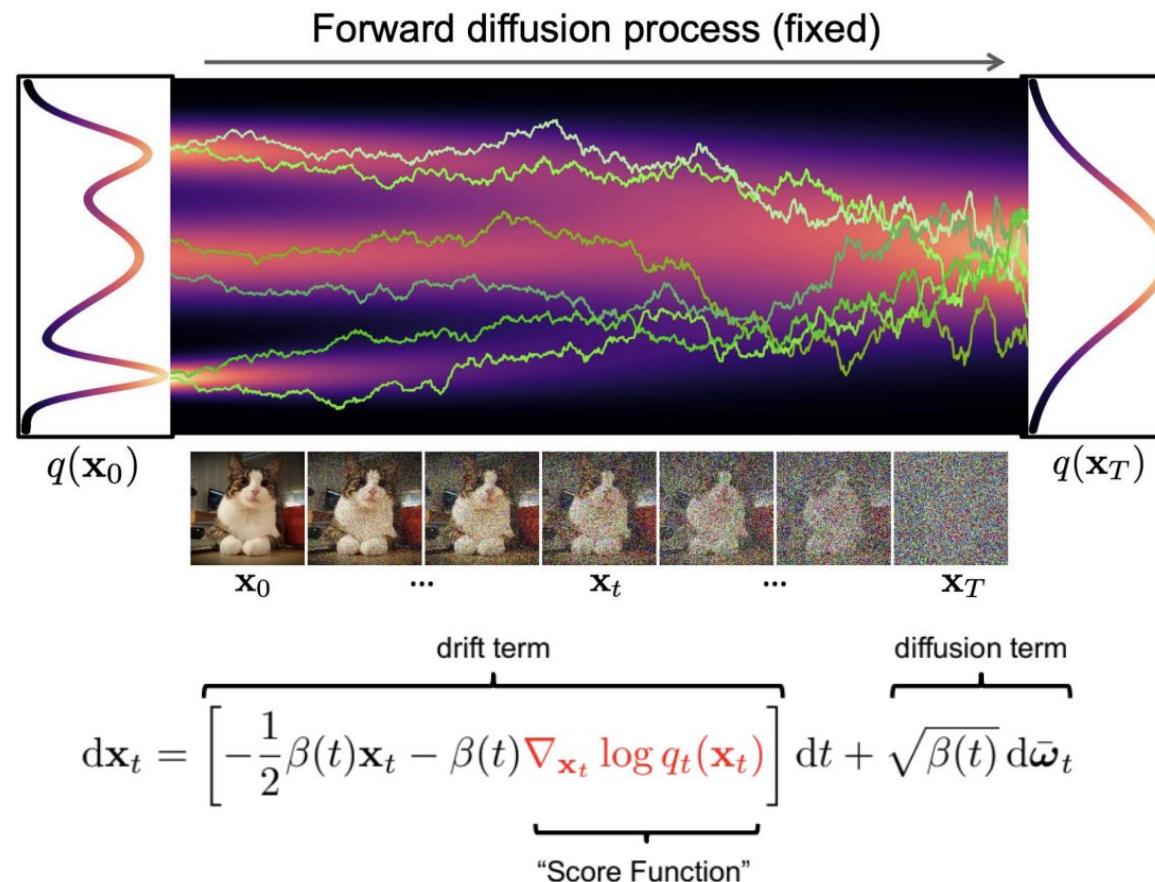


Figure credit to: <https://yang-song.net/blog/2021/score/>

# Generative Reverse SDEs



Slide credit to: <https://cvpr2022-tutorial-diffusion-models.github.io/>

# Generative Reverse SDEs

---

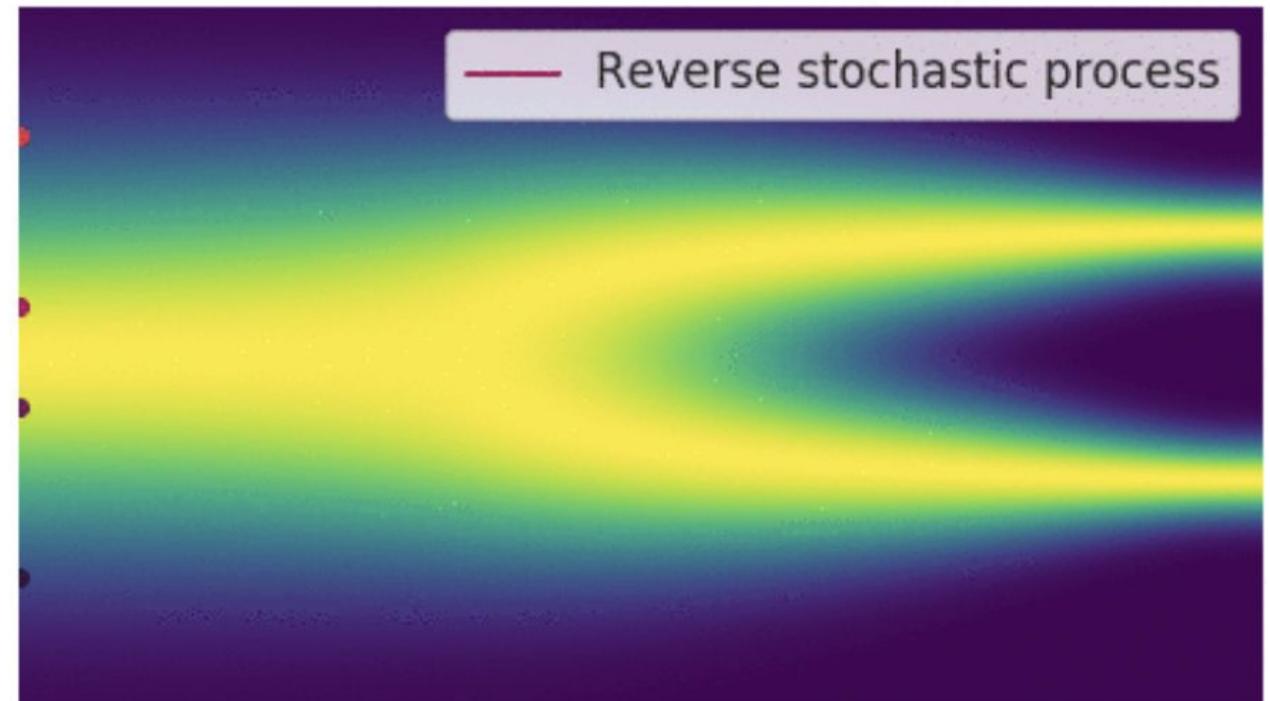
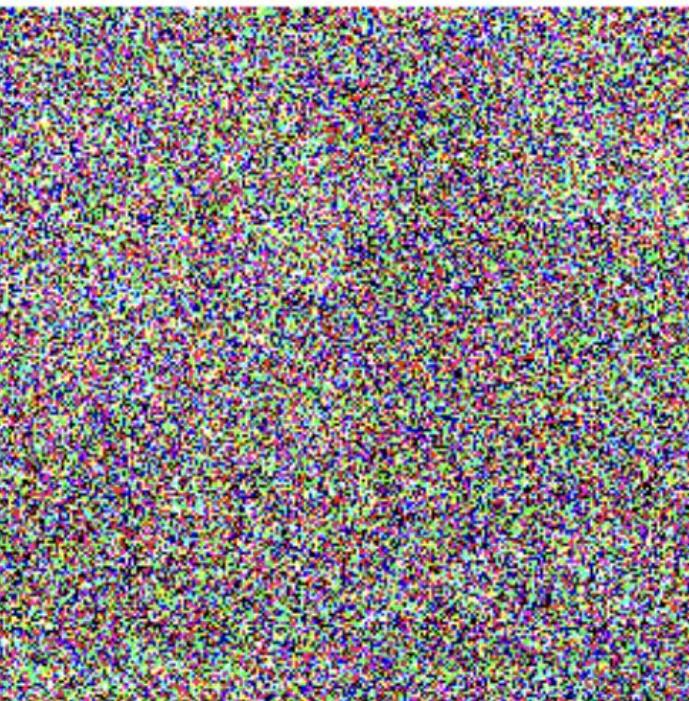


Figure credit to: <https://yang-song.net/blog/2021/score/>

# The Score Function: Why It Matters

---

## Traditional Approach:

Learn  $p(x)$  directly → requires normalizing constant  $Z$  →  $Z$  is intractable for complex data

## Score-Based Approach:

Learn  $\nabla_x \log p(x)$  instead → no  $Z$  needed! →  $Z$  cancels in gradient

# The Score Function: Why It Matters

---

## Connection to Diffusion:

Predicting noise  $\epsilon \Leftrightarrow$  Estimating score  $\nabla_x \log p(x)$

Same objective, different interpretation

Score tells us direction toward high probability

# Score Matching

---

- General form of probability density function

$$p_{\theta}(\mathbf{x}) = \frac{e^{-f_{\theta}(\mathbf{x})}}{Z_{\theta}}$$

- Maximizing the log-likelihood requires us to know  $Z_{\theta}$ 
  - Often intractable
- Instead, we can model the score function

$$\nabla_{\mathbf{x}} \log p(\mathbf{x})$$

Slide credit to: Hao Chen [https://deeplearning.cs.cmu.edu/S25/document/slides/lec23.diffusion\\_s25.pdf](https://deeplearning.cs.cmu.edu/S25/document/slides/lec23.diffusion_s25.pdf)

# Denoising Score Matching

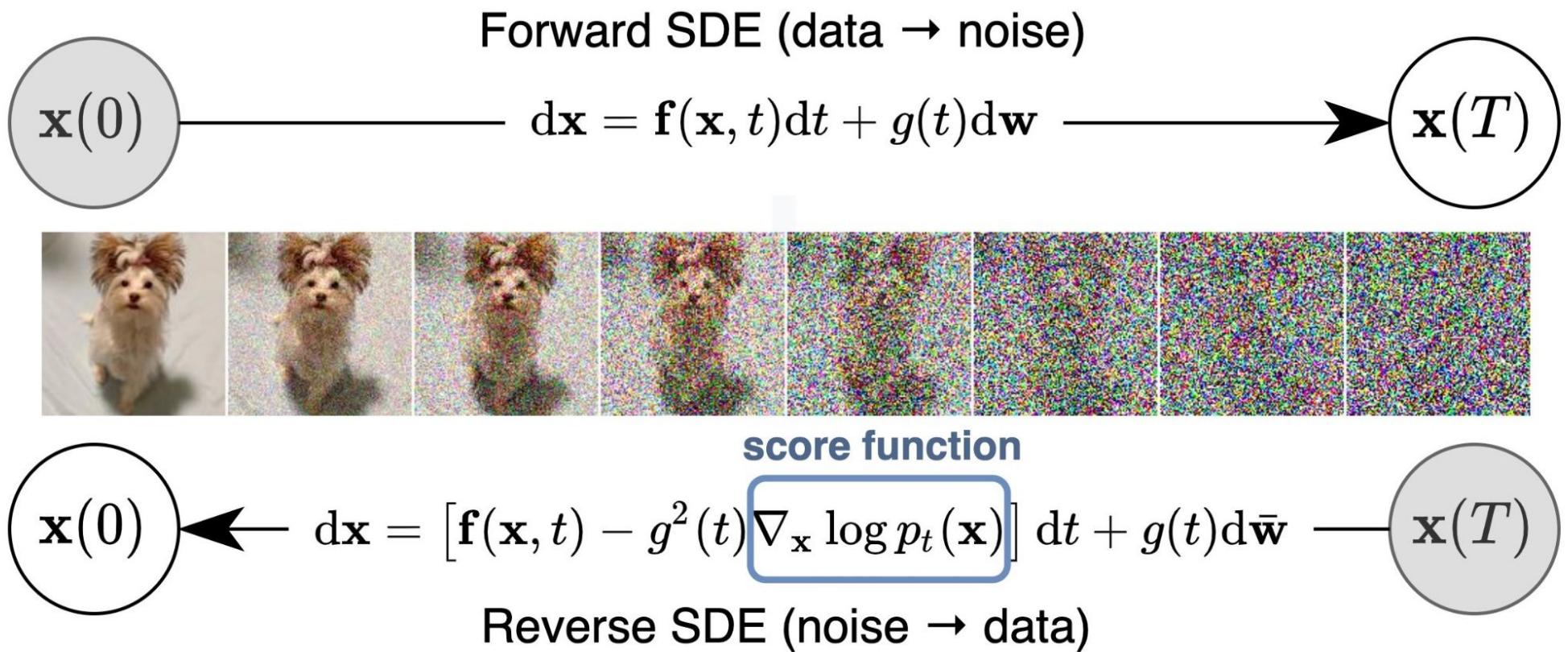


Figure credit to: <https://yang-song.net/blog/2021/score/>

# Weighted Diffusion Objective

---

Denoising score matching objective with loss weighting

$$\min_{\theta} \mathbb{E}_{t \sim \mathcal{U}(0, T)} \mathbb{E}_{\mathbf{x}_0 \sim q_0(\mathbf{x}_0)} \mathbb{E}_{\epsilon \sim \mathcal{N}(\mathbf{0}, \mathbf{I})} \frac{\lambda(t)}{\sigma_t^2} \|\epsilon - \epsilon_{\theta}(\mathbf{x}_t, t)\|_2^2$$

Problem: AND Why weighting  $\lambda(t)$ ?

- Different timesteps contribute differently
- Early timesteps (high noise): easy to predict, less important
- Late timesteps (low noise): hard to predict, very important

Common choices:

- $\lambda(t) = 1$ : Simple, uniform weighting (DDPM)
- $\lambda(t) = \sigma_t^2$ : Better perceptual quality
- $\lambda(t) = \beta(t)$ : Maximum likelihood

## Poll 4

---

The score function  $\nabla \log p(x)$  points toward:

- A) Random directions
- B) Low probability regions
- C) High probability regions (data)
- D) The origin

## Poll 4

---

The score function  $\nabla \log p(x)$  points toward:

- A) Random directions
- B) Low probability regions
- C) High probability regions (data)
- D) The origin

# Outline

---

- Definition of Diffusion
- Importance of Diffusion
- Explaining the process of diffusion
- Denoising Diffusion Implicit Models (DDIM)
- Diffusion Models from Stochastic Differential Equations and Score Matching Perspective
- **Classifier-free Guidance (CFG)**
- Performance Metrics: FID, IS, Precision and Recall
- Applications of Diffusion Models

# Conditional Generation: The Control Problem

---

## **Unconditional Diffusion:**

- Sample: Random noise → Random image from training distribution
- No control over output class/content

## **The Need for Control:**

- "Draw a sunset over mountains"
- "Create a portrait in Van Gogh style"

# Conditional Generation: The Control Problem

---

1. **Classifier Guidance:** Train separate classifier, use gradients
  - $\times$  Requires training extra model
  - $\times$  Classifier can be noisy at high  $t$
2. **Classifier-Free Guidance (CFG):** Train one model for both
  - $\checkmark$  Single model
  - $\checkmark$  Better quality

# Classifier-free Guidance (CFG)

---

**Limited Class Control:** Previous method is incapable of generating an image for a given class

**Purpose of CFG:** it allows for targeted generation of images by conditioning the model on a specific class label.

# How Classifier-Free Guidance Works

---

## Training:

- Randomly drop condition 10-20% of time
- Model learns both:
  - $\varepsilon_\theta(x_t, \text{class}) \leftarrow \text{conditional}$
  - $\varepsilon_\theta(x_t, \emptyset) \leftarrow \text{unconditional}$

## Sampling:

$$\tilde{\varepsilon} = \varepsilon_{\text{uncond}} + \gamma(\varepsilon_{\text{cond}} - \varepsilon_{\text{uncond}})$$

↑                      ↑      ↑  
baseline      scale   direction toward class

## Guidance scale $\gamma$ :

- $\gamma = 0$ : Ignore condition (random)
- $\gamma = 1$ : Standard conditioning
- $\gamma > 1$ : Stronger conditioning (more typical of class)

# Classifier-free Guidance (CFG)

---

---

**Algorithm 1** Classifier guided diffusion sampling, given a diffusion model  $(\mu_\theta(x_t), \Sigma_\theta(x_t))$ , classifier  $p_\phi(y|x_t)$ , and gradient scale  $s$ .

---

Input: class label  $y$ , gradient scale  $s$   
 $x_T \leftarrow$  sample from  $\mathcal{N}(0, \mathbf{I})$   
**for all**  $t$  from  $T$  to 1 **do**  
     $\mu, \Sigma \leftarrow \mu_\theta(x_t), \Sigma_\theta(x_t)$   
     $x_{t-1} \leftarrow$  sample from  $\mathcal{N}(\mu + s\Sigma \nabla_{x_t} \log p_\phi(y|x_t), \Sigma)$   
**end for**  
**return**  $x_0$

---

# Outline

---

- Definition of Diffusion
- Importance of Diffusion
- Explaining the process of diffusion
- Denoising Diffusion Implicit Models (DDIM)
- Diffusion Models from Stochastic Differential Equations and Score Matching Perspective
- Classifier-free Guidance (CFG)
- **Performance Metrics: FID, IS, Precision and Recall**
- Applications of Diffusion Models

# Performance Metrics

---

## **Fréchet Inception Distance (FID)**

Measures the distance between feature vectors of real and generated images; lower scores indicate better image quality and similarity to real images.

## **Inception Score (IS)**

Assesses the diversity and clarity of generated images using a pre-trained model; higher scores denote better image clarity and variety.

## **Precision and Recall**

Evaluates the quality and diversity of generated images; high precision indicates realistic images, while high recall shows variety close to the actual dataset.

# Outline

---

- Definition of Diffusion
- Importance of Diffusion
- Explaining the process of diffusion
- Denoising Diffusion Implicit Models (DDIM)
- Diffusion Models from Stochastic Differential Equations and Score Matching Perspective
- Classifier-free Guidance (CFG)
- Performance Metrics: FID, IS, Precision and Recall
- Applications of Diffusion Models

# Latent DDPM

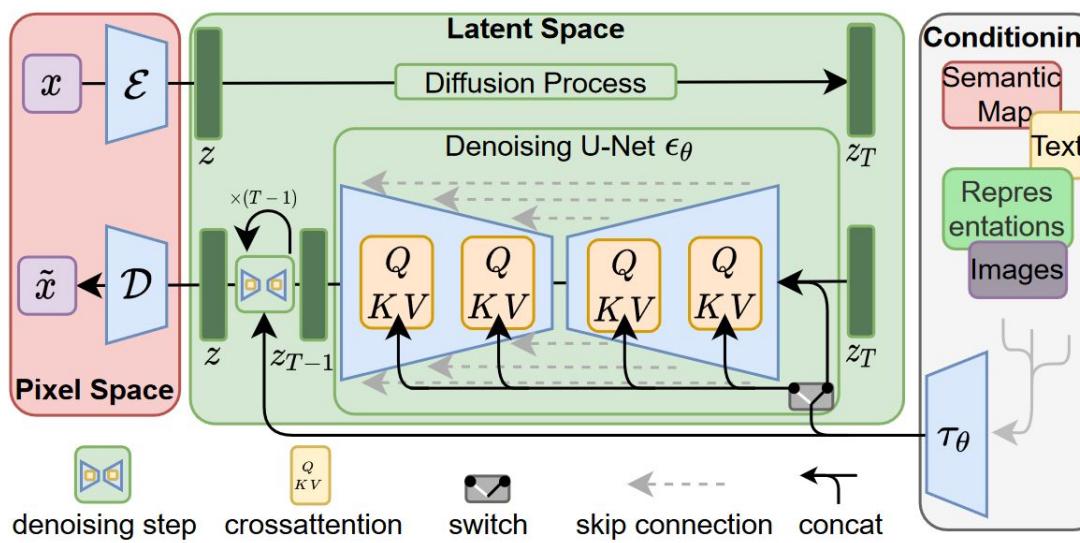
---

## Why Latent Space?

- **Efficiency:** Running diffusion models directly in pixel space is computationally expensive. Latent Diffusion Models (LDMs) operate in a compressed latent space, drastically reducing computational cost and complexity.
- **Data Compression:** LDMs capture the essential ***features*** of high-dimensional data like images in a more compact and structured form, enhancing processing efficiency.
- **Improved Performance:** Working in latent space focuses the model on **relevant aspects** of the *data*, improving both the efficiency and effectiveness of the generation process.

# Latent DDPM

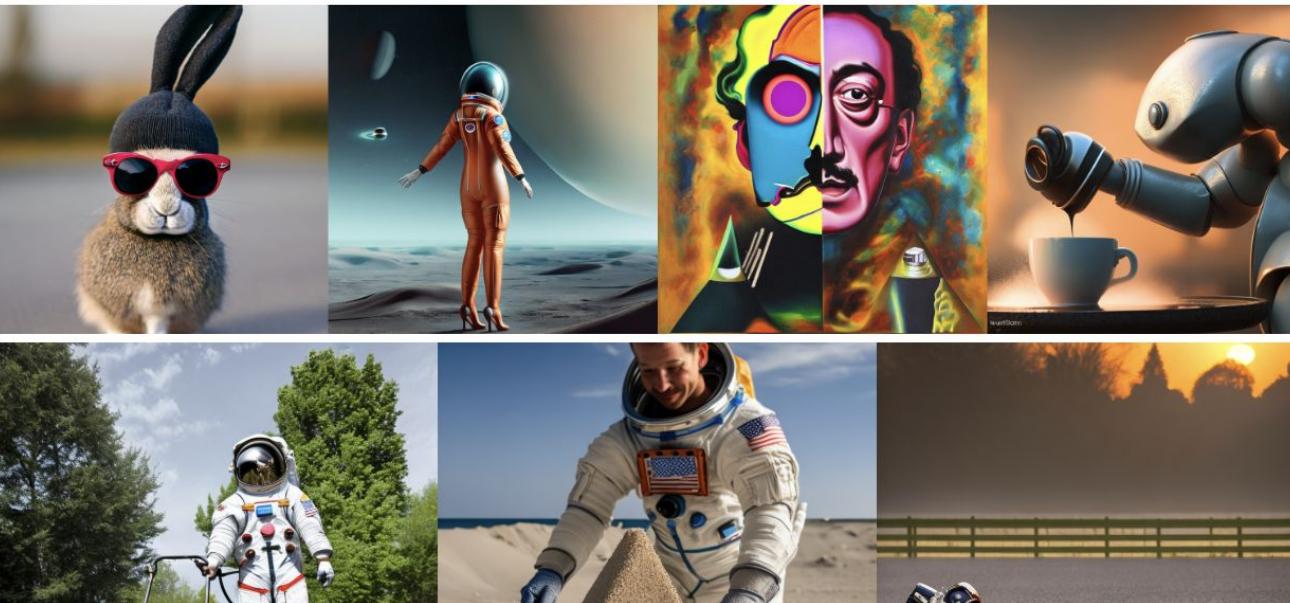
- Map data into latent space using VAE or any other similar approach
1. **Encoder:** Compresses input data into a latent representation.
  2. **Decoder:** Reconstructs the original data from the latent representation.



# Stable Diffusion

---

- Stable Diffusion is a latent text-to-image diffusion model.

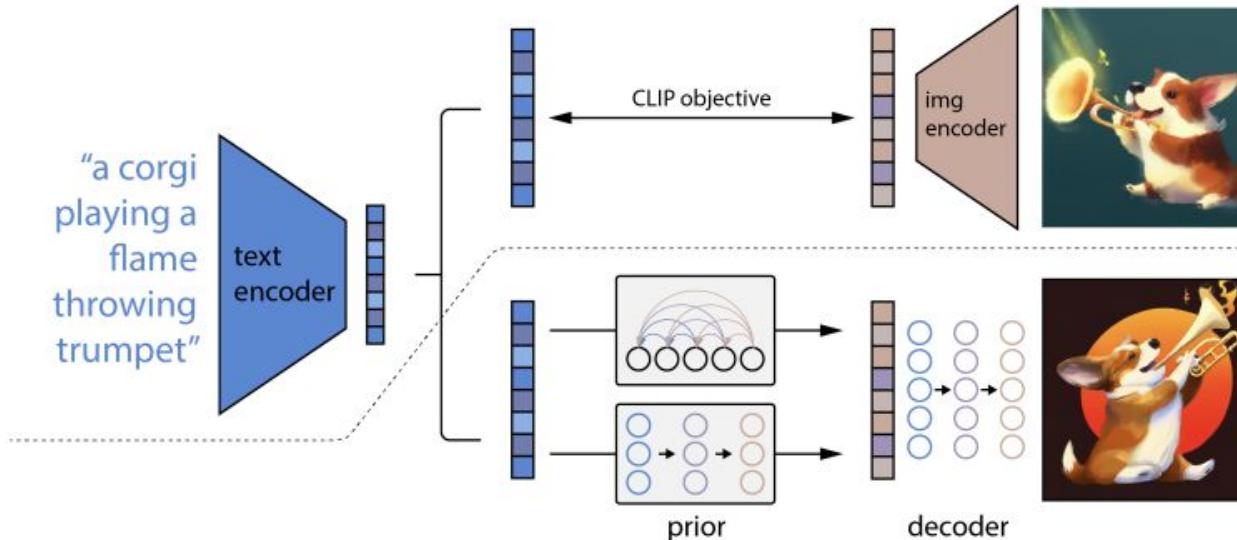


Stability AI. <https://github.com/Stability-AI/stablediffusion>

# DALLE

---

- DALLE is a text-to-image generative model that creates coherent, high-resolution images from natural-language prompts.



Ramesh et al. Hierarchical Text-Conditional Image Generation with CLIP Latents

# DiT

---

- DiT (Diffusion Transformer) is a diffusion model architecture that replaces U-Nets with Transformers to more effectively denoise and generate high-quality images.



Peebles et al. Scalable Diffusion Models with Transformers. 2020.

# MAR

---

- An autoregressive model with diffusion loss



Li et al. Autoregressive Image Generation without Vector Quantization. 2024.

# Key Takeaways

---

## **1. Many Small Steps Beat One Big Jump**

VAEs try to generate in one step and fail (blurry images). Diffusion uses 1000 tiny steps instead. Each small change is easy to learn. Result: sharp, high-quality outputs.

## **2. Forward is Fixed, Reverse is Learning**

Forward process: add noise using a fixed equation (no training needed). Reverse process: train neural network to predict the noise. Training objective: MSE loss.

## **3. Three Views, Same Model**

Denoising, score-based, and SDEs are equivalent perspectives. Understanding one helps understand all.

---

Thank you :-)