Outline

- Generative Models Basics
- Autoregressive Models
- Autoencoder and Variational Autoencoder
- Generative Adversarial Network
- **Diffusion Models**

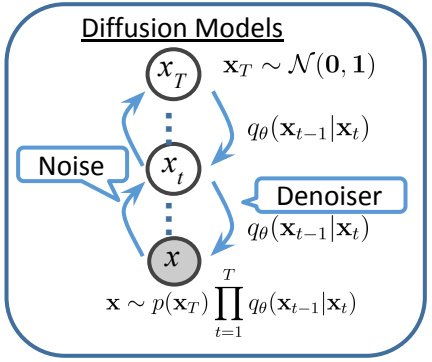
Denoising Diffusion Models

"Diffusion Models Beat GANs on Image Synthesis"
Dhariwal & Nichol, OpenAI, 2021

"Cascaded Diffusion Models for High Fidelity Image Generation" Ho et al., Google, 2021

Slide credit: Kreis, Gao, Vahdat, CVPR 2022 Tutorial

Denoising Diffusion Models



Forward diffusion process (fixed)

Data

Noise

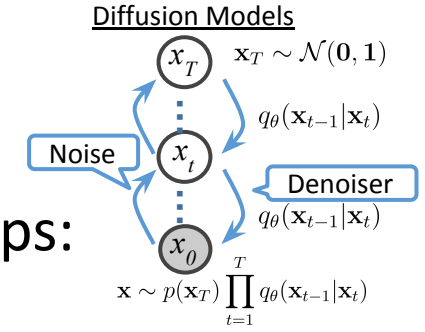
Reverse denoising process (generative)

Denoising Diffusion Models

Sohl-Dickstein et al., Deep Unsupervised Learning using Nonequilibrium Thermodynamics, ICML 2015
Ho et al., Denoising Diffusion Probabilistic Models, NeurIPS 2020
Song et al., Score-Based Generative Modeling through Stochastic Differential Equations, ICLR 2021

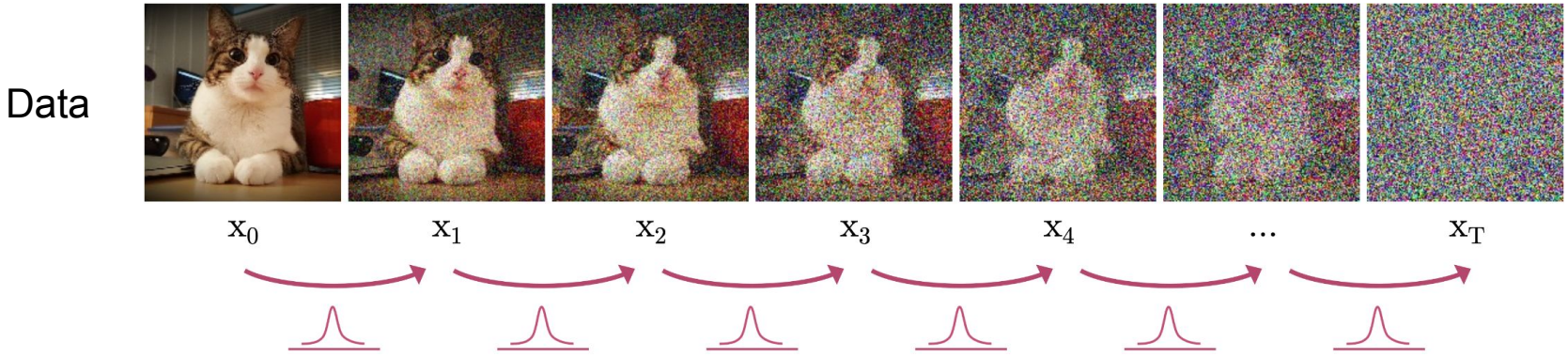
Slide credit: Kreis, Gao, Vahdat, CVPR 2022 Tutorial

Forward Diffusion Process



The formal definition of the forward process in T steps:

Forward diffusion process (fixed)

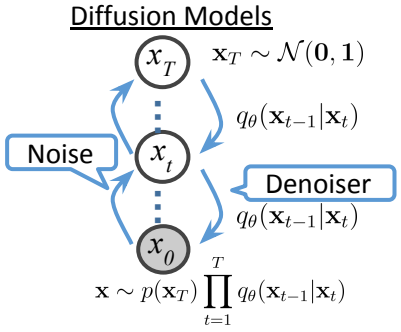


$$q(\mathbf{x}_t|\mathbf{x}_{t-1}) = \mathcal{N}(\mathbf{x}_t; \sqrt{1-\beta_t}\mathbf{x}_{t-1}, \beta_t\mathbf{I}) \quad \rightarrow \quad q(\mathbf{x}_{1:T}|\mathbf{x}_0) = \prod_{t=1}^{T} q(\mathbf{x}_t|\mathbf{x}_{t-1}) \qquad \text{(joint)}$$

β_s controls the *noise schedule* such that $\mathbf{x}_T$ is complete noise.

Slide credit: Kreis, Gao, Vahdat, CVPR 2022 Tutorial

Diffusion Kernel



Forward diffusion process (fixed)

Data

Noise

$\mathbf{x}_0$ $\mathbf{x}_1$ $\mathbf{x}_2$ $\mathbf{x}_3$ $\mathbf{x}_4$ ... $\mathbf{x}_T$

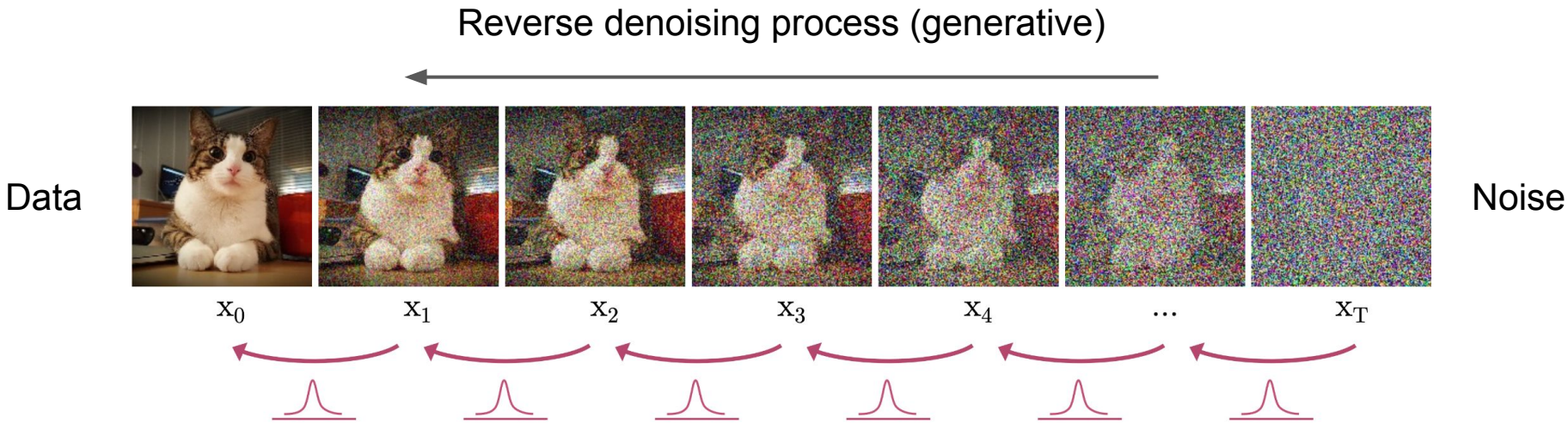
"Skip" *t* steps **(Diffusion Kernel)**

$$q(\mathbf{x}_t|\mathbf{x}_0) = \mathcal{N}(\mathbf{x}_t; \sqrt{\bar{\alpha}_t}\mathbf{x}_0, (1-\bar{\alpha}_t)\mathbf{I})$$

$$\text{Define } \bar{\alpha}_t = \prod_{s=1}^{t}(1-\beta_s)$$

Slide credit: Kreis, Gao, Vahdat, CVPR 2022 Tutorial

Reverse Denoising Process



$$q(\mathbf{x}_{t-1}|\mathbf{x}_t) \propto q(\mathbf{x}_{t-1})q(\mathbf{x}_t|\mathbf{x}_{t-1})$$

Intractable... 

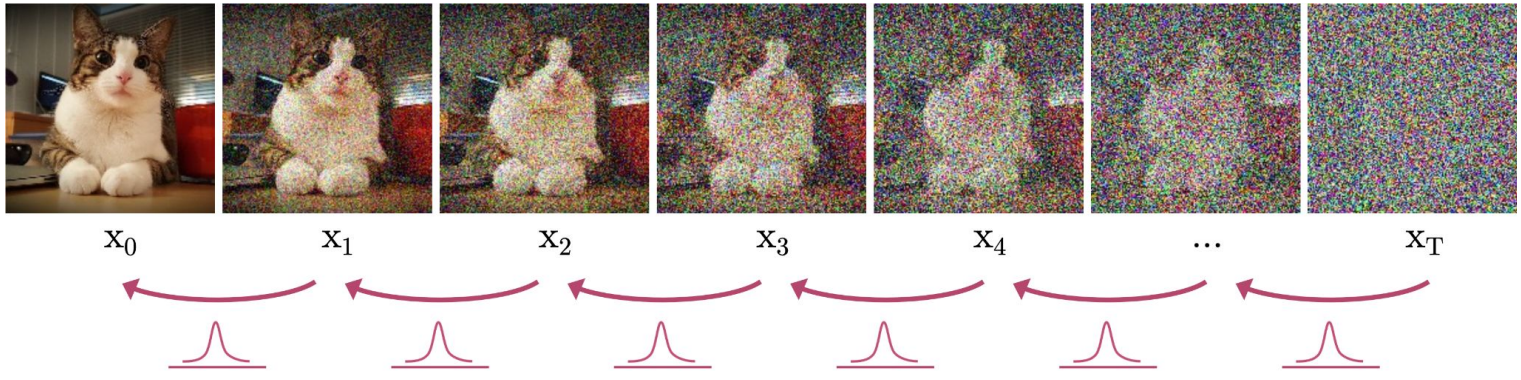
Idea: Approximate $q(\mathbf{x}_{t-1}|\mathbf{x}_t)$. Use a Normal distribution if β_t is small in each forward diffusion step.

Slide credit: Kreis, Gao, Vahdat, CVPR 2022 Tutorial

Reverse Denoising Process

Formal definition of forward and reverse processes in T steps:

Reverse denoising process (generative)



$$p_\theta(\mathbf{x}_T) = \mathcal{N}(\mathbf{x}_T; \mathbf{0}, \mathbf{I})$$

$$p_\theta(\mathbf{x}_{t-1}|\mathbf{x}_t) = \mathcal{N}(\mathbf{x}_{t-1}; \underbrace{\mu_\theta(\mathbf{x}_t, t)}, \sigma_t^2\mathbf{I})$$

$$\frac{1}{\sqrt{\alpha_t}}\left(\mathbf{x}_t - \frac{1-\alpha_t}{\sqrt{1-\bar{\alpha}_t}}\underbrace{\boldsymbol{\epsilon}(\mathbf{x}_t, t)}\right)$$

Trainable network

$$\rightarrow \quad p_\theta(\mathbf{x}_{0:T}) = p(\mathbf{x}_T)\prod_{t=1}^{T} p_\theta(\mathbf{x}_{t-1}|\mathbf{x}_t)$$

Apply $\mu_\theta(\cdot)$ T times!

Slide credit: Kreis, Gao, Vahdat, CVPR 2022 Tutorial

U-Net: Image-to-Image CNN network

The trainable noise prediction network is usually a U-Net

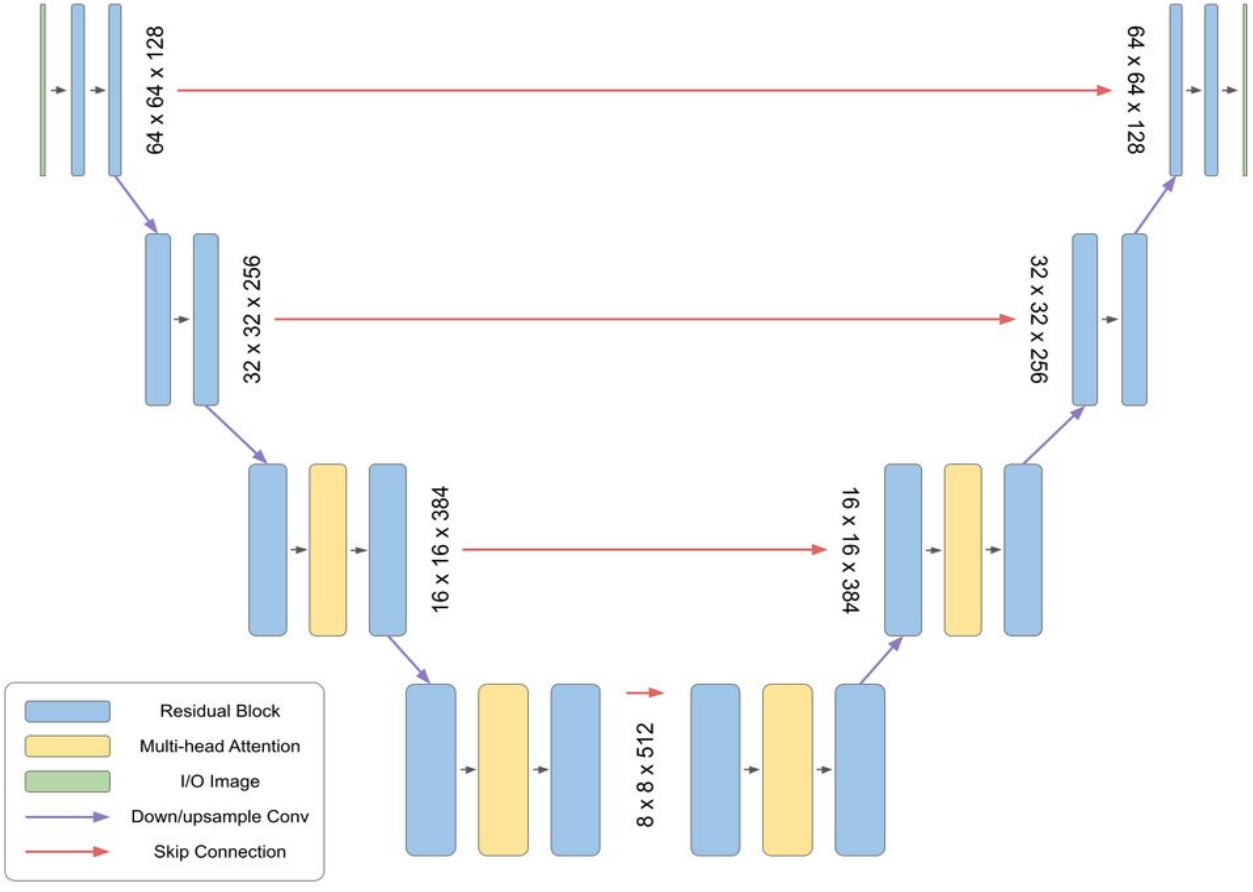


Image Source: https://www.assemblyai.com/blog/minimagen-build-your-own-imagen-text-to-image-model/

U-Net: Image-to-Image CNN network

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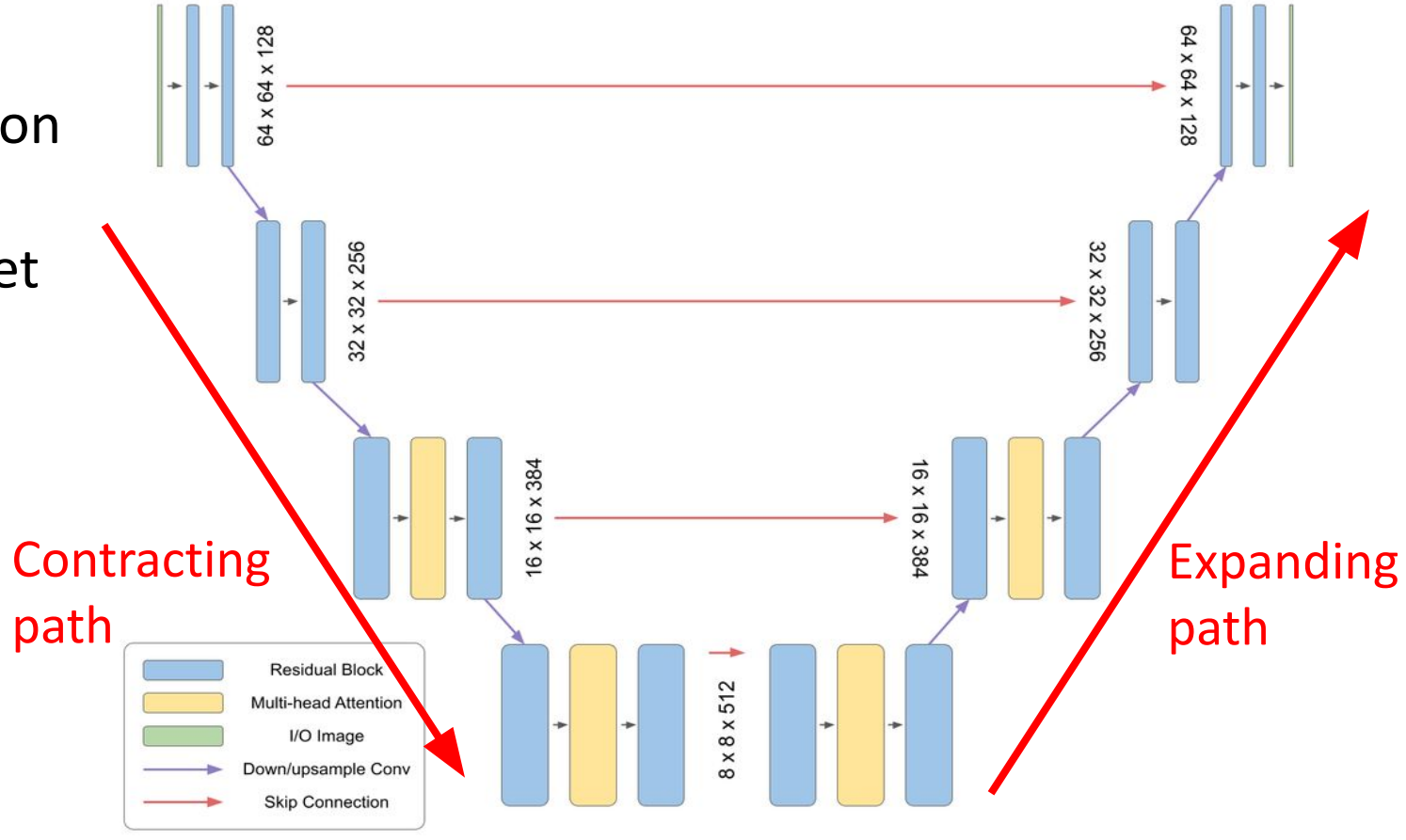
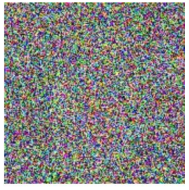


Image Source: https://www.assemblyai.com/blog/minimagen-build-your-own-imagen-text-to-image-model/

Training

Repeat:

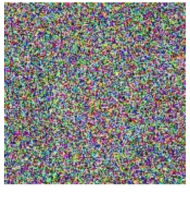
Sample:  $\mathbf{x}_0$  $\boldsymbol{\epsilon} \sim \mathcal{N}(\mathbf{0}, \mathbf{I})$

$$t \sim \text{Uniform}(\{1, \ldots, \text{T}\})$$

Diffusion kernel: 

$$\sqrt{\bar{\alpha}_t}\mathbf{x}_0 + \sqrt{1-\bar{\alpha}_t}\boldsymbol{\epsilon}$$

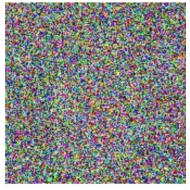
Add noise to clean image (with variance equivalent to t steps of noise)

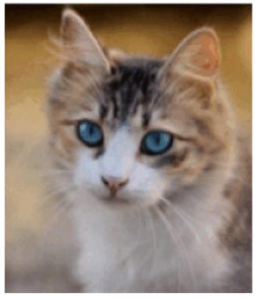
Gradient update: $\nabla_\theta \left\| \right.$  $- \boldsymbol{\epsilon}_\theta \Big($  $, t \Big) \left. \right\|^2$

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

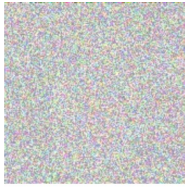
Give model noisy image, ask it to predict the added noise. Train with L2 loss between predicted noise and original noise

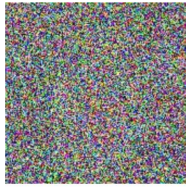
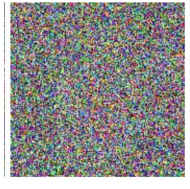
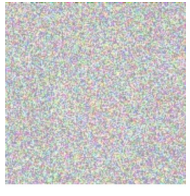
Sampling

Sample:  $\mathbf{x}_T \sim \mathcal{N}(\mathbf{0}, \mathbf{I})$



for t = 1 … T **do**:

Sample noise:  $\mathbf{z}_t \sim \mathcal{N}(\mathbf{0}, \sigma_t \mathbf{I})$

Predict next step:  $= \frac{1}{\sqrt{\alpha_t}} \left(\right.$  $- \frac{1-\alpha_t}{\sqrt{1-\bar{\alpha}_t}} \boldsymbol{\epsilon}_\theta \left(\right.$  $, t \left. \right) \left. \right) +$ 

$$x_{t-1} = \frac{1}{\sqrt{\alpha_t}} \left(\mathbf{x}_t - \frac{1-\alpha_t}{\sqrt{1-\bar{\alpha}_t}} \boldsymbol{\epsilon}_\theta(\mathbf{x}_t, t) \right) + \mathbf{z}_t$$

Summary: Training and Sample Generation

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
$$\nabla_\theta \left\| \boldsymbol{\epsilon} - \boldsymbol{\epsilon}_\theta(\sqrt{\bar{\alpha}_t}\mathbf{x}_0 + \sqrt{1-\bar{\alpha}_t}\boldsymbol{\epsilon}, t) \right\|^2$$
6: **until** converged

Algorithm 2 Sampling

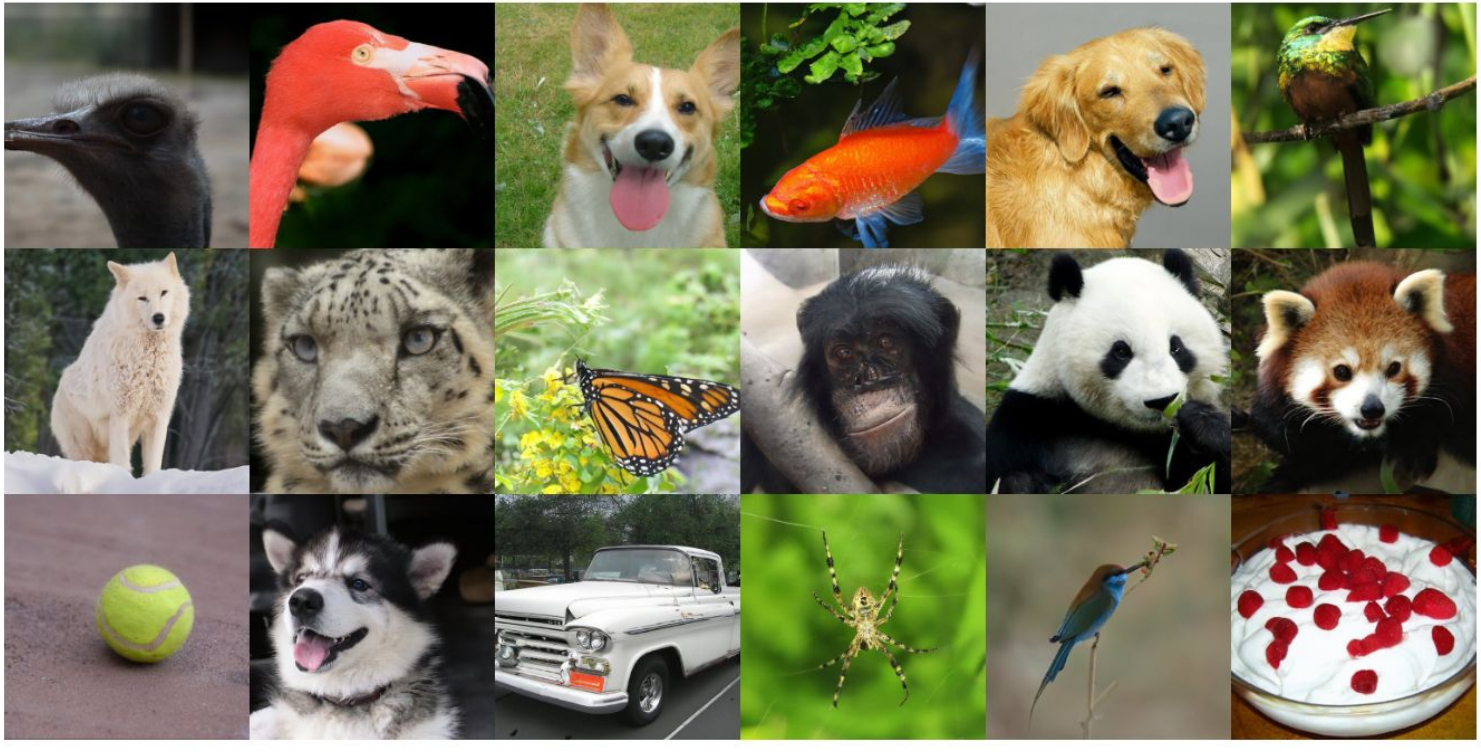
1: $\mathbf{x}_T \sim \mathcal{N}(\mathbf{0}, \mathbf{I})$
2: **for** $t = T, \dots, 1$ **do**
3: $\mathbf{z} \sim \mathcal{N}(\mathbf{0}, \mathbf{I})$
4: $\mathbf{x}_{t-1} = \frac{1}{\sqrt{\alpha_t}} \left(\mathbf{x}_t - \frac{1-\alpha_t}{\sqrt{1-\bar{\alpha}_t}} \boldsymbol{\epsilon}_\theta(\mathbf{x}_t,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

Slide credit: Kreis, Gao, Vahdat, CVPR 2022 Tutorial

Improvements: Conditional Generation

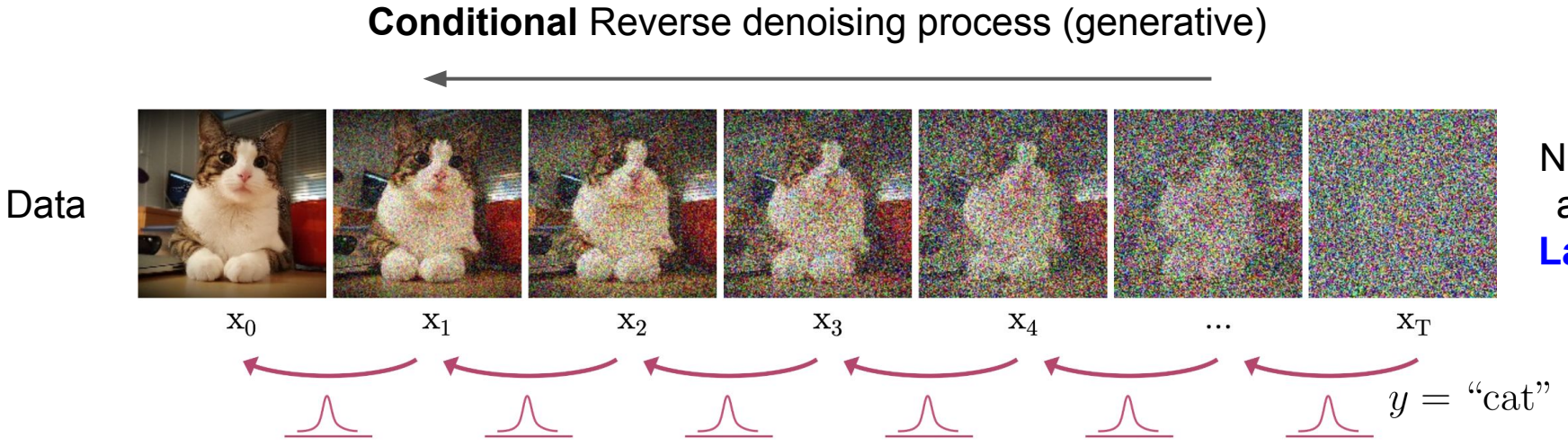
Guided Diffusion

512 x 512 ImageNet Conditioned Samples



Dhariwal and Nichol “Diffusion Models Beat GANs on Image Synthesis” NeurIPS 2021

Improvements: Conditional Generation

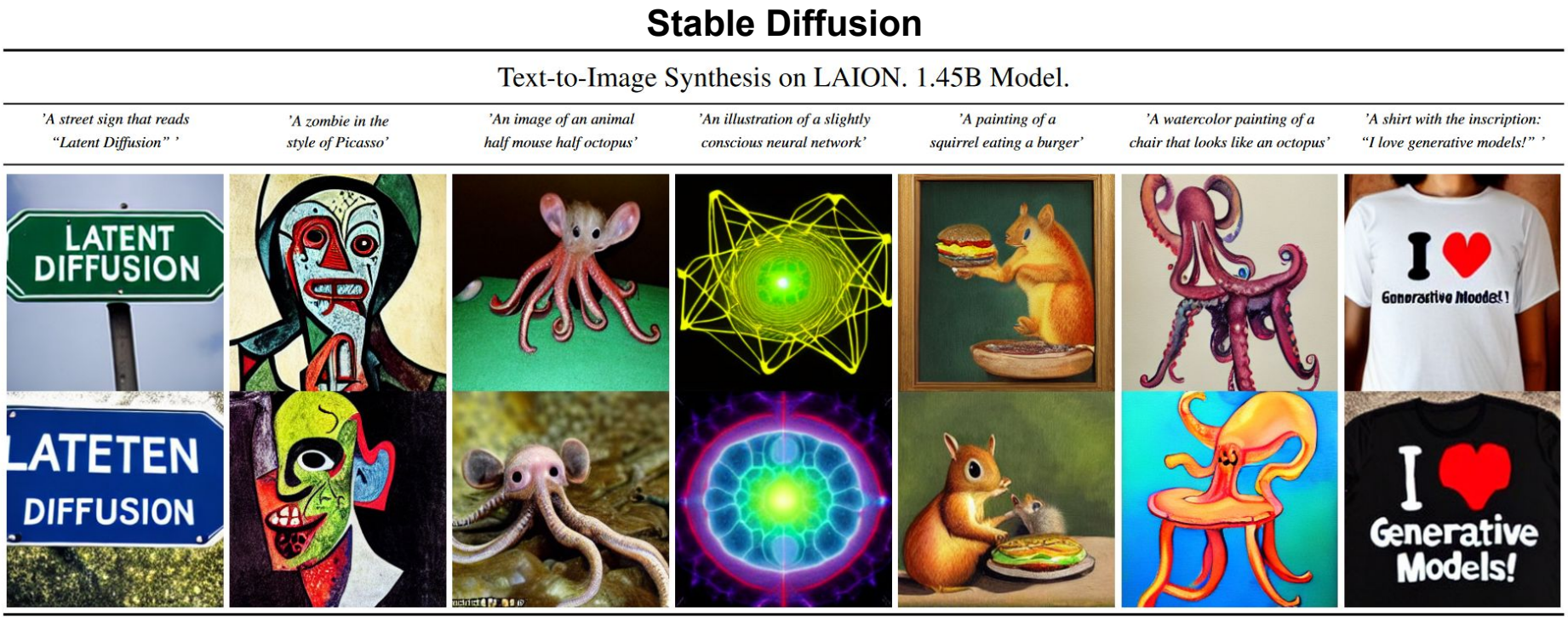


We pass the condition (y=cat) to the noise-predicting model in addition to the noisy image and step number to "control" the generation (i.e. force the generated image to be a cat instead of a dog)

Text-to-image generation: train the diffusion model to condition on text!

Improvements: Latent Diffusion Models

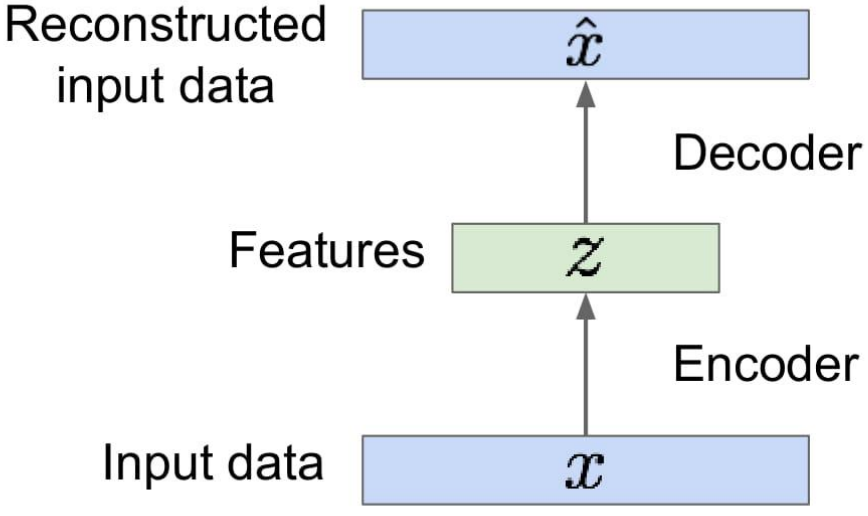
(Combined with text-image models such as CLIP)



Rombach et al. "High-Resolution Image Synthesis with Latent Diffusion Models" CVPR 2022

Improvements: Latent Diffusion Models

Use an **autoencoder** to compress the data!

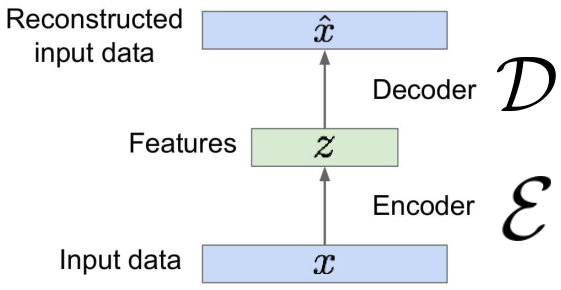


Original Image: 256x256x3

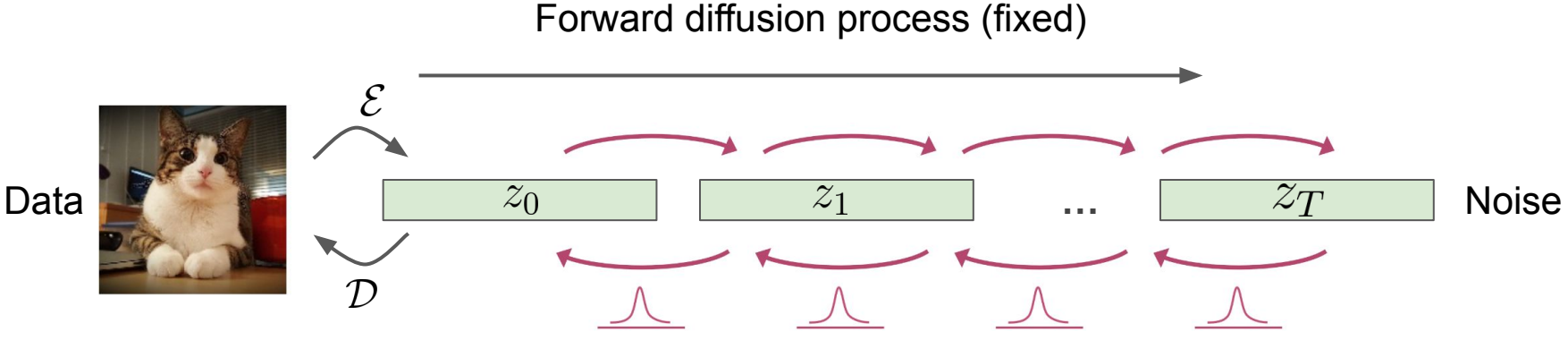
Latent Representation: 32x32x4

Reconstructed Image: 256x256x3

Improvements: Latent Diffusion Models



Use an **autoencoder** to compress the data!



Most popular text-to-image generation models nowadays are Latent Diffusion Models

For example,

- OpenAI DALLE-2 and DALLE-3, and likely all images generated from text within ChatGPT
- Midjourney
- Imagen (by Google)

Diffusion models can also be used for text-to-video generation.

- OpenAI Sora is a latent diffusion model with transformer architecture

Recap

		Pros	Cons
Autoregressive (PixelRNN, PixelCNN)	Explicit density model	Optimizes exact likelihood	Inefficient sequential generation, sample quality not the best
Variational Autoencoders (VAE)	Optimize variational lower bound on likelihood	Useful latent representation, inference queries	Sample quality not the best
Generative Adversarial Networks (GANs)	Game-theoretic approach	Great samples	Tricky and unstable to train, no inference queries
Diffusion Models	Optimize variational lower bound for denoising	Best samples! Easy to train	Inefficient sequential generation

Slide credit: Justin Johnson