Machine Learning Department
School of Computer Science
Carnegie Mellon University

Latent Diffusion Models
(and other text-to-image models)

Matt Gormley
Lecture 13
Feb. 28, 2024

Reminders

- **Homework 3: Applying and Adapting LLMs**
  - **Out: Wed, Feb 21**
  - **Due: Thu, Feb 29 at 11:59pm**
- **Homework 4:** Prompt-to-Prompt
  - **Out: Mon, Mar 11**
  - **Due: Fri, Mar 22 at 11:59pm**
- **Looking ahead…**
  - **Wed, Feb 28: Matt's last lecture**
  - **Mon, Mar 11: Yuanzhi's first lecture**

Syllabus Highlights

- **Grading**: 40% homework, 10% quizzes, 20% exam, 25% project, 5% participation
- **Exam**: in-class exam, Wed, Mar. 27
- **Homework**: 5 assignments
  - 6 grace days for homework assignments
  - Late submissions: 75% day 1, 50% day 2, 25% day 3
  - No submissions accepted after 3 days w/o extension
  - Extension requests: for emergency situations, see syllabus
- **Recitations**: Fridays, same time/place as lecture (optional, interactive sessions)
- **Readings**: required, online PDFs, recommended for after lecture
- **Technologies**:
  - Piazza (discussion),
  - Gradescope (homework),
  - Google Forms (polls),
  - Zoom (livestream),
  - Panopto (video recordings)
- **Academic Integrity**:
  - Collaboration encouraged, but must be documented
  - Solutions must always be written independently
  - No re-use of found code / past assignments
  - Severe penalties (i.e.. failure)
  - (Policies differ from 10-301/10-601)
- **Office Hours**: posted on Google Calendar on "Office Hours" page

Homework

There will be 5 homework assignments during the semester. The assignments will consist of both conceptual and programming problems.

	Main Topic	Implementation	Application Area	Type
HW0	PyTorch Primer	image classifier + Text classifier	vision + language	written + programming
HW1	Large Language Models	TransformerLM with sliding window attn.	char-level text gen	written + programming
HW2	Image Generation	diffusion model	unconditional image gen	written + programming
HW3	Adapting LLMs	GPT-2 + LoRA	instruction fine-tuning	written + programming
HW4	Multimodal Foundation Models	Prompt-to-Prompt	text-to-image generation	written + programming
HW623	(10-623 only)	read / analyze a recent research paper	genAI	video presentation

Project

- Goals:
  - Explore a generative modeling technique of your choosing
  - Deeper understanding of methods in real-world application
  - Work in teams of 3 students

CONDITIONAL IMAGE GENERATION

Image Generation

- Class-conditional generation
- Super resolution
- Image Editing
- Style transfer
- Text-to-image (TTI) generation

Figure from Razavi et al. (2019)

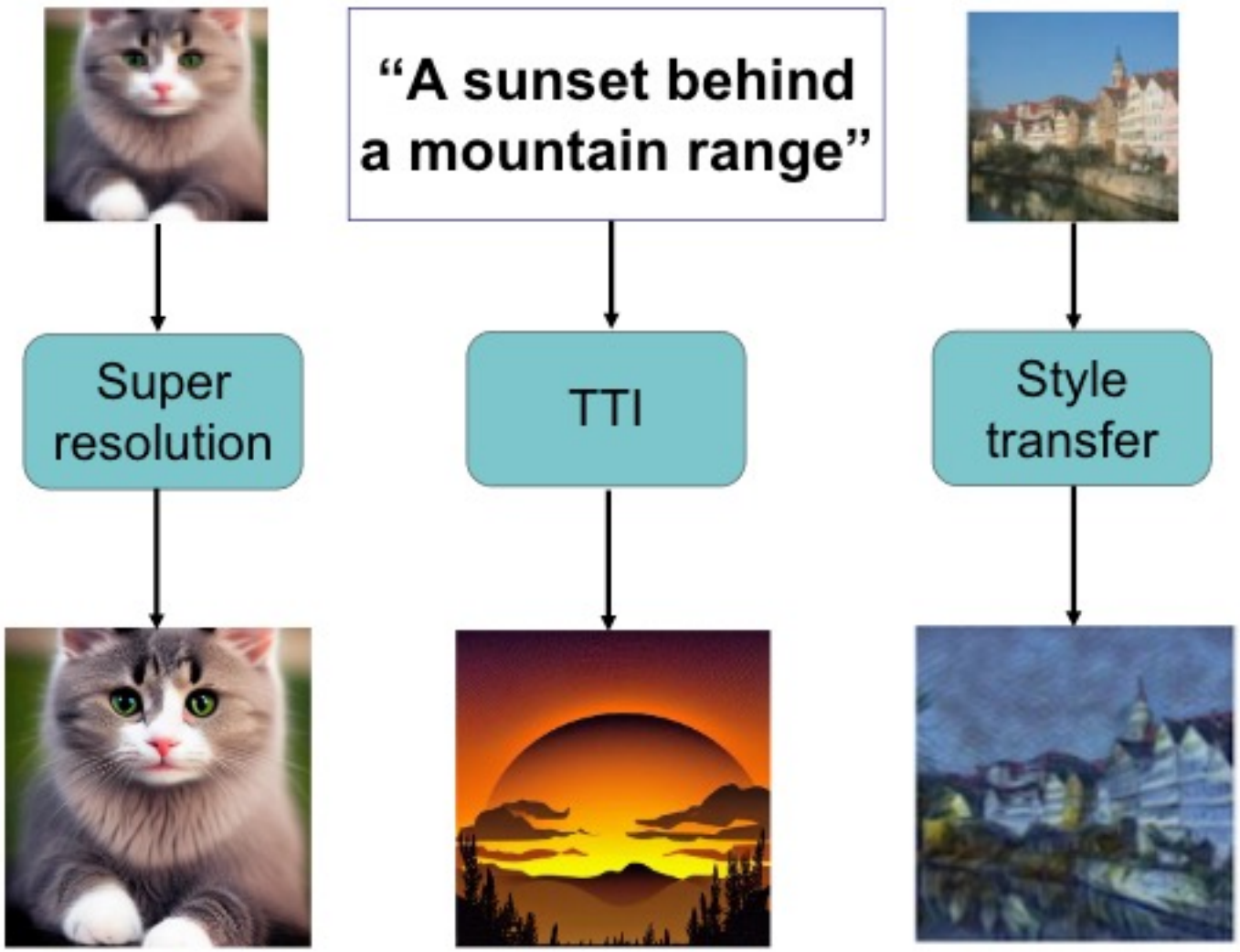


Figure from Bie et al. (2023)

Class Conditional Generation

- **Task**: Given a class label indicating the image type, sample a new image from the model with that type
- Image classification is the problem of taking in an image and predicting its label p(y|x)
- Class conditional generation is doing this in reverse p(x|y)

sea anemone

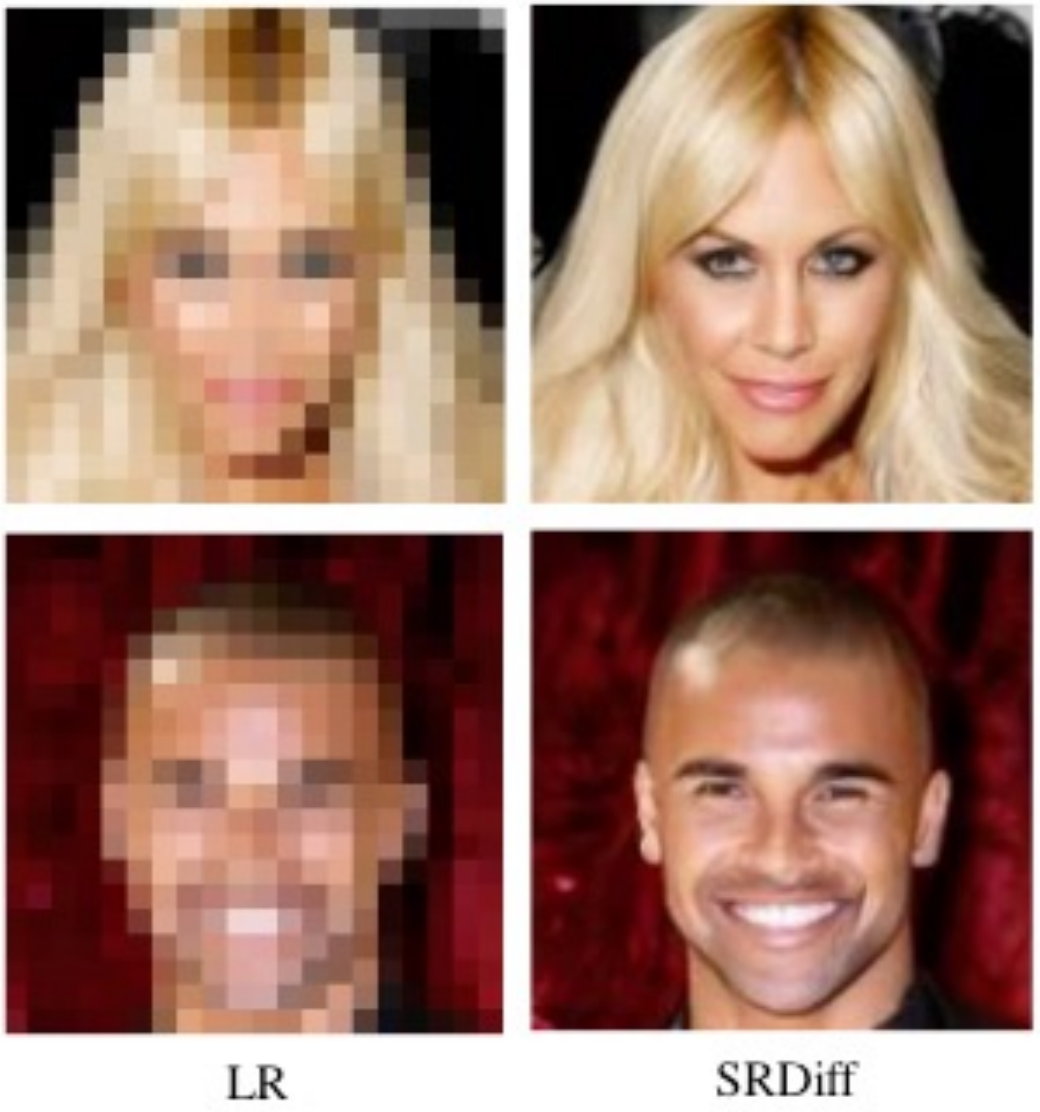
brain coral

slug

goldfinch

Figure from Razavi et al. (2019)

Super Resolution



- Given a low resolution image, generate a high resolution reconstruction of the image
- Compelling on low resolution inputs (see example to the left) but also effective on high resolution inputs

Figure from Li et al. (2021)

Image Editing

A variety of tasks involve automatic editing of an image:

- **Inpainting** fills in the (pre-specified) missing pixels
- **Colorization** restores color to a greyscale image
- **Uncropping** creates a photo-realistic reconstruction of a missing side of an image

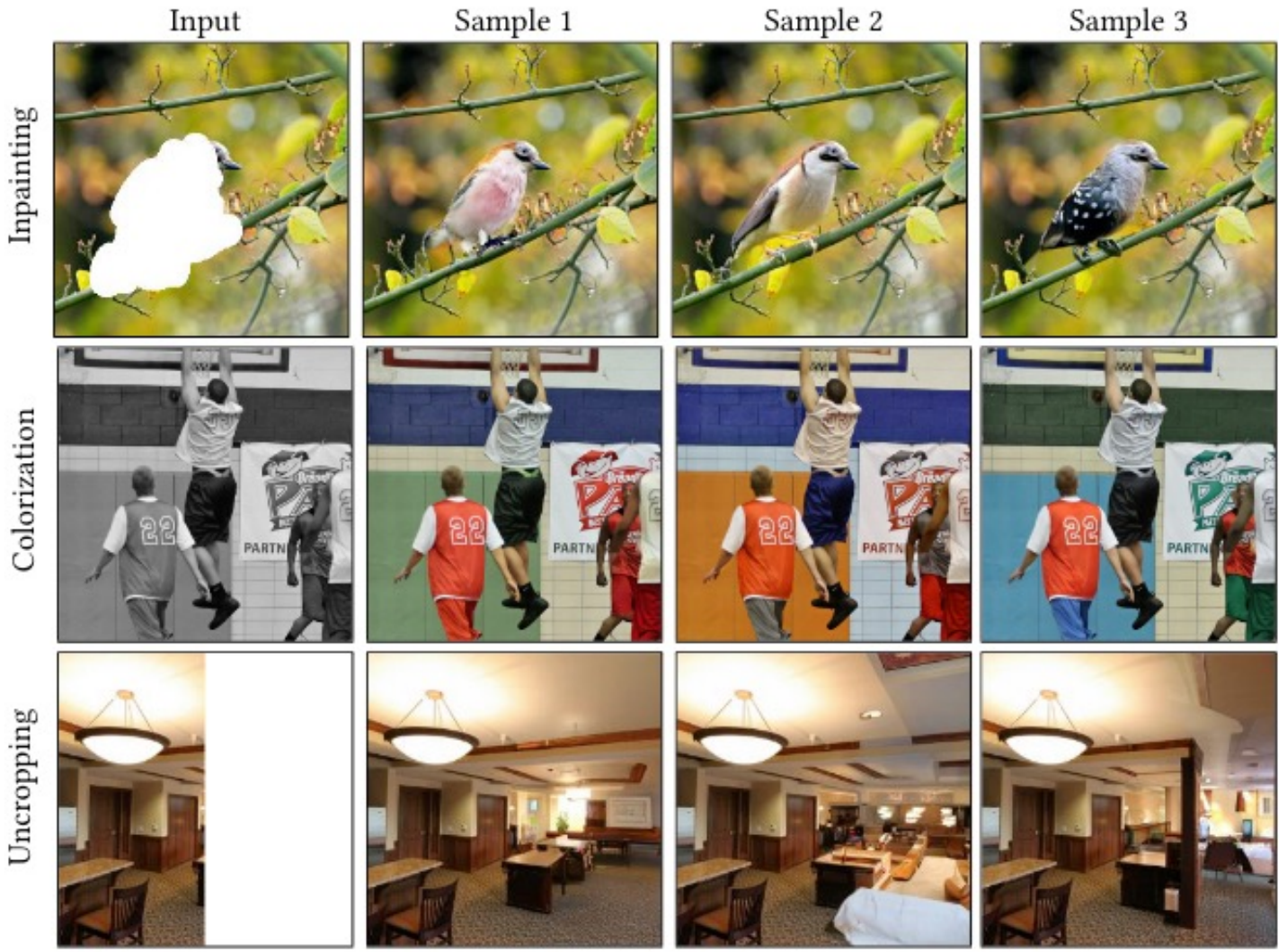


Figure from Saharia et al. (2022)

Style Transfer

- The goal of style transfer is to blend two images
- Yet, the blend should retain the semantic content of the source image presented in the style of another image

Figure from Gatys et al. (2016)

Figure 3. Images that combine the content of a photograph with the style of several well-known artworks. The images were created by finding an image that simultaneously matches the content representation of the photograph and the style representation of the artwork. The original photograph depicting the Neckarfront in Tübingen, Germany, is shown in **A** (Photo: Andreas Praefcke). The painting that provided the style for the respective generated image is shown in the bottom left corner of each panel. **B** *The Shipwreck of the Minotaur* by J.M.W. Turner, 1805. **C** *The Starry Night* by Vincent van Gogh, 1889. **D** *Der Schrei* by Edvard Munch, 1893. **E** *Femme nue assise* by Pablo Picasso, 1910. **F** *Composition VII* by Wassily Kandinsky, 1913.

Text-to-Image Generation

- Given a text description, sample an image that depicts the prompt
- The following images are samples from SDXL with refinement

Prompt: A propaganda poster depicting a cat dressed as french emperor napoleon holding a piece of cheese.

Figure from Podell et al. (2023)

Timeline: Text-to-Image Generation

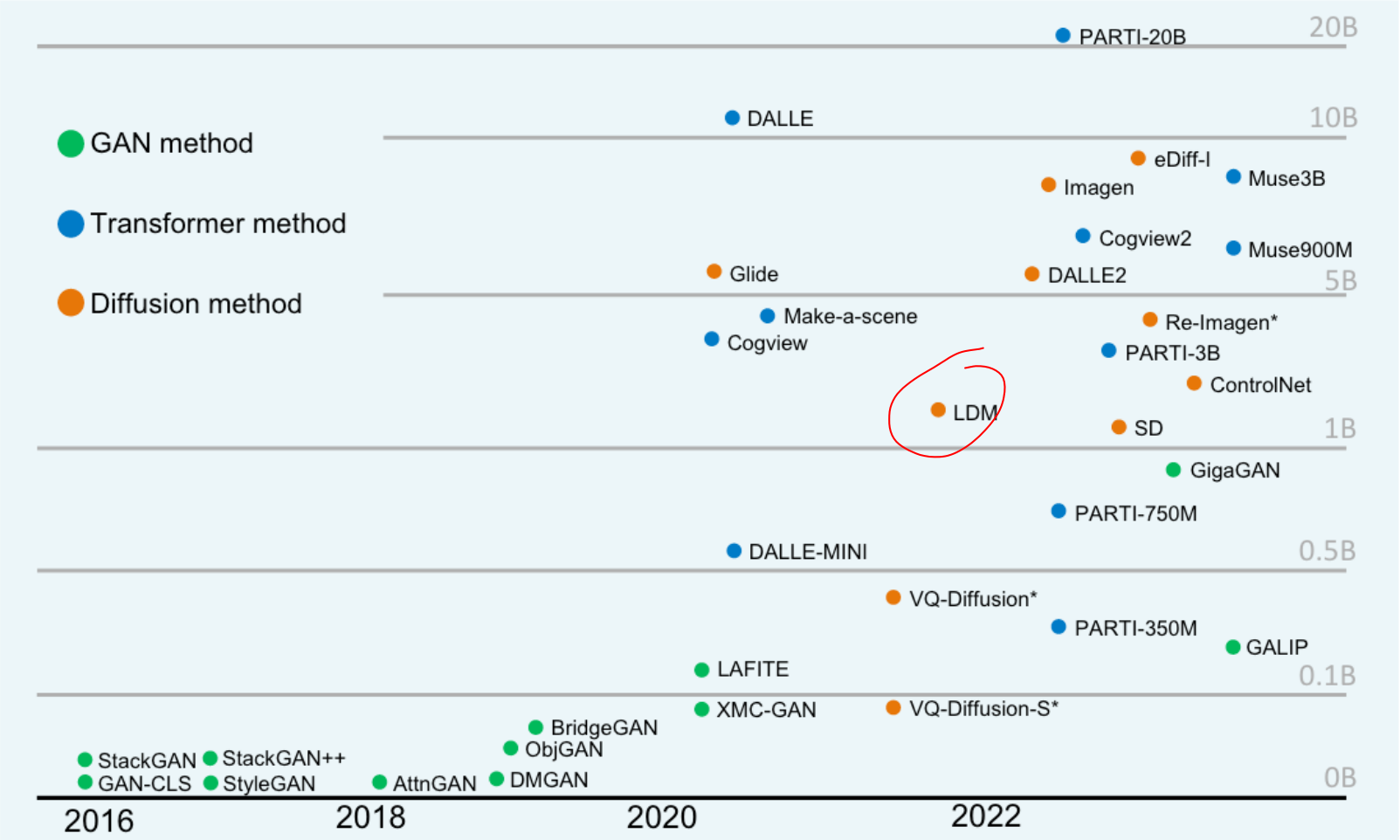


Fig. 5. Timeline of TTI model development, where green dots are GAN TTI models, blue dots are autoregressive Transformers and orange dots are Diffusion TTI models. Models are separated by their parameter, which are in general counted for all their components. Models with asterisk are calculated without the involvement of their text encoders.

Figure from Bie et al. (2023) http://arxiv.org/abs/2309.00810

Timeline: Text-to-Image Generation

A comparison of the text to image methods discussed highlighting their date published, model configuration and evaluation results. For the model type, green dot refers to the GAN model TTI, blue dot refers to the autoregressive TTI and orange dot indicates the Diffusion TTI. For evaluation metrics, IS and FID score are provided under the evaluation of MSCOCO dataset in a zero-shot fashion. The last column provides the specific model size in scale of Million(M) or Billion(B); ⋆ : no zero-shot results found, use standard results instead.

Method	Date	Model Type	Data Size	Open Source	IS evaluation	FID evaluation	Model size
AttnGAN [33]	11/2017	🟢	120K	✗	20.80	35.49 ⋆	13M
StyleGAN [34]	11/2017	🟢	120K	✗	20.80	35.49 ⋆	-
Obj-GAN [220]	09/2019	🟢	120K	✓	24.09	36.52 ⋆	34M
Control-GAN [221]	09/2019	🟢	120K	✓	23.61	33.10 ⋆	-
DM-GAN [35]	04/2019	🟢	120K	✓	32.32	27.34 ⋆	21M
XMC-GAN [165]	01/2021	🟢	120K	✗	30.45	9.33 ⋆	90M
LAFITE [44]	11/2021	🟢	-	✓	26.02	26.94	150M
Retreival-GAN [208]	08/2022	🟢	120K	✗	29.33	9.13 ⋆	25M
GigaGAN [46]	01/2023	🟢	-	✗	-	10.24	650M
GALIP [45]	03/2023	🟢	3M-12M	✓	-	12.54	240M
DALLE [39]	02/2021	🔵	250M	✗	-	27.5	12B
Cogview [189]	06/2021	🔵	300M	✓	-	27.1	4B
Make-A-Scene	03/2022	🔵	35M	✗	-	11.84	4B
Cogview2 [43]	05/2022	🔵	300M	✓	-	24.0	6B
PARTI-350M [5]	06/2022	🔵	∼1000M	✗	-	14.10	350M
PARTI-20B [5]	06/2022	🔵	∼1000M	✗	-	7.23	20B
DALLE-mini [187]	07/2021	🔵	250M	✗	-	-	∼500M
MUSE-3B [31]	03/2023	🔵	∼1000M	✗	-	7.88	7.6B
GLIDE [40]	12/2021	🟠	250M	✓	-	12.24	5B
VQ-diffusion-F [68]	11/2021	🟠	>7M	✓	-	13.86 ⋆	370M
DALLE-2 [4]	04/2022	🟠	250M	✗	-	10.39	5.2B
Imagen [30]	05/2022	🟠	∼860M	✗	-	7.27	7.6B
LDM [3]	08/2022	🟠	400M	✓	30.29	12.63	1.45B
eDiff-I [197]	11/2022	🟠	1000M	✗	-	6.95	9B
Shift Diffusion[158]	08/2022	🟠	900M	✓	-	10.88	-
Re-Imagen[203]	09/2022	🟠	50M	✗	-	6.88	∼8B
ControlNet [159]	03/2023	🟠	-	✓	-	-	∼2.2B

Figure from Bie et al. (2023) http://arxiv.org/abs/2309.00810

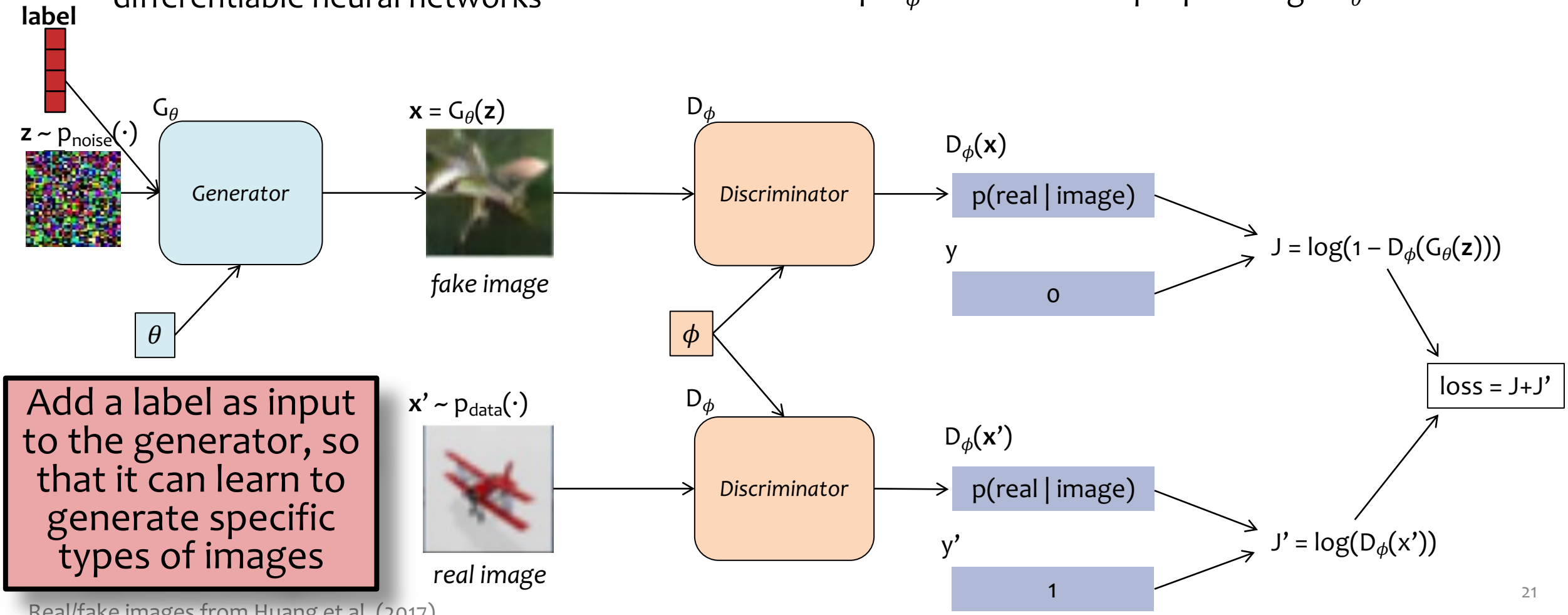
TEXT-TO-IMAGE: GANS

Class-conditional GANs

- Objective function is a simple differentiable function
- We chose G and D to be differentiable neural networks

Training alternates between:

- Keep G_θ fixed and backprop through D_ϕ
- Keep D_ϕ fixed and backprop through G_θ



Add a label as input to the generator, so that it can learn to generate specific types of images

Real/fake images from Huang et al. (2017)

Generative adversarial text to image synthesis

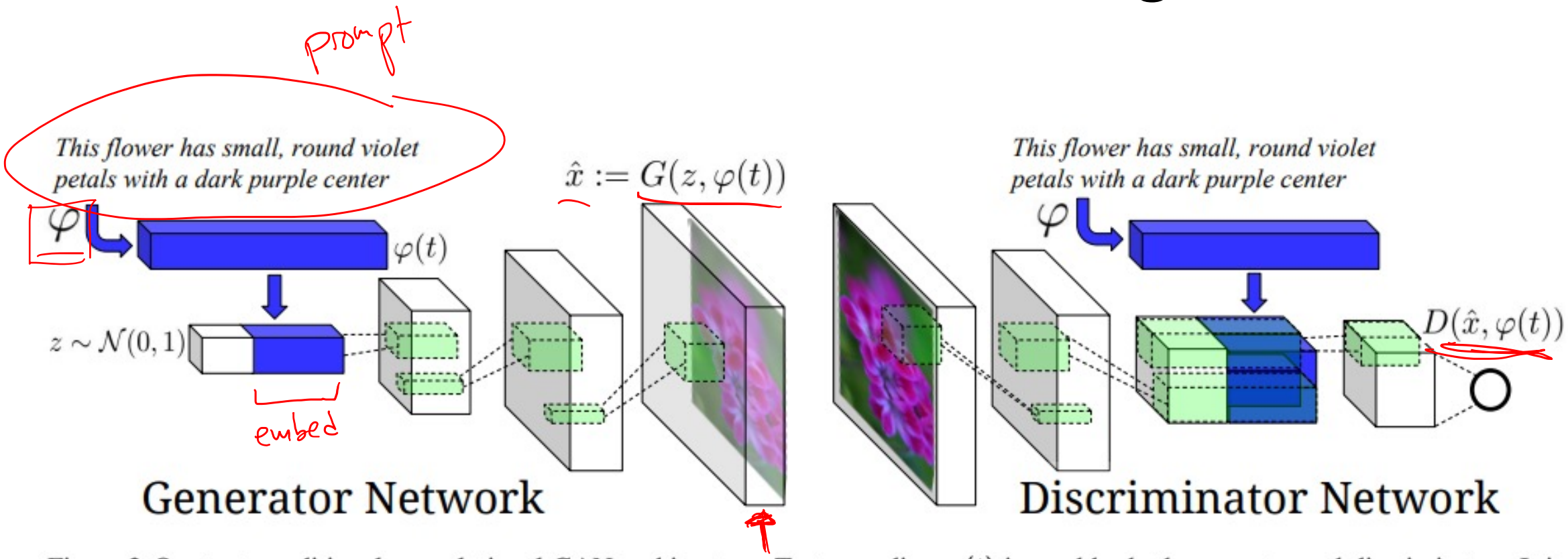


Figure 2. Our text-conditional convolutional GAN architecture. Text encoding $\varphi(t)$ is used by both generator and discriminator. It is projected to a lower-dimensions and depth concatenated with image feature maps for further stages of convolutional processing.

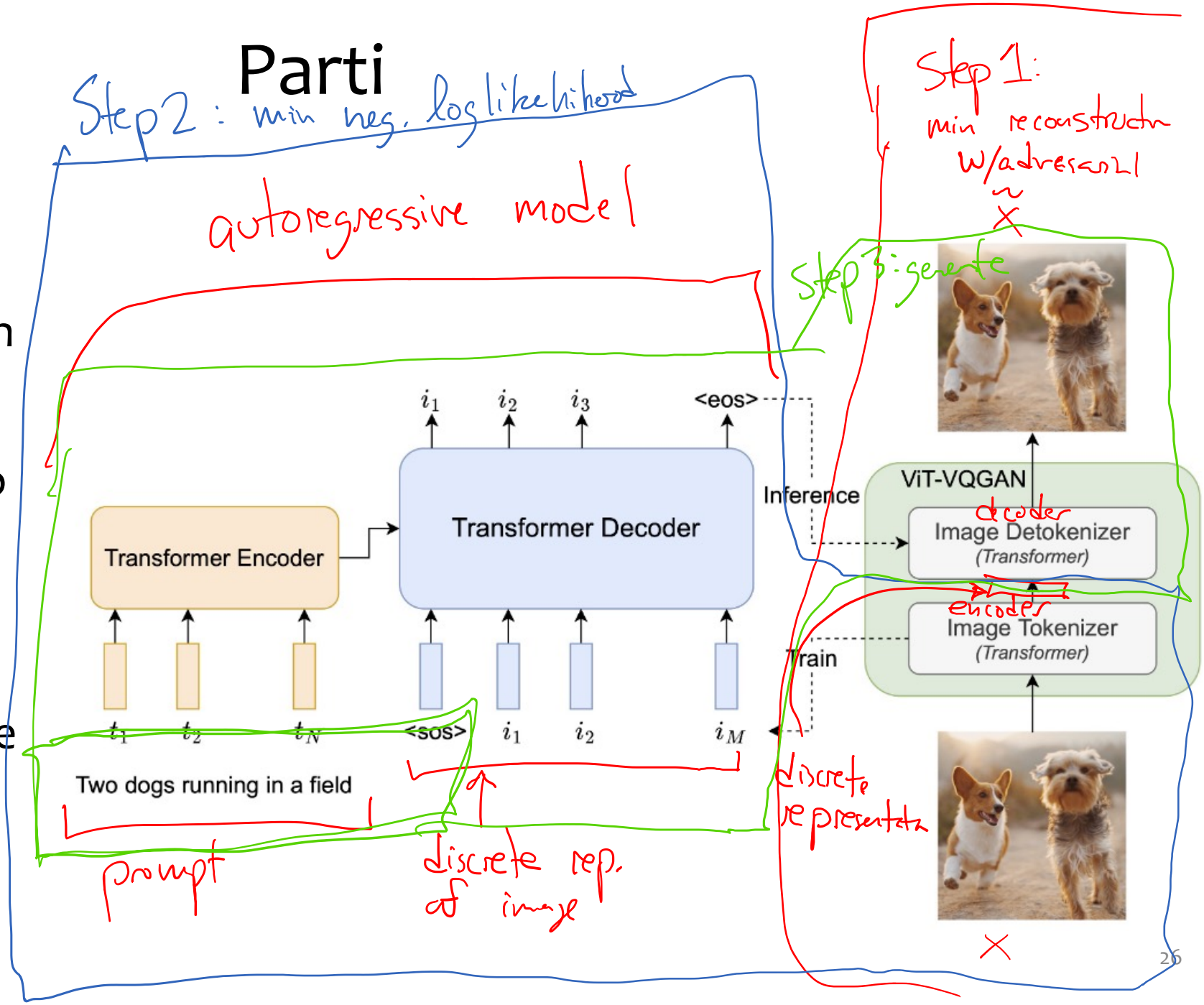
Figure from Reed et al. (2016)

TEXT-TO-IMAGE: AUTOREGRESSIVE MODELS

Parti

The Pathways Autoregressive Text-to-Image (Parti) model:

- treat image generation as a sequence-to-sequence problem
- text prompt is input to encoder
- sequence of image tokens is output of decoder
- ViT-VQGAN takes in the image tokens and generates a high-quality image



TEXT-TO-IMAGE: DIFFUSION MODELS

CLIP (background for Dall-E 2)

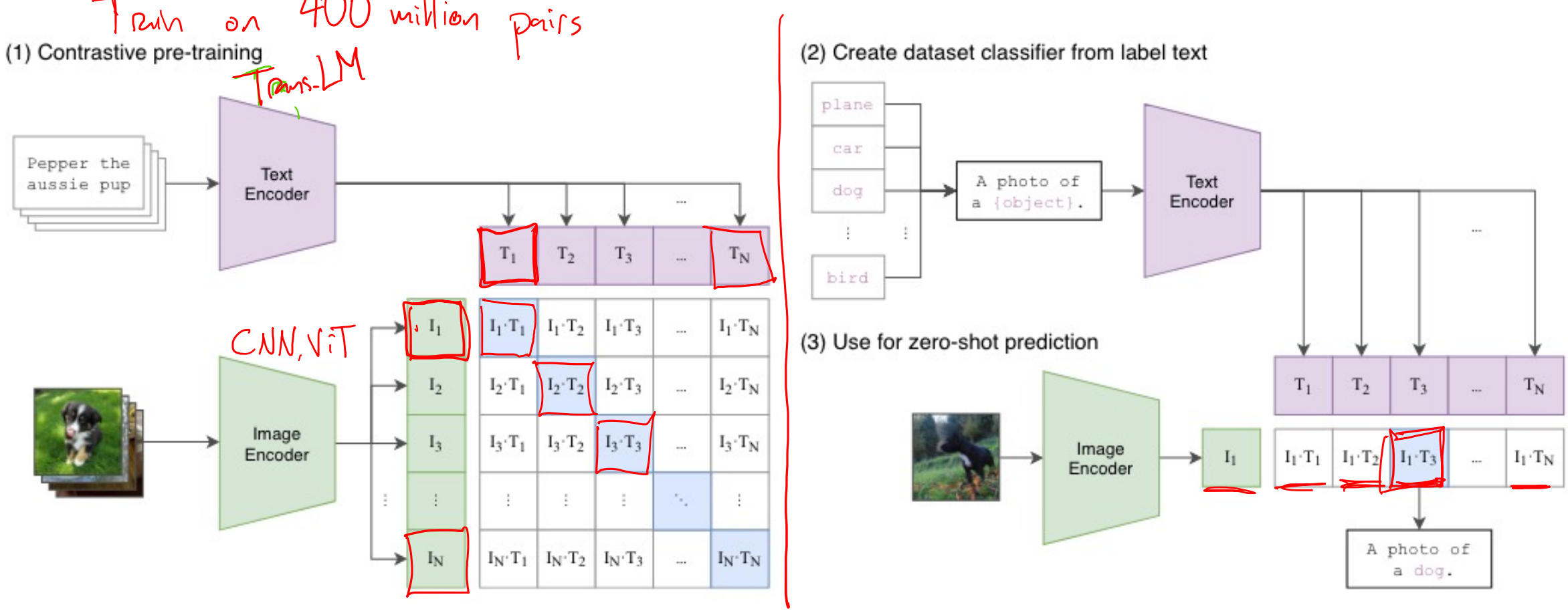


Figure 1. Summary of our approach. While standard image models jointly train an image feature extractor and a linear classifier to predict some label, CLIP jointly trains an image encoder and a text encoder to predict the correct pairings of a batch of (image, text) training examples. At test time the learned text encoder synthesizes a zero-shot linear classifier by embedding the names or descriptions of the target dataset's classes.

Figure from Radford et al. (2021)

Dall-E 2

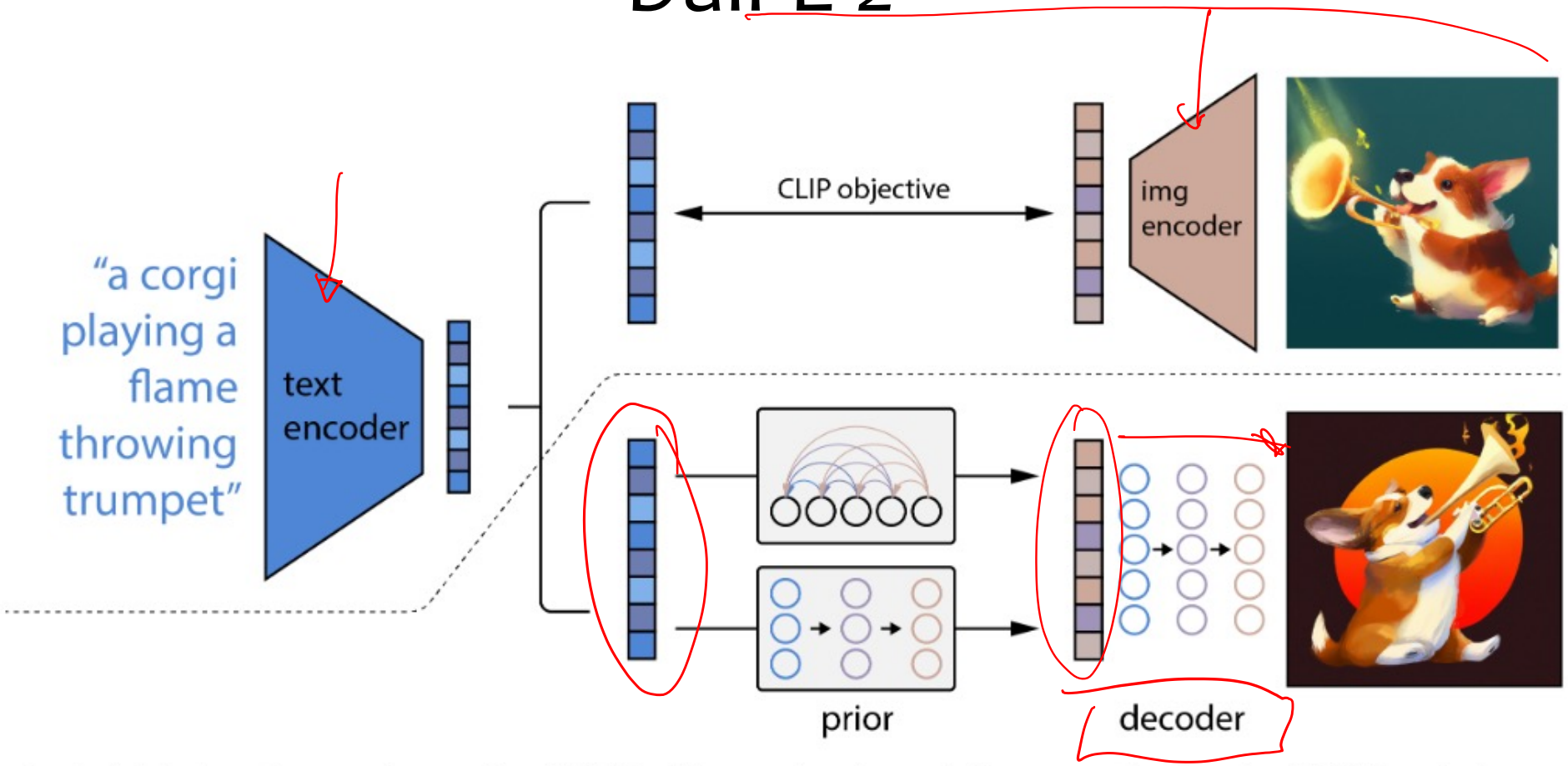


Figure 2: A high-level overview of unCLIP. Above the dotted line, we depict the CLIP training process, through which we learn a joint representation space for text and images. Below the dotted line, we depict our text-to-image generation process: a CLIP text embedding is first fed to an autoregressive or diffusion prior to produce an image embedding, and then this embedding is used to condition a diffusion decoder which produces a final image. Note that the CLIP model is frozen during training of the prior and decoder.

Figure from Ramesh et al. (2022)

Imagen

- Imagen uses a text-to-image diffusion model coupled with a super-resolution diffusion model
- All the models operate in pixel space
- While effective, the compute requirements are very high

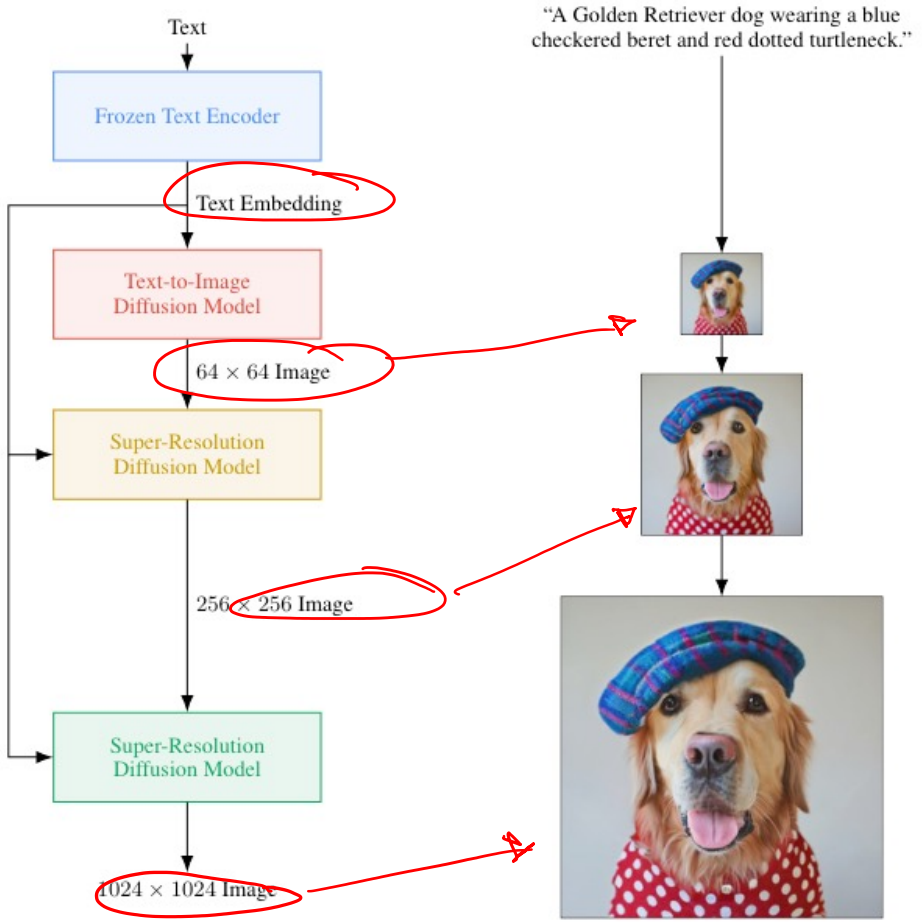


Figure A.4: Visualization of Imagen. Imagen uses a frozen text encoder to encode the input text into text embeddings. A conditional diffusion model maps the text embedding into a 64×64 image. Imagen further utilizes text-conditional super-resolution diffusion models to upsample the image, first $64 \times 64 \rightarrow 256 \times 256$, and then $256 \times 256 \rightarrow 1024 \times 1024$.

LATENT DIFFUSION MODEL (LDM)

Latent Diffusion Model

Motivation:

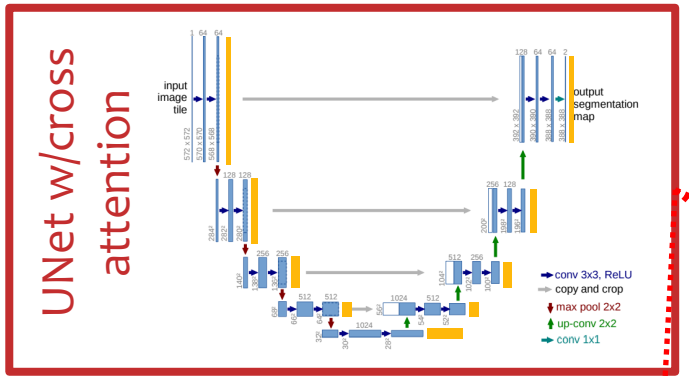
- diffusion models typically operate in pixel space
- yet, training typically takes hundreds of GPU days
  - 150 – 1000 V100 days [Guided Diffusion] (Dhariwal & Nichol, 2021)
  - 256 TPU-v4s for 4 days = 1000 TPU days [Imagen] (Sharia et al., 2022)
- inference is also slow
  - 50k samples in 5 days on A100 GPU [Guided Diffusion] (Dhariwal & Nichol, 2021)
  - 15 seconds per image

Key Idea:

- train an autoencoder (i.e. encoder-decoder model) that learns an efficient latent space that is perceptually equivalent to the data space
- keeping the autoencoder fixed, train a diffusion model on the latent representations of real images z_0 = encoder(x)
  - forward model: latent representation z_0 → noise z_T
  - reverse model: noise z_T → latent representation z_0
- to generate an image:
  - sample noise z_T
  - apply reverse diffusion model to obtain a latent representation z_0
  - decode the latent representation to an image x
- condition on prompt via cross attention in latent space

Latent Diffusion Model

Latent Diffusion Model (LDM)



prompt space

latent space

pixel space

$\widehat{\mathbf{y}}$

τ_θ

LLM

$\mathbf{y}$

orange cat

$p_\theta(\mathbf{z}_T)$

$p_\theta(\mathbf{z}_{T-1} \mid \mathbf{z}_T, \widehat{\mathbf{y}})$

$p_\theta(\mathbf{z}_t \mid \mathbf{z}_{t+1}, \widehat{\mathbf{y}})$

$p_\theta(\mathbf{z}_0 \mid \mathbf{z}_1, \widehat{\mathbf{y}})$

z_T z_{T-1} ... z_{t+1} z_t ... z_1 z

$\mathcal{D}$

Decoder

$\widetilde{\mathbf{x}}$

$\sim N(0, I)$

$q_\phi(\mathbf{z}_T \mid \mathbf{z}_{T-1})$

$q_\phi(\mathbf{z}_{t+1} \mid \mathbf{z}_t)$

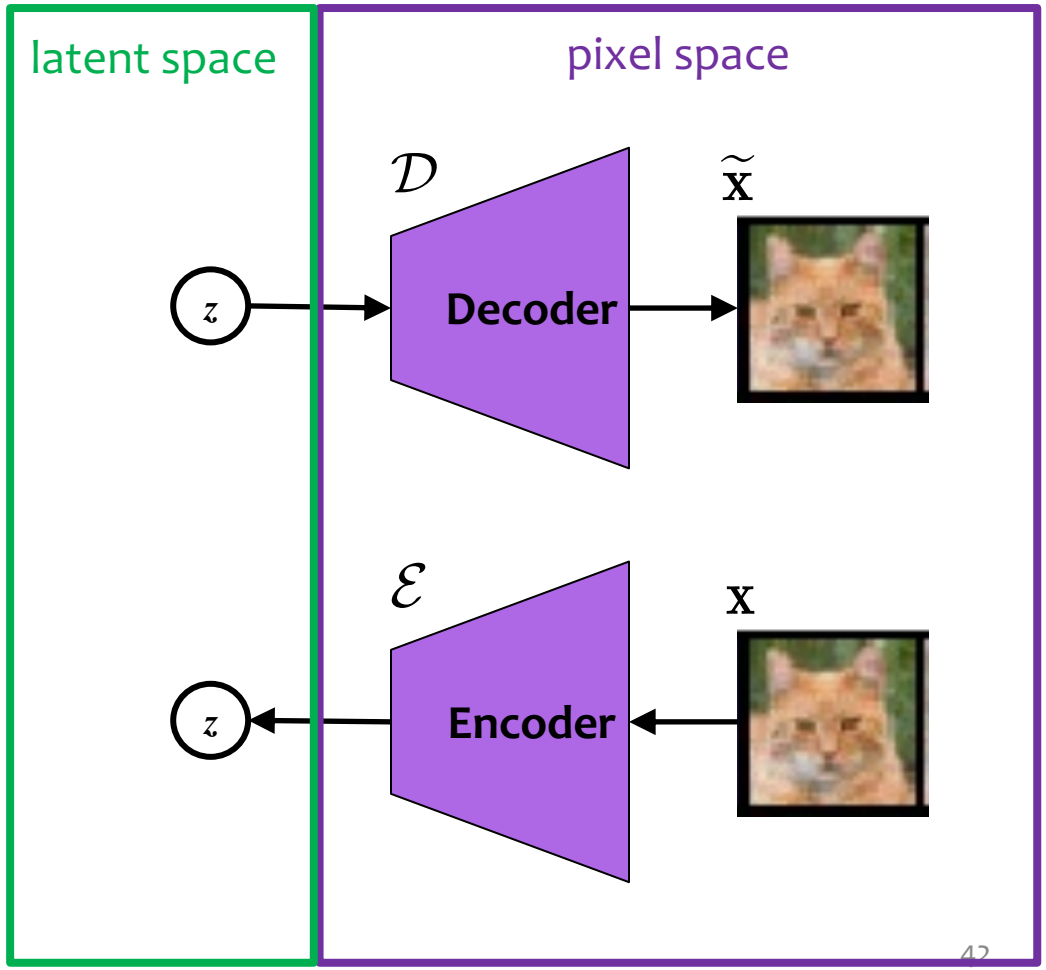
$q_\phi(\mathbf{z}_1 \mid \mathbf{z}_0)$

$\mathcal{E}$

Encoder

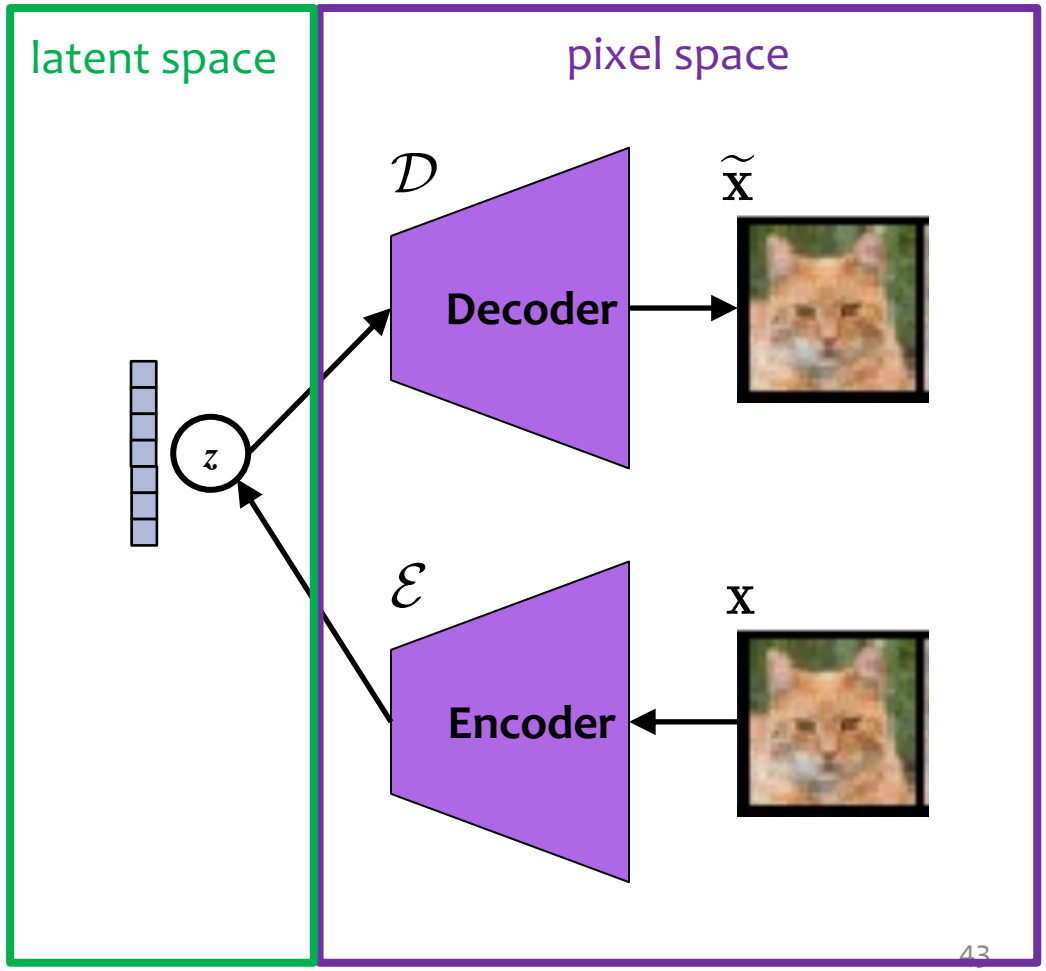
$\mathbf{x}$

LDM: Autoencoder



LDM: Autoencoder

- The autoencoder is chosen so that it can project high dimensional images (e.g. 1024x1024) down to low dimensional **latent space** and **faithfully** project back up to **pixel space**
- The original LDM paper considers two options:
  1. a VAE-like model (regularizes the noise towards a Gaussian)
  2. a VQGAN (performs vector quantization in the decoder; i.e., it uses a discrete codebook)
- This model is trained ahead of time just on raw images (no text prompts) and then frozen
- The frozen encoder-decoder can be reused for all subsequent LDM training

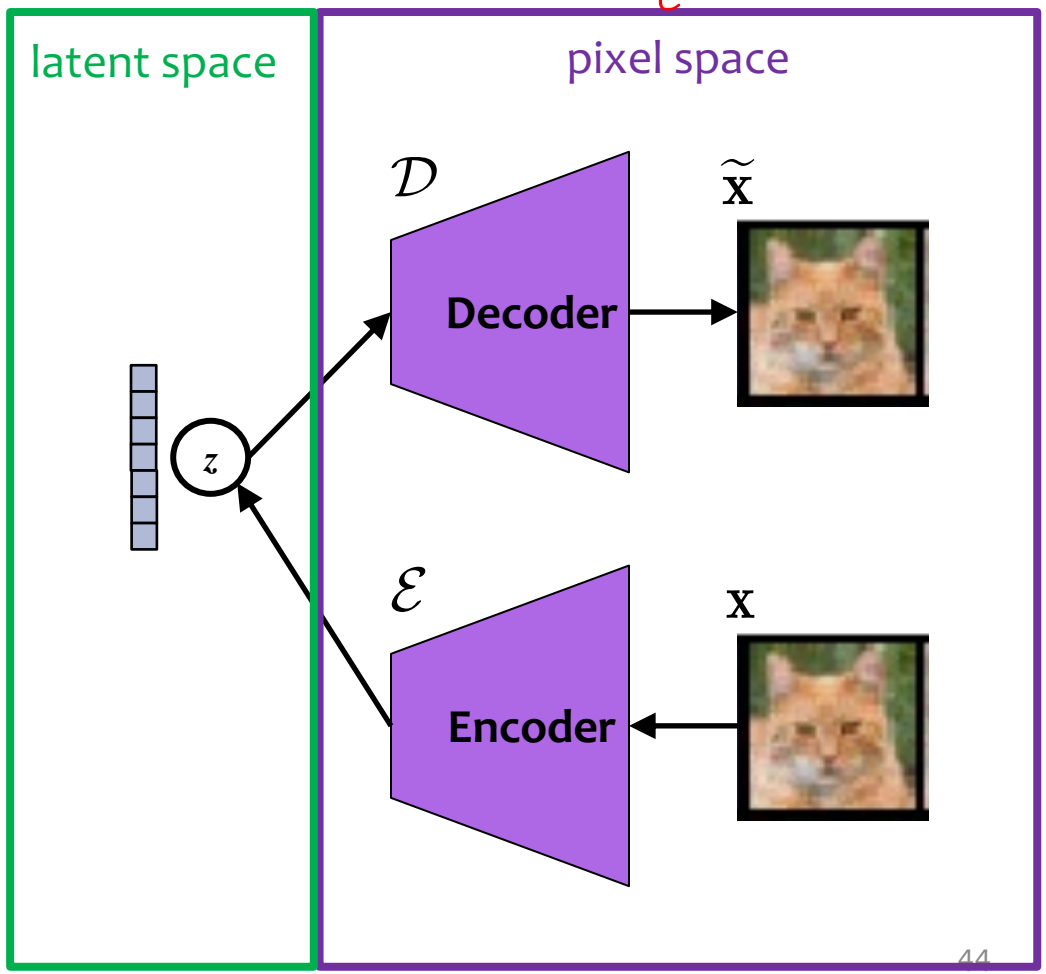


LDM: Autoencoder

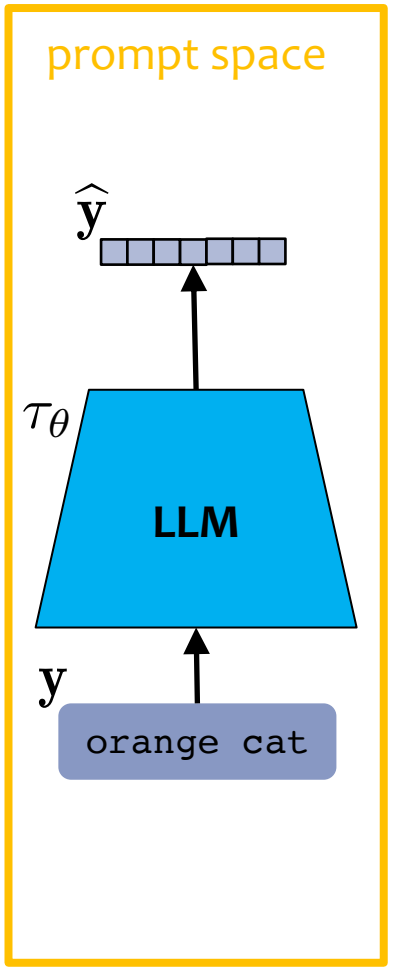
- After trying a zoo of autoencoder options, the original paper picked one that offered a good level of compression without much loss of information

f	$\lvert\mathcal{Z}\rvert$	c	R-FID ↓	R-IS ↑	PSNR ↑	PSIM ↓	SSIM ↑
16 *VQGAN* [23]	16384	256	4.98	–	19.9 ±3.4	1.83 ±0.42	0.51 ±0.18
16 *VQGAN* [23]	1024	256	7.94	–	19.4 ±3.3	1.98 ±0.43	0.50 ±0.18
8 *DALL-E* [66]	8192	-	32.01	–	22.8 ±2.1	1.95 ±0.51	0.73 ±0.13
32	16384	16	31.83	40.40 ±1.07	17.45 ±2.90	2.58 ±0.48	0.41 ±0.18
16	16384	8	5.15	144.55 ±3.74	20.83 ±3.61	1.73 ±0.43	0.54 ±0.18
8	16384	4	1.14	201.92 ±3.97	23.07 ±3.99	1.17 ±0.36	0.65 ±0.16
8	256	4	1.49	194.20 ±3.87	22.35 ±3.81	1.26 ±0.37	0.62 ±0.16
4	8192	3	0.58	224.78 ±5.35	27.43 ±4.26	0.53 ±0.21	0.82 ±0.10
4†	8192	3	1.06	221.94 ±4.58	25.21 ±4.17	0.72 ±0.26	0.76 ±0.12
4	256	3	0.47	223.81 ±4.58	26.43 ±4.22	0.62 ±0.24	0.80 ±0.11
2	2048	2	0.16	232.75 ±5.09	30.85 ±4.12	0.27 ±0.12	0.91 ±0.05
2	64	2	0.40	226.62 ±4.83	29.13 ±3.46	0.38 ±0.13	0.90 ±0.05
32	KL	64	2.04	189.53 ±3.68	22.27 ±3.93	1.41 ±0.40	0.61 ±0.17
32	KL	16	7.3	132.75 ±2.71	20.38 ±3.56	1.88 ±0.45	0.53 ±0.18
16	KL	16	0.87	210.31 ±3.97	24.08 ±4.22	1.07 ±0.36	0.68 ±0.15
16	KL	8	2.63	178.68 ±4.08	21.94 ±3.92	1.49 ±0.42	0.59 ±0.17
8	KL	4	0.90	209.90 ±4.92	24.19 ±4.19	1.02 ±0.35	0.69 ±0.15
4	KL	3	0.27	227.57 ±4.89	27.53 ±4.54	0.55 ±0.24	0.82 ±0.11
2	KL	2	0.086	232.66 ±5.16	32.47 ±4.19	0.20 ±0.09	0.93 ±0.04

Table 8. Complete autoencoder zoo trained on OpenImages, evaluated on ImageNet-Val. † denotes an attention-free autoencoder.

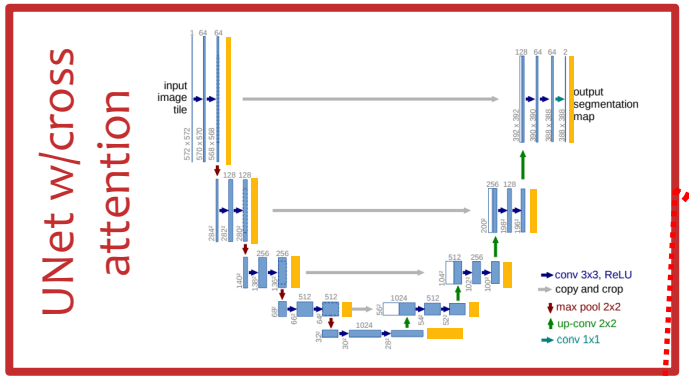


LDM: the Prompt Model



- The prompt model is just a Transformer LM
- We learn its parameters alongside the diffusion model
- The goal is to build up good representations of the text prompts such that they inform the latent diffusion process

LDM: with DDPM



prompt space

latent space

$\hat{\mathbf{y}} = \tau_\theta(y)$

τ_θ

LLM

$\mathbf{y}$

orange cat

$p_\theta(\mathbf{z}_{T-1} \mid \mathbf{z}_T, \hat{\mathbf{y}})$

$p_\theta(\mathbf{z}_t \mid \mathbf{z}_{t+1}, \hat{\mathbf{y}})$

$p_\theta(\mathbf{z}_0 \mid \mathbf{z}_1, \hat{\mathbf{y}})$

$p_\theta(\mathbf{z}_T)$

z_T z_{T-1} ... z_{t+1} z_t ... z_1 z

$q_\phi(\mathbf{z}_T \mid \mathbf{z}_{T-1})$

$q_\phi(\mathbf{z}_{t+1} \mid \mathbf{z}_t)$

$q_\phi(\mathbf{z}_1 \mid \mathbf{z}_0)$

LDM: with DDPM

Noise schedule:

We choose α_t to follow a fixed schedule s.t. $q_\phi(\mathbf{x}_T) \sim \mathcal{N}(\mathbf{0}, \mathbf{I})$, just like $p_\theta(\mathbf{x}_T)$.

Here we let $\mathbf{z}_0 = \mathbf{z}$, the output of the encoder from our autoencoder

Forward Process:

$$q_\phi(\mathbf{z}_{1:T}) = q(\mathbf{z}_0) \prod_{t=1}^{T} q_\phi(\mathbf{z}_t \mid \mathbf{z}_{t-1})$$

$$q(\mathbf{z}_0) = \text{data distribution}$$

$$q_\phi(\mathbf{z}_t \mid \mathbf{z}_{t-1}) \sim \mathcal{N}(\sqrt{\alpha_t}\mathbf{z}_{t-1}, (1-\alpha_t)\mathbf{I})$$

(Learned) Reverse Process:

$$p_\theta(\mathbf{z}_{1:T}) = p_\theta(\mathbf{z}_T) \prod_{t=1}^{T} p_\theta(\mathbf{z}_{t-1} \mid \mathbf{z}_t, \tau_\theta(y))$$

$$p_\theta(\mathbf{z}_T) \sim \mathcal{N}(\mathbf{0}, \mathbf{I})$$

$$p_\theta(\mathbf{z}_{t-1} \mid \mathbf{z}_t, \tau_\theta(y)) \sim \mathcal{N}(\mu_\theta(\mathbf{z}_t, t, \tau_\theta(y)), \mathbf{\Sigma}_\theta(\mathbf{z}_t, t))$$

LDM: with DDPM

Noise schedule:

We choose α_t to follow a fixed schedule s.t. $q_\phi(\mathbf{x}_T) \sim \mathcal{N}(\mathbf{0}, \mathbf{I})$, just like $p_\theta(\mathbf{x}_T)$.

Here we let $\mathbf{z}_0 = \mathbf{z}$, the output of the encoder from our autoencoder

Forward Process:

$$q_\phi(\mathbf{z}_{1:T}) = q(\mathbf{z}_0) \prod_{t=1}^{T} q_\phi(\mathbf{z}_t \mid \mathbf{z}_{t-1})$$

$$q(\mathbf{z}_0) = \text{data}$$

$$q_\phi(\mathbf{z}_t \mid \mathbf{z}_{t-1}) \sim \mathcal{N}(\sqrt{$$

Question: How do we define the mean to condition on the prompt representation?

(Learned) Reverse Process:

$$p_\theta(\mathbf{z}_{1:T}) = p_\theta(\mathbf{z}_T) \prod_{t=1}^{T} p_\theta(\mathbf{z}_{t-1} \mid \mathbf{z}_t, \tau_\theta(y))$$

$$p_\theta(\mathbf{z}_T) \sim \mathcal{N}(\mathbf{0}, \mathbf{I})$$

$$p_\theta(\mathbf{z}_{t-1} \mid \mathbf{z}_t, \tau_\theta(y)) \sim \mathcal{N}(\mu_\theta(\mathbf{z}_t, t, \tau_\theta(y)), \mathbf{\Sigma}_\theta(\mathbf{z}_t, t))$$

Properties of forward and *exact* reverse process

Recall...

Property #1:

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

$$\text{where } \bar{\alpha}_t = \prod_{s=1}^{t} \alpha_s$$

$\Rightarrow$ we can sample $\mathbf{x}_t$ from $\mathbf{x}_0$ at any timestep t efficiently in closed form

$\Rightarrow$ $\mathbf{x}_t = \sqrt{\bar{\alpha}_t}\mathbf{x}_0 + (1-\bar{\alpha}_t)\boldsymbol{\epsilon}$ where $\boldsymbol{\epsilon} \sim \mathcal{N}(\mathbf{0}, \mathbf{I})$

Property #2: Estimating $q(\mathbf{x}_{t-1} \mid \mathbf{x}_t)$ is intractable because of its dependence on $q(\mathbf{x}_0)$. However, conditioning on $\mathbf{x}_0$ we can efficiently work with:

$$q(\mathbf{x}_{t-1} \mid \mathbf{x}_t, \mathbf{x}_0) = \mathcal{N}(\tilde{\mu}_q(\mathbf{x}_t, \mathbf{x}_0), \sigma_t^2 \mathbf{I})$$

$$\begin{aligned}\text{where } \tilde{\mu}_q(\mathbf{x}_t, \mathbf{x}_0) &= \frac{\sqrt{\bar{\alpha}_t}(1-\alpha_t)}{1-\bar{\alpha}_t}\mathbf{x}_0 + \frac{\sqrt{\alpha_t}(1-\bar{\alpha}_t)}{1-\bar{\alpha}_t}\mathbf{x}_t \\ &= \alpha_t^{(0)}\mathbf{x}_0 + \alpha_t^{(t)}\mathbf{x}_t\end{aligned}$$

$$\sigma_t^2 = \frac{(1-\bar{\alpha}_{t-1})(1-\alpha_t)}{1-\bar{\alpha}_t}$$

Property #3: Combining the two previous properties, we can obtain a different parameterization of $\tilde{\mu}_q$ which has been shown empirically to help in learning p_θ.

Rearranging $\mathbf{x}_t = \sqrt{\bar{\alpha}_t}\mathbf{x}_0 + (1-\bar{\alpha}_t)\boldsymbol{\epsilon}$ we have that:

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

Substituting this definition of $\mathbf{x}_0$ into property #2's definition of $\tilde{\mu}_q$ gives:

$$\begin{aligned}\tilde{\mu}_q(\mathbf{x}_t, \mathbf{x}_0) &= \alpha_t^{(0)}\mathbf{x}_0 + \alpha_t^{(t)}\mathbf{x}_t \\ &= \alpha_t^{(0)}\left((\mathbf{x}_0 + (1-\bar{\alpha}_t)\boldsymbol{\epsilon}) / \sqrt{\bar{\alpha}_t}\right) + \alpha_t^{(t)}\mathbf{x}_t \\ &= \frac{1}{\sqrt{\alpha_t}}\left(\mathbf{x}_t - \frac{(1-\alpha_t)}{\sqrt{1-\bar{\alpha}_t}}\boldsymbol{\epsilon}\right)\end{aligned}$$

Parameterizing the *learned* reverse process

Recall...

Recall: $p_\theta(\mathbf{x}_{t-1} \mid \mathbf{x}_t) \sim \mathcal{N}(\mu_\theta(\mathbf{x}_t, t), \mathbf{\Sigma}_\theta(\mathbf{x}_t, t))$

Later we will show that given a training sample $\mathbf{x}_0$, we want

$$p_\theta(\mathbf{x}_{t-1} \mid \mathbf{x}_t)$$

to be as close as possible to

$$q(\mathbf{x}_{t-1} \mid \mathbf{x}_t, \mathbf{x}_0)$$

Intuitively, this makes sense: if the *learned* reverse process is supposed to subtract away the noise, then whenever we're working with a specific $\mathbf{x}_0$ it should subtract it away exactly as *exact* reverse process would have.

Idea #1: Rather than learn $\Sigma_\theta(\mathbf{x}_t, t)$ just use what we know about $q(\mathbf{x}_{t-1} \mid \mathbf{x}_t, \mathbf{x}_0) \sim \mathcal{N}(\tilde{\mu}_q(\mathbf{x}_t, \mathbf{x}_0), \sigma_t^2 \mathbf{I})$:

$$\Sigma_\theta(\mathbf{x}_t, t) = \sigma_t^2 \mathbf{I}$$

Idea #2: Choose μ_θ based on $q(\mathbf{x}_{t-1} \mid \mathbf{x}_t, \mathbf{x}_0)$, i.e. we want $\mu_\theta(\mathbf{x}_t, t)$ to be close to $\tilde{\mu}_q(\mathbf{x}_t, \mathbf{x}_0)$. Here are three ways we could parameterize this:

Option C: Learn a network that approximates the $\boldsymbol{\epsilon}$ that gave rise to $\mathbf{x}_t$ from $\mathbf{x}_0$ in the forward process from $\mathbf{x}_t$ and t:

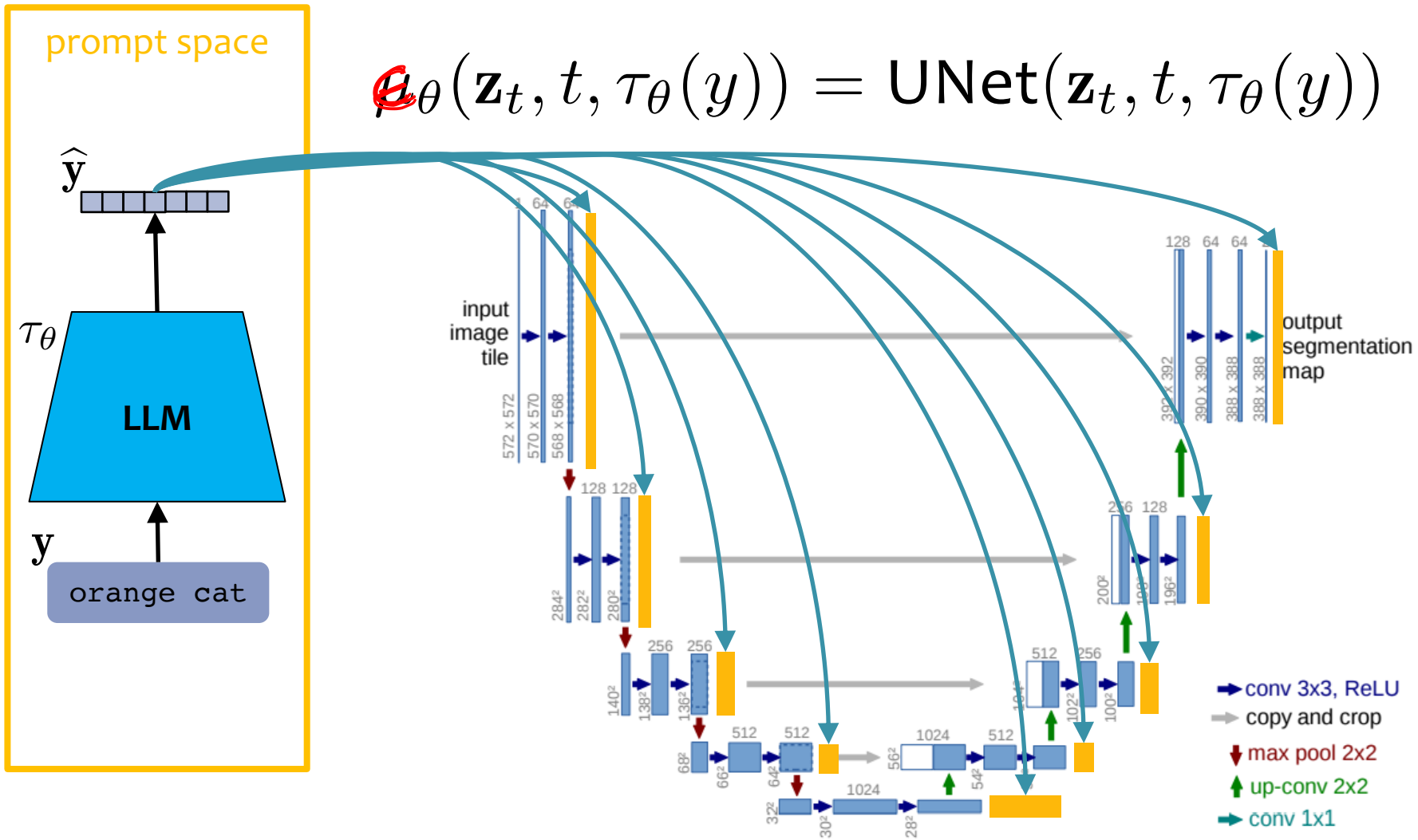
$$\mu_\theta(\mathbf{x}_t, t) = \alpha_t^{(0)} \mathbf{x}_\theta^{(0)}(\mathbf{x}_t, t) + \alpha_t^{(t)} \mathbf{x}_t$$

where $\mathbf{x}_\theta^{(0)}(\mathbf{x}_t, t) = (\mathbf{x}_t + (1 - \bar{\alpha}_t)\boldsymbol{\epsilon}_\theta(\mathbf{x}_t, t)) / \sqrt{\bar{\alpha}_t}$

where $\boldsymbol{\epsilon}_\theta(\mathbf{x}_t, t) = \text{UNet}_\theta(\mathbf{x}_t, t)$

LDM: Noise Model

$$\epsilon_\theta(\mathbf{z}_t, t, \tau_\theta(y)) = \text{UNet}(\mathbf{z}_t, t, \tau_\theta(y))$$



- The noise model includes **cross attention** (yellow boxes) to the representation of the prompt text
- During training we optimize both the parameters of the UNet noise model and the parameters of the LLM simultaneously

LDM: Cross-Attention in Noise Model

- The cross-attention is placed within a larger Transformer layer

Transformer Layer inside UNet

input	$\mathbb{R}^{h \times w \times c}$
LayerNorm	$\mathbb{R}^{h \times w \times c}$
Conv1x1	$\mathbb{R}^{h \times w \times d \cdot n_h}$
Reshape	$\mathbb{R}^{h \cdot w \times d \cdot n_h}$
$\times T$ { SelfAttention	$\mathbb{R}^{h \cdot w \times d \cdot n_h}$
$\times T$ { MLP	$\mathbb{R}^{h \cdot w \times d \cdot n_h}$
$\times T$ { CrossAttention	$\mathbb{R}^{h \cdot w \times d \cdot n_h}$
Reshape	$\mathbb{R}^{h \times w \times d \cdot n_h}$
Conv1x1	$\mathbb{R}^{h \times w \times c}$

- The cross-attention modifies the keys and values to be the prompt representation
- The queries are the current layer of UNet

$$\text{Attention}(Q, K, V) = \text{softmax}\left(\frac{QK^T}{\sqrt{d}}\right) \cdot V, \text{ with}$$

$$Q = W_Q^{(i)} \cdot \varphi_i(z_t),\ K = W_K^{(i)} \cdot \tau_\theta(y),\ V = W_V^{(i)} \cdot \tau_\theta(y).$$

Here, $\varphi_i(z_t) \in \mathbb{R}^{N \times d_\epsilon^i}$ denotes a (flattened) intermediate representation of the UNet implementing ϵ_θ and $W_V^{(i)} \in \mathbb{R}^{d \times d_\epsilon^i}$, $W_Q^{(i)} \in \mathbb{R}^{d \times d_\tau}$ & $W_K^{(i)} \in \mathbb{R}^{d \times d_\tau}$ are learnable projection matrices [36, 97]. See Fig. 3 for a visual depiction.

Figure from http://arxiv.org/abs/2112.10752

LDM: Learning the Diffusion Model + LLM

Given a training sample $\mathbf{z}_0$, we want

$$p_\theta(\mathbf{z}_{t-1} \mid \mathbf{z}_t, \tau_\theta(y))$$

to be as close as possible to

$$q(\mathbf{z}_{t-1} \mid \mathbf{z}_t, \mathbf{z}_0)$$

Intuitively, this makes sense: if the *learned* reverse process is supposed to subtract away the noise, then whenever we're working with a specific $\mathbf{z}_0$ it should subtract it away exactly as *exact* reverse process would have.

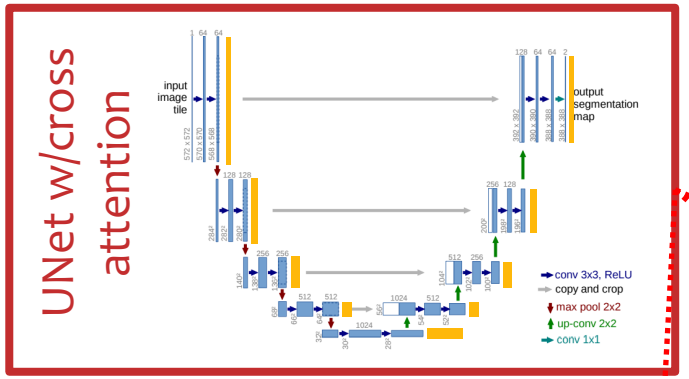
Objective Function:

$$L_{LDM} := \mathbb{E}_{\mathcal{E}(x),y,\epsilon\sim\mathcal{N}(0,1),t}\left[\|\epsilon - \epsilon_\theta(z_t, t, \tau_\theta(y))\|_2^2\right]$$

Algorithm 1 Training

1: initialize θ
2: **for** $e \in \{1, \ldots, E\}$ **do**
3: **for** $x_0, y \in \mathcal{D}$ **do**
4: $t \sim \text{Uniform}(1, \ldots, T)$
5: $\boldsymbol{\epsilon} \sim \mathcal{N}(\mathbf{0}, \mathbf{I})$
6: $\mathbf{x}_t \leftarrow \sqrt{\bar{\alpha}_t}\mathbf{x}_0 + \sqrt{1-\bar{\alpha}_t}\boldsymbol{\epsilon}$
7: $\ell_t(\theta) \leftarrow \|\boldsymbol{\epsilon} - \boldsymbol{\epsilon}_\theta(\mathbf{x}_t, t, \tau_\theta(\mathbf{y}))\|^2$
8: $\theta \leftarrow \theta - \nabla_\theta \ell_t(\theta)$

Latent Diffusion Model (LDM)



prompt space

latent space

pixel space

$\widehat{\mathbf{y}}$

τ_θ

LLM

$\mathbf{y}$

orange cat

$p_\theta(\mathbf{z}_T)$

$p_\theta(\mathbf{z}_{T-1} \mid \mathbf{z}_T, \widehat{\mathbf{y}})$

$p_\theta(\mathbf{z}_t \mid \mathbf{z}_{t+1}, \widehat{\mathbf{y}})$

$p_\theta(\mathbf{z}_0 \mid \mathbf{z}_1, \widehat{\mathbf{y}})$

z_T z_{T-1} ... z_{t+1} z_t ... z_1 z

$\mathcal{D}$

Decoder

$\widetilde{\mathbf{x}}$

$\mathcal{E}$

Encoder

$\mathbf{x}$

$q_\phi(\mathbf{z}_T \mid \mathbf{z}_{T-1})$

$q_\phi(\mathbf{z}_{t+1} \mid \mathbf{z}_t)$

$q_\phi(\mathbf{z}_1 \mid \mathbf{z}_0)$

LDM Results

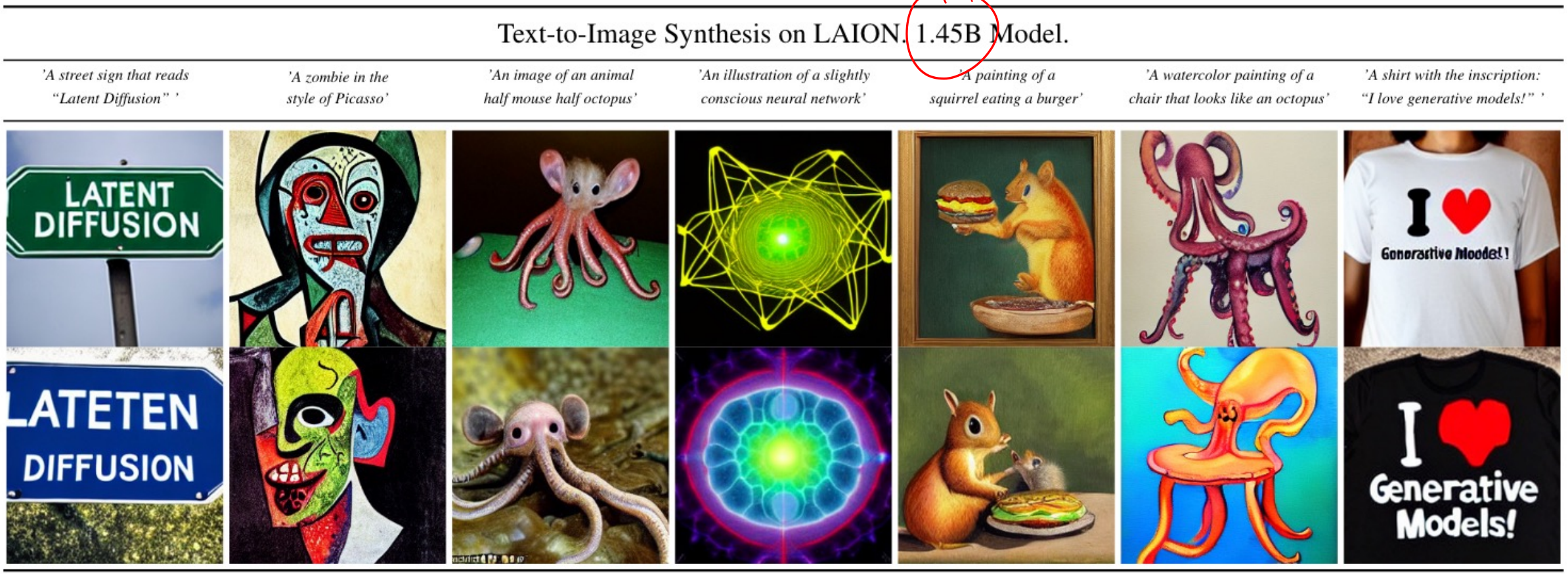


Figure 5. Samples for user-defined text prompts from our model for text-to-image synthesis, *LDM-8 (KL)*, which was trained on the LAION [78] database. Samples generated with 200 DDIM steps and $\eta = 1.0$. We use unconditional guidance [32] with $s = 10.0$.

Figure from http://arxiv.org/abs/2112.10752

LDM Results

- The result models obtain very high quality FID / IS scores with many fewer parameters than competing models
- The models are much more efficient than vanilla diffusion models because the most computationally intensive step happens in low dimensional latent space, instead of high dimensional pixel space

Text-Conditional Image Synthesis				
Method	FID ↓	IS↑	N_{params}	
CogView† [17]	27.10	18.20	4B	self-ranking, rejection rate 0.017
LAFITE† [109]	26.94	<u>26.02</u>	75M	
GLIDE* [59]	<u>12.24</u>	-	6B	277 DDIM steps, c.f.g. [32] $s = 3$
Make-A-Scene* [26]	**11.84**	-	4B	c.f.g for AR models [98] $s = 5$
LDM-KL-8	23.31	$20.03_{\pm 0.33}$	1.45B	250 DDIM steps
*LDM-KL-8-G**	12.63	$\mathbf{30.29}_{\pm 0.42}$	1.45B	250 DDIM steps, c.f.g. [32] $s = 1.5$

Table 2. Evaluation of text-conditional image synthesis on the 256×256-sized MS-COCO [51] dataset: with 250 DDIM [84] steps our model is on par with the most recent diffusion [59] and autoregressive [26] methods despite using significantly less parameters. †/*:Numbers from [109]/ [26]

Figure from http://arxiv.org/abs/2112.10752