# Scalable Diffusion Models with Transformers

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#### Intro & Motivation

- 1. Existing diffusion models adopt convolutional U-Net as backbone
- 2. Explore alternative architecture choices for generative modeling research
- 3. Introduce diffusion transformers (DiTs) and study the scaling behavior of transformers with respect to network complexity vs. sample quality



#### **Diffusion Transformers - Preliminaries**

Forward Noising Process:

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

**Reverse Denoising Process:** 

$$p_{ heta}(\mathbf{x}_{t-1}|\mathbf{x}_{t}) = \mathcal{N}(\mathbf{x}_{t-1}; oldsymbol{\mu}_{ heta}(\mathbf{x}_{t}, t), oldsymbol{\Sigma}_{ heta}(\mathbf{x}_{t}, t))$$



Noise

Reverse denoising process (generative)

Data

#### **Diffusion Transformers - Preliminaries**

Learning via variational lower bound:

- ightarrow Reparameterize  $\mu_{ heta}$  as  $\epsilon_{ heta}$  and train with L\_simple
- ightarrow Learn reverse process covariance  $\Sigma_{ heta}$  with L

$$\mathcal{L}(\theta) = -p(x_0|x_1) + \sum_t \mathcal{D}_{KL}(q^*(x_{t-1}|x_t, x_0)||p_{\theta}(x_{t-1}|x_t))$$
  
$$\mathcal{L}_{simple}(\theta) = ||\epsilon_{\theta}(x_t) - \epsilon_t||_2^2$$

Algorithm 1 Training	Algorithm 2 Sampling
1: repeat 2: $\mathbf{x}_0 \sim q(\mathbf{x}_0)$ 3: $t \sim \text{Uniform}(\{1, \dots, T\})$ 4: $\boldsymbol{\epsilon} \sim \mathcal{N}(0, \mathbf{I})$ 5: Take gradient descent step on $\nabla_{\boldsymbol{\theta}} \  \boldsymbol{\epsilon} - \boldsymbol{\epsilon}_{\boldsymbol{\theta}}(\sqrt{\bar{\alpha}_t}\mathbf{x}_0 + \sqrt{1 - \bar{\alpha}_t}\boldsymbol{\epsilon}, t) \ ^2$ 6: until converged	1: $\mathbf{x}_T \sim \mathcal{N}(0, \mathbf{I})$ 2: for $t = T,, 1$ do 3: $\mathbf{z} \sim \mathcal{N}(0, \mathbf{I})$ if $t > 1$ , else $\mathbf{z} = 0$ 4: $\mathbf{x}_{t-1} = \frac{1}{\sqrt{\alpha_t}} \left( \mathbf{x}_t - \frac{1-\alpha_t}{\sqrt{1-\overline{\alpha}_t}} \boldsymbol{\epsilon}_{\theta}(\mathbf{x}_t, t) \right) + \sigma_t \mathbf{z}$ 5: end for 6: return $\mathbf{x}_0$

#### **Diffusion Transformers - Preliminaries**

**Classifier-Free Guidance** 

 $\rightarrow$  Reverse process conditioned on class c:  $p_{\theta}(x_{t-1}|x_t,c)$ 

 $\rightarrow \hat{\epsilon}_{\theta}(x_t,c) = \epsilon_{\theta}(x_t,\emptyset) + s \cdot \nabla_x \log p(x|c) \propto \epsilon_{\theta}(x_t,\emptyset) + s \cdot (\epsilon_{\theta}(x_t,c) - \epsilon_{\theta}(x_t,\emptyset))$ 

Latent Diffusion Models

 $\rightarrow$  Learn an autoencoder with a learned encoder E

- $\rightarrow$  Train a diffusion model of representations z = E(x) instead of x (E is frozen)
- → Off-the-shelf convolutional VAEs & transformer-based DDPMs

# **Diffusion Transformer Design Space**

DiT is largely based on Vision Transformers (ViTs)

For Patching:

- $\rightarrow$  Input is the encoded image z
- $\rightarrow$  For 256x256x3 image, input z has shape 32x32x4
- $\rightarrow$  Patch size p = 2, 4, 8, Token length T = (I / p) \*\* 2
- $\rightarrow$  Additionally apply sine-cosine positional embeddings





# Diffusion Transformer Design Space

Incorporating Timestamp & Class in DiT Blocks:

- $\rightarrow$  Timestampe t and class c are appended as special tokens
- $\rightarrow$  Additional cross attention layer for (t, c) above self-attention
- $\rightarrow$  Adaptive layer norm (adaLN) by regressing gamma / beta upon (t, c)
- $\rightarrow$  Zero-initializing the final layer in adaLN prior to residual connections Decoder Blocks:
  - $\rightarrow$  Linear decoder to predict noise and covariance (p x p x 2C shape)

#### **Diffusion Transformer Design Space**



Latent Diffusion Transformer

DiT Block with adaLN-Zero

DiT Block with Cross-Attention

DiT Block with In-Context Conditioning

# **Training Diffusion Transformers**

- 1. Leverage pretrained VAE from stable diffusion
- 2. After sampling, decode pixels using VAE decoder
- 3. Evaluate with FID scores and 250 DDPM sampling steps



### Model and Patch Size

Increasing transformer size



#### Model and Patch Size / DiT Gflops



# Model and Patch Size / DiT Gflops

Model / Patch Size:

- 1. Improvements in FID are obtained by making transformer models larger
- 2. Improvements in FID are obtained by reducing patch sizes (scaling tokens)

Gflops:

- 1. Scaling model Gflops is the key to improved FID performance
- 2. Given constant Gflops, different DiT configs obtain similar FID values



# DiT Model Size / Efficiency

Training compute:

Model Gflops \* batch \* training steps \* 3

- 1. Given constant training Gflops, larger DiT models are more efficient
- 2. Models that are identical except for patch size have different performance profiles even when controlling training Gflops
- 3. Similar observation on qualitative examples



### **Further Qualitative Examples**

Increasing transformer size



Decreasing patch size

# Scaling & Sampling

Scaling Model vs. Sampling Compute

- Consider constant sampling compute (DiT-XL/2 w/ 128 steps OR DiT-L/2 w/ 1000 steps)
- 2. In most cases, scaling-up sampling compute (steps) cannot compensate for the lack of model compute



#### **Comparison to SOTA Models**

For both 256 x 256 and 512 x 512 image resolution, the proposed DiT can

outperform existing methods and

remains compute efficient.

Class-Conditional ImageNet 512×512					
Model	FID↓	sFID↓	IS↑	Precision↑	Recall↑
BigGAN-deep [2]	8.43	8.13	177.90	0.88	0.29
StyleGAN-XL [53]	2.41	4.06	267.75	0.77	0.52
ADM [9]	23.24	10.19	58.06	0.73	0.60
ADM-U	9.96	5.62	121.78	0.75	0.64
ADM-G	7.72	6.57	172.71	0.87	0.42
ADM-G, ADM-U	3.85	5.86	221.72	0.84	0.53
DiT-XL/2	12.03	7.12	105.25	0.75	0.64
DiT-XL/2-G (cfg=1.25)	4.64	5.77	174.77	0.81	0.57
DiT-XL/2-G (cfg=1.50)	3.04	5.02	240.82	0.84	0.54

Class-Conditional ImageNet 256×256					
Model	FID↓	sFID↓	IS↑	Precision↑	Recall↑
BigGAN-deep [2]	6.95	7.36	171.4	0.87	0.28
StyleGAN-XL [53]	2.30	4.02	265.12	0.78	0.53
ADM [9]	10.94	6.02	100.98	0.69	0.63
ADM-U	7.49	5.13	127.49	0.72	0.63
ADM-G	4.59	5.25	186.70	0.82	0.52
ADM-G, ADM-U	3.94	6.14	215.84	0.83	0.53
CDM [20]	4.88	-	158.71	-	-
LDM-8 [48]	15.51	-	79.03	0.65	0.63
LDM-8-G	7.76	-	209.52	0.84	0.35
LDM-4	10.56	-	103.49	0.71	0.62
LDM-4-G (cfg=1.25)	3.95	-	178.22	0.81	0.55
LDM-4-G (cfg=1.50)	3.60	-	247.67	0.87	0.48
DiT-XL/2	9.62	6.85	121.50	0.67	0.67
DiT-XL/2-G (cfg=1.25)	3.22	5.28	201.77	0.76	0.62
DiT-XL/2-G (cfg=1.50)	2.27	4.60	278.24	0.83	0.57

# **Additional Results**

#### Metrics

- 1. Additional metrics: sFID, Inception Score, Precision, Recall
- 2. FID-driven analysis in the paper generalizes to the other metrics across every metric, scaled-up DiT models are more compute-efficient and model Gflops are highly-correlated with performance. In particular, Inception Score and Precision benefit heavily from increased model scale.



# **Additional Results**

#### **Training Loss**

- 1. Training curves for different DiT model sizes
- 2. Increasing DiT model Gflops (via transformer size or number of input tokens) causes the training loss to decrease more rapidly and saturate at a lower value. This phenomenon is consistent with trends observed with language models, where scaled up transformers demonstrate both improved loss curves and downstream performance.



# **Additional Results**

VAE Decoder Ablations

1. XL/2 continues to outperform all prior models when using the LDM decoder.

Class-Conditional ImageNet 256×256, DiT-XL/2-G (cfg=1.5)					
Decoder	FID↓	sFID↓	IS↑	Precision <sup>↑</sup>	<b>Recall</b> ↑
original	2.46	5.18	271.56	0.82	0.57
ft-MSE	2.30	4.73	276.09	0.83	0.57
ft-EMA	2.27	4.60	278.24	0.83	0.57



# Conclusion

- 1. Introduced Diffusion Transformers (DiTs), a simple transformer-based backbone for diffusion models that outperforms prior U-Net models.
- 2. Incorporated a series of improvements upon ViT such as cross attention, adaptive layer norm etc., leading to improved generation performance
- 3. Given the promising scaling results in this paper, future work should continue to scale DiTs to larger models and token counts. Alternatively, DiT could also be explored as a drop-in backbone for text-to-image models like DALLE.

# **Thoughts & Discussion**

- 1. Conditioned on class
  - $\rightarrow$  Text prior is not considered
- 2. Only experimented on ImageNet
  - $\rightarrow$  Further experiments and evaluation may be beneficial
- 3. Focuses on the scaling of DiT rather than efficiency
  - $\rightarrow$  Sampling is still somewhat expensive
  - $\rightarrow$  Maybe MoE or fewer steps like InstaFlow
  - $\rightarrow$  Model size are much smaller than LLMs (33M to 675M)